Foreword

Paradise Valley Community College is proud to present this 2-volume set of the *22nd Annual Mancini Science Symposium*. This symposium was held on May 12, 2016 in the Center for Performing Arts (CPA).

Dr. Hank Mancini started the *Annual Science Symposium* in 1995. The first publication contained papers from his Organic Chemistry class. This annual symposium has since grown to include papers from Physics, Engineering, Astronomy, Chemistry, Biology, and Math. What began with 9 research papers in the first year has evolved into 84 papers this year. Dr. Mancini retired in 2012. After his retirement, the symposium was renamed the *Annual Mancini Science Symposium*.

Students enrolled in Astronomy, Chemistry, and Physics classes from PVCC participated in the event this year. Each contributor was responsible for selecting and researching his/her topic and preparing a paper. This 2-volume set contains all 84 papers (24 in Astronomy, 4 in Chemistry, and 56 in Physics). A few students gave oral presentations of their project to their peers. Students chose the oral presentation topics.

I would like to thank the following faculty members for participating in this event:
Scott Massey (PhD) – Chemistry,
Bill Sherry (PhD) – Astronomy,
Mike Swingler (MS) – Physics, and
Sydney Wilson (PhD) – Physics.

As instructors and faculty advisors for this symposium, we want to thank and congratulate each participant for his/her effort, courage, and dedication. By participating, these individuals perpetuate this event annually. We are proud and honored to present the work of these individuals.

Casey Durandet, PhD
*Symposium Coordinator*
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Maintenance of Arrhythmia Through the Use of a Pacemaker

Sean Koester

PHY 112
Dr. Casey Durandet
April 21, 2016
ABSTRACT

The following discusses how pacemakers are capable of managing the life threatening heart conditions of arrhythmias. When the heart is unable to contract in a coordinated and consistent manner, its ability to supply the body with blood becomes compromised. Without a sufficient supply of blood the deprived tissue will begin to die. As a result maintaining a consistent and effective heart rhythm is imperative to the patients survival. A pacemaker monitors the heart rhythm by measuring the voltage increases across and the frequency in which these increases happen. When the pacemaker detects an irregular or ineffective heart rhythm it will generate a calculated electrical impulse that will help to jump start the heart's regular rhythm. Advancements are constantly being made in the field of pacemaker research which are leading to increasingly more effective devices.

DISCUSSION

Cardiovascular System

The natural function of the human body relies on the proper operation of a wide range of interwoven organ systems throughout the body. One of the most important of these systems is the cardiovascular system. The cardiovascular system consists of the heart and an extensive network of arteries and veins that run throughout the body. The goal of this system is to transport oxygen saturated blood to all parts of the body. The key component of the cardiovascular system is the heart which serves as a pump in the transportation of blood throughout the body. The human heart consists of four chambers, the upper left and right atrium and the left and right ventricle below. The atria serve as collecting chambers for the blood returning to the heart. In contrast the ventricles serve as collecting chambers for the blood that is about to be pumped to the rest of the body.

The circulation of blood begins when oxygen deprived blood is pumped into the right atrium. After the right atrium fills, the blood travels into the right ventricle. From this point the blood is pumped into the lungs where it can acquire a fresh supply of oxygen. The freshly oxygenated blood returns to the heart and enters the left atrium. After filling the left atrium, the blood is pumped into the left ventricle. From here the blood is pumped out through the aorta and travels to the rest of the body.

Understanding the electrical conduction system of the heart requires knowledge of the interconnection between voltage, current, and impedance. Voltage is the “force that causes electrons to move through a circuit.” Voltage is measured in the unit of volts. The passage of these electrons through a circuit is known as the current. The flow of the energy caused by these electrons is measured in amperes. The impedance of a circuit is “the sum of all resistance to the flow of current.” These three elements are interdependent, a change in the value of one will have an effect on the other two. The relationship between these elements is described in Ohm's law, $V=IR$, where voltage is equal to the current times the impedance. In the same fashion voltage is directly proportional to both current and impedance.

Biological elements such as tissue are capable of conducting electricity in much the same way as a circuit. The conductivity of the myocardial tissue of the heart plays an incredibly important role in the function of the cardiovascular system. In much the same way as a traditional electrical circuit, the flow of current through tissue is a function of the applied voltage and the effective impedance of the tissue. In a biological circuit there is only a small variance in the voltage flowing through the tissue. The
changes happen so slowly that it is viewed to be quasistatic. This means that the changes happen so slowly that “capacitive and inductive effects and the finite speed of electromagnetic radiation” can be ignored\(^6\). A major factor to consider when evaluating the conductivity of tissue is that the cell tissue and the interstitial space are made of different materials. As a result, the effective impedance of the two materials differs\(^6\). This can lead to variances in the current flow as it passes through each material.

However, the heart has a structure to help bypass this potential conductivity problem. The heart cells are connected by a piece of tissue known as a syncytium, in which the “intracellular spaces of adjacent cells are coupled through inter-cellular channels, so that current can pass between any two intracellular points without crossing the cell membrane\(^6\)”. This biological feature helps to ensure an even flow of current through the myocardial tissue of the heart. This even flow of current is vital to promoting a coordinated contraction for the heart.

The pumping of the heart functions through the process of a coordinated contraction. The contraction is initiated by the sinoatrial node, which generates an electrical impulse. The impulse travels through the atria and then the ventricles. There is a brief pause when the impulse reaches the atrioventricular node which lies between the atria and ventricles. This pause is necessary because it allows for the ventricles to fill before contracting\(^7\). After the impulse travels past the atrioventricular node it travels across the bundle of His and Purkinje fibers which serve as a conduction system for travel through the ventricles. As the electrical impulse travels across the heart, it increases the voltage at each individual cell it passes. The impulse simulates a wavelike rising and falling of voltage through the heart as it travels. The voltages of each cell increase as the impulse comes in contact with them and falls back to normal levels shortly thereafter. This rising and falling pattern creates what is known as an action potential. The spreading of this action potential causes the cells within the heart muscle to release calcium. The influx of calcium causes the heart muscle to contract\(^5,9\).

### Disorders That Interrupt Normal Cardiac Flow

Conditions that can interrupt or hamper the cardiovascular system typically come in the form of irregular heart rhythms. The heart typically contracts in a uniform and coordinated manner. Any deviations from this normal rhythm, known as sinus rhythm, can have severe consequences\(^11\). Being able to detect these rhythm irregularities is of the utmost importance. An important tool in the analysis of the heart rhythm is the electrocardiogram. The electrocardiogram produces a wave function that represents the different phases of heart depolarization and contraction\(^11\). The vertical axis of the electrocardiogram graph is the voltage measured in the heart and it is measured in units of milliamps. The horizontal axis of the graph measures the time in units of milliseconds. By comparing the two variables the voltage increases in the heart and their associated time intervals can be evaluated. The different sections of the typical heart rhythm on an electrocardiogram can be seen in figure 1. A sample electrocardiogram can be seen in figure 2. The wave function created can be divided into several main sections, the P wave, the QRS complex and the ST segment and T wave. The P wave is the initial portion of the wave and represents the depolarization and contraction of the atria. The QRS complex follows the P wave and is associated with the depolarization and contraction of the ventricles.

Anytime the heart's rhythm varies from this coordinated and predictable pattern it is called an arrhythmia. These irregular heart rhythms come in a number of forms such as ventricular tachycardia, ventricular fibrillation, atrial fibrillation, atrial flutter, and bradycardia.

Ventricular tachycardia is a condition of a fast and irregular heartbeat. In this condition the heart beats at a rate of over 100 beats per minute and often at an irregular rate\(^12\). Common symptoms associated with ventricular tachycardia are angina, syncope, light-headedness, palpitations, and shortness of
breath\textsuperscript{12}. Angina is chest pain that is caused by insufficient blood flow through the coronary arteries and myocardium\textsuperscript{12}. Syncope is a sudden loss of consciousness that is often associated with the brain receiving insufficient oxygen. This deprivation of oxygen is due to the heart being unable to pump sufficient blood throughout the body\textsuperscript{15}. The brain relies on the heart providing it with this oxygen rich blood. If it is unable to receive the required amount of oxygen, the brain's function will be severely impaired. The symptom of light-headedness is often attributed to the same causes as the more sudden syncope. Heart palpitations are a condition when a person feels like their heart is pounding. The sensation can often be felt in the chest, throat, or neck\textsuperscript{14}. Shortness of breath is also caused by the heart's inability to sufficiently pump blood to the body. If the body doesn't receive sufficient levels of oxygenated blood its function will be impaired. Ventricular tachycardia typically shows up on an electrocardiogram as an extended and uneven QRS complex\textsuperscript{9}. Premature ventricular contractions can also be witnessed on an electrocardiogram\textsuperscript{9}. Ventricle tachycardia can be seen in the electrocardiogram shown in figure 3.

Ventricular fibrillation is a severe condition that results from a significant impairment of heart function. It is an arrhythmia that results from the impairment of the heart's electrical activity and the resulting contractions become ineffective\textsuperscript{9}. This condition condition causes blood pressure to plummet and will result in death of the patient, due to insufficient cardiac output\textsuperscript{9}. Ventricle tachycardia can lead to ventricular fibrillation if not addressed. One potential cause for this condition is the repeated activation of electrical activity which lead to spiral waves of electrical active passing through the ventricles. These repeated stimuli prevent the heart from contracting in an organized manner which severely inhibits its ability to pump blood\textsuperscript{9}. Ventricular fibrillation can be identified on an electrocardiogram by a repeated wavelike baseline. Also there is no clearly identifiable QRS complex. An example of ventricular fibrillation can be seen in figure 4.

Atrial fibrillation is a condition in which the heart beats rapidly and irregularly \textsuperscript{16}. In this condition the atria contract rapidly and out of sync of the ventricles. The ventricles are not able to fill up before they contract. As a result the heart is no longer able to efficiently pump blood throughout the body. Atrial fibrillation can be seen displayed on an electrocardiogram in figure 5.

Atrial flutter is a condition under which the atria undergo a repeated loop of depolarization. This arrhythmia is similar to atrial fibrillation, but the increased rate of contractions happen at a regular pace\textsuperscript{15}. A key identifier for this condition on an electrocardiogram is a sawtooth pattern. An example of this condition displayed on an electrocardiogram can be seen in figure 6.

In contrast to these other arrhythmias, bradycardia is an abnormally slow heart rhythm. Unlike the other arrhythmias, bradycardia is not always a life threatening condition. It is commonly seen in athletes\textsuperscript{17}. However, there are some cases where it is due to a harmful condition. In these cases the condition is likely due to problems with the electrical conduction system within the heart\textsuperscript{17}.

**Pacemakers**

The management of these arrhythmias is vital to a patient's survival. A key tool that has gone a significant way to helping manage these life threatening conditions is the artificial pacemaker. The pacemaker is a device that is planted underneath the skin. It is the job of this device to correct these abnormal heart rhythms. The pacemaker performs this task by sending electrical impulses to the heart in an attempt to re-establish regular heart rhythm\textsuperscript{7}. The pacemaker detects whether a shock is necessary to be delivered by measuring the voltage produced by the heart during contraction. The generated voltage is typically very small and is measured in millivolts.
The pacemaker uses an intricate detection mechanism to accurately evaluate the heart's rhythm. In order for the pacemaker to effectively detect the heart's rhythm, specific voltage thresholds are set for each specific portion of the heart contraction. The pacemaker also measures the time interval that these voltages are present. By comparing these two values the pacemaker can make accurate interpretation of the hearts contraction rhythm. The use of these two variables helps the pacemaker to avoid misinterpreting similar voltages that it may detect, but may be present for differing time periods. An example of this is that the P wave has a voltage range that falls within the same range as the T wave, but happens over a considerably briefer time period. These factors also help the pacemaker to disregard the electrical signals detected from other sources such as the skeletal muscle myopotentials of the nearby pectoral muscles. An example of the different electrical signal ranges that can be detected by the pacemaker can be seen in figure 7.

The anatomy of a pacemaker consists of a battery that serves as the power source, insulated wires that connect the power source to the myocardial tissue of the heart, and a computer chip that interprets and evaluates the sensory information received on the heart's rhythm. The typical pacemaker uses a lithium battery as its power source. The lithium battery is an ideal choice because it provides a long lasting and efficient battery. This characteristic is important because the replacement of the battery requires a surgical procedure. There are two distinct categories of pacemakers, single chamber and dual chamber pacemakers. The pacemaker is capable of measuring the electrical impulses generated in either the atria, ventricle, or both. The particular part of the heart that the pacemaker monitors determines which section of the heart it is capable of pacing. The single chamber pacemaker is ideal for heart conditions that are isolated in either the atria or the ventricle. On the other hand, the dual chamber pacemaker is designed for a heart that requires both pacing of atria and ventricles. The type of pacemaker will determine where the leads are attached to the myocardial tissue. The design of the pacemaker is attuned to the condition the heart is suffering. In the case of single chamber pacemakers, a single lead is connected from the battery into the myocardial tissue of either the atria or ventricle. A dual chamber pacemaker is ideal in a situation where both the atria and the ventricle contraction rhythms are compromised. This pacemaker employs two separate leads, one lead is connected to both the atria and ventricle. This pacemaker is capable of monitoring and ensuring that both the atria and ventricles contract in a coordinated manner. There is also a distinction between types of pacemakers depending on whether the pacemaker inhibits or stimulates a response. If an inhibitory pacemaker senses a natural heart rhythm it will inhibit its own stimulatory response until it senses an abnormal heart rhythm. In contrast, the stimulatory pacemaker waits until it senses an abnormal heart rhythm and it will initiate its electrical stimulus to attempt to restore normal heart rhythm.

The pacemaker's ability to stimulate the heart relies on the concept of the stimulus threshold. The stimulus threshold is the “minimum electrical stimulus needed to consistently capture the heart outside of the refractory period”. The refractory period for the heart is the duration under which the heart is unable to receive another stimulus. The purpose of the refractory period is to prevent the heart from being over stimulated. When the pacemaker is able to provide sufficient stimulus to initiate a contraction of the heart it is execute what is known as a myocardial capture. The pacemaker's ability to trigger myocardial capture are determined by two main factors, the strength of the impulse and the duration of the current flow. The impulse strength must be high enough to cause depolarization of the heart to occur, but not too strong to damage the tissue. The duration of the current flow must be long enough for the depolarization to spread throughout the heart. These two characteristics are interconnected. For example, the stronger the impulse given by the pacemaker the shorter the required duration of the current flow. The relationship between these two characteristics can be seen in figure 8. Impedance also plays an important role in the effectiveness of the pacemaker. If the impedance is too
high, it will require a significantly higher voltage to be applied to generate the necessary current.

The pacemaker's sensory mechanism plays a vital role in maintaining a regular heart rhythm. The patient's voltage levels during heart contraction are analyzed and a baseline is set for each individual section. The differing waves of the heart contraction have unique voltage and duration parameters. When the pacemaker senses a voltage increase beyond the baseline established for that specific wave it will identify it as that particular wave. The pacemaker also takes into account the duration of the current flow through the heart in an attempt to help filter out similar voltage readings.

When a pacemaker senses an irregular heart rhythm, it must calculate the amount of voltage to apply to the heart. The energy applied during the shock must surpass what is known as the defibrillation threshold. When the shock is applied the myocardium is captured and the heart contracts or it does not. There is no halfway point in applying an impulse to the heart, it is an "all or none phenomenon." Any given energy amount is generally not guaranteed to defibrillate the heart. Therefore, the chance of successfully applying the impulse to the heart is a factor of probability. Figure 9 illustrates the relationship between increasing energy values and the chance of successfully capturing the heart during the delivery of an impulse.

Another factor that has an effect on the success of the impulse re-establishing a regular heart rhythm is the type of shock applied. There are two main categories of shock waveforms. The first type is a monophasic shock in which the shock flows in a single direction after being applied. The other type of shock is the biphasic shock. The energy in a biphasic shock "reverses direction during the discharge." The biphasic shock has several advantages over its monophasic counterpart. The use of a biphasic shock delivery lowers the defibrillation thresholds, lowers the chance of short-term heart tissue injury, and has a faster recovery time to normal heart rhythm.

When a pacemaker is able to function under proper conditions it is capable of sufficiently ensuring a regular heartbeat in the patient. However, there are a number of possible conditions that can hinder the pacemaker's function. One possible complication for the pacemaker is damage to the leads. This damage can come in the form of insulation breaks or wire fractures. When the insulation on the wiring breaks, it exposes the wire to body fluids. Bodily fluids have a lower resistance than the wiring which will cause impedance values to fall. While having a low impedance value is important, but the exposure of the wire has a significantly negative effect on the pacemaker. The breaking of the insulation will cause the current to drain into the body, which will deplete the battery. In the case of wire fractures, there will be the opposite effect on the impedance value. The fraying of wires will cause impedance values within the wire to increase, resulting in a serious complication. If the impedance value becomes too high the current supplied by the battery will become insufficient and the pacemaker will be ineffective. Maintaining structural integrity of the pacemaker is imperative to its proper function.

A major concern of any patient who has an implanted pacemaker is the effect of electromagnetic interference. Encountering electromagnetic interference can significantly impair the function of a pacemaker. Electromagnetic interference is defined as "any signal, either biologic or non-biologic, that falls within a frequency spectrum that are being detected by the sensing circuitry of the pacemaker." Electromagnetic interference occurs when "electromagnetic waves emitted by one electronic source or device impede the normal function of another electronic device." There are three potential scenarios for the effect of electromagnetic interference. Electromagnetic interference can cause the pacemaker to deliver irregular impulses, ignore the heart's rhythm and deliver its own stimulus, or stop the delivery of the stimulus completely. A number of electronic devices can deliver
these interfering signals such as electrocautery, magnetic resonance imaging, and radiofrequencies. The electromagnetic interference can come in the form of direct contact such as the surgical cautery procedure where the patient comes into direct contact with the current. On the other hand, in the case of a strong magnetic field like the one present in magnetic resonance imaging, contact is not required to cause interference. Any electrical device that produces a signal within the 10 to 60 Hz range can interfere with the pacemaker because it will overlap with the cardiac signal range. It is imperative for any patient with a pacemaker to avoid electromagnetic interference to ensure the device's proper function. There have been a number of advancements made to limit the damage of electromagnetic interference. The changes made to mitigate this effect include “titanium shielding, signal filtering, interference rejection circuits, and programmable parameters.” The inclusion of bipolar leads has also helped to mitigate the effect of interference on the pacemaker. The bipolar leads are more effective than their unipolar counterparts because the distance between the electrode and the antenna is shorter, leaving less time for electromagnetic interference to take effect.

The field of pacemakers is constantly advancing and there is a number of experimental models undergoing research. The most potentially promising of these new fields of research are the biological pacemaker and the leadless pacemaker. Neither of these two prospective pacemaker designs are fully developed and are still undergoing experimentation. The leadless pacemaker is the most developed and has begun clinical trials in humans. In contrast to the pacemaker being implanted in a pocket in the chest close to the heart, the leadless pacemaker is implanted directly inside the heart. The pacemaker will monitor the heart's rhythm and emit small pulses of electricity whenever it determines that there is an abnormal heart rhythm. The leadless pacemaker can be seen implanted inside the heart in figure 10. Like the leadless pacemaker, the biological pacemaker is still in its early development stages. However, the biological pacemaker has not advanced far enough to begin human trials. Due to its early stages of development, there are several potential methods that are being attempted for biological pacemakers. One process for the biological pacemaker is the implanting of “pacemaker” embryonic cells into the myocardium. The use of this gene therapy causes an over expression of pacemaker channel genes. The most recent method being tested for biological pacemaker is the use of the transcription factor TBX18 for somatic reprogramming of ventricular myocytes to induce sinoatrial cells. These reprogramming cells will resemble the endogenous pacemaker cells. The newly reprogrammed cells will be able to help facilitate the normal heart contraction rhythm that the damaged cells are no longer able to maintain. While the biological pacemaker shows promise for the future, there are a number of concerns that are present. One concern is that since the implanted cells are biological in nature they are not perfectly controllable and may migrate outside of the desired area of impact. The other concern is that the use of TBX18 reprogramming potentially runs the risk of causing heart arrhythmias. However, there have been no documented cases of induced arrhythmias in animal testing of TBX18 reprogramming. Both methods show promise for the advancing field of pacemakers, but there is a considerable amount of experimentation left to perfect them.

CONCLUSION

Heart arrhythmias are a serious condition that affect a significant portion of the population. When the heart is unable to contract in a coordinated and regular manner, its ability to provide a sufficient supply of oxygenated blood to the body is comprised. The development of the pacemakers has gone a long way to mitigating the harm of these conditions. The pacemaker can constantly monitor the heart rhythm and can emit a calculated electrical impulse to stimulate the heart's contraction process.

One of the most intriguing aspects I found was how the conductivity of tissue was very similar to that of an typical electric circuit. It was interesting to see biological versions of both batteries and resistors.
within tissue. However, the calculations for the conductivity value were difficult to grasp due to them going well beyond this level of physics.

The current advancements in the field of pacemakers has been rather intriguing. I find the experimentation with both leadless and biological pacemakers to be exciting. Once these techniques are perfected there will be alternatives to the traditional pacemaker. The leadless pacemaker is appealing because it has fewer parts than the traditional pacemaker, therefore less of a chance for something to break requiring a surgical procedure to repair. Without the traditional pacemaker wires there will no longer be a concern of wire fraying or fracture causing unintended current to run into the myocardial tissue. I also find the concept of being able to emit an electrical pulse from inside the heart and directly simulate the sinoatrial node to be fascinating. Overall, I find the field of pacemaker advancement to be very promising, with many excellent improvements coming down the pipeline.
Figures

Figure 1 – Different sections of the typical ECG. Available from http://www.anaesthesiak.com/SearchRender.aspx?DocId=1304&Index=D%3a%5cdtSearch%5cUser Data%5cAUK&HitCount=11&hits=4+5+12+13+32+5c+60+64+a8+16e+177+

Figure 2 – Normal electrocardiogram. Available from http://www.ecglibrary.com/norm.php

Figure 3 - Ventricular Tachycardia. Available from http://www.anaesthesiak.co.uk/SearchRender.aspx?DocId=1311&Index=D%3a%5cdtSearch%5cUser Data%5cAUK&HitCount=18&hits=3+b+24+29+43+69+6d+d2+103+104+12d+133+134+183+1f3+1f6+20c+24e+
Figure 4 – Ventricular Fibrillation. Available from http://www.anaesthesiaweb.co.uk/SearchRender.aspx?DocId=1311&Index=D%3a%5cdtSearch%5cUser Data%5cUK&HitCount=18&hits=3+b+24+29+43+69+6d+d2+103+104+12d+133+134+183+1f3+1f6+20c+24e+

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<th>P Wave</th>
<th>PR interval (in seconds)</th>
<th>QRS (in seconds)</th>
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<td>Absent</td>
<td>N/A</td>
<td>Fibrillatory baseline</td>
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Figure 5 – Atrial Fibrillation. Available from http://www.anaesthesiaweb.co.uk/article.aspx?articleid=100686

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<th>Rhythm</th>
<th>P Wave</th>
<th>PR interval (in seconds)</th>
<th>QRS (in seconds)</th>
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<td>Irregular</td>
<td>Fibrillatory (fine to course)</td>
<td>N/A</td>
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Figure 6 – Atrial Flutter. Available from http://www.anaesthesiaku.co.uk/article.aspx?articleid=100686


Figure 10 – Leadless Pacemaker. Available from
https://www.sjm.com/leadlesspacing/intl/options/leadless-pacing
References


Parallel Universes

Neda Kooklani

04/21/2016

Physics 112

Dr. Durandet
Abstract:

This research paper is about the theory of parallel universes. This theory offers an explanation about an important problem of quantum mechanics presented by Schrödinger. Based on the dual functionality of a particle and uncertainty principle, he hypnotized that the probability of something happening or not happening at a specific time is equal. So, a cat can be dead and alive at the same time! In order to explain this, scientists came up with many different explanations. One of the most famous ones is the idea of parallel universes. In this paper, first a history of quantum mechanics is presented in order to understand how scientists came to face this question. Then the idea of parallel universes is presented and at the end, the author discusses her own ideas about this theory.

History of Quantum Mechanics:

The fundamental idea that initialized the idea of quantum mechanics was the dual functionality of light. What started this idea was the results of a very interesting experiment called double slit experiment.

Double slit experiment was originally done by an English scientist named Thomas Young. He was born in 1773 and died in 1829. The original experiment was done in 1801. This experience showed that light has a wave nature. Later on, Einstein’s theory of light as photons had conflict with this theory. Therefore, many other scientists did this experiment in many different forms because this experiment is known to provide some understandings of basic fundamental concepts of quantum mechanics.

Imagine a screen with two narrow holes in it. A machine gun sprays bullets at the screen. There is another screen right behind this one that bullets will land on, that is shown in figure number 1. The experiment is done in three different steps. In one step, only slit one is open, for instance, slit 1. Every bullet that leaves the Machin gun will land somewhere on the backstop, and probably somewhere opposite of the position of the open slit. Of course, not all of them land exactly at the same spot, but roughly around an area right behind slit 1:

\[ P_1(x) = \text{probability of a bullet landing in the range } (X, X+\Delta X) \]

And if the open slit is slit 2:

\[ P_2(x) = \text{probability of a bullet landing in the range } (X, X+\Delta X) \]

Finally, both slits are opened. The bullets will randomly choose to go through slit 1 or slit 2. The possibility of bullets landing somewhere behind the screen would be:

\[ P_{12}(x) = P_1(x) + P_2(x) \]

So far, it makes perfect sense. The same experiment done with light provides interesting results. Imagine two beams of light coming from one source. These two beams of light have exactly the same wavelength because they are coming from the same source. The slits are very small, so
only a beam of light can pass through them. They are also very close together in a way that their distance is comparable to the wavelength of light. This time, both slits are open at the same time as it shown in figure number 2. Light will not travel as a straight line. It will spread out two dimensions as shown in the picture from both slits and the result will be overlapping (interfering) of the two patterns. When waves interfere instructively they add up and when they interfere destructively they cancel each other out.

The appearance of the second screen cannot be explained with the $P_{12}(x) = P_1(x) + P_2(x)$ formula. The only way to explain this is considering light as a wave function. So, when constructive interference happens there is a bright spot; when destructive interference happens there is a dark spot, and medium bright spots are the result of half constructive half destructive waves. So, light may well be a wave function.

This dual functionality of light made scientists to think about what is the true nature of light. Many different explanations were offered but the most important one was presented by Max Plank during the study of the black body radiation problem.

Quantum mechanics probably was started with Max Plank. He was a German physicist who was born in 1858 and died in 1947. In early 1900s, he was working on one of the most important problems of that time: The black body radiation problem. Imagine a body that absorbs heat and transfer the heat energy to light energy; as the body gets hotter, according to the rules of the classical physics, the light energy increases and as a result, infinite amount of ultraviolet radiation will be produce. That means that if something has been heated long enough it should eventually give off the frequency of the ultra violet light meaning that it eventually disappears! So, why it does not make sense? Max plank came up with a theory that was the fundamental theory of the quantum mechanics. He theorized that energy is actually “quantized”; meaning that it comes in portions but millions and millions of them at one time. He also came up with this formula which would mathematically explain a relationship between the energy and the wavelength:

$$E = n \times h \times f$$

Where $n$ is a positive number called quantum number, $f$ is the frequency of the light produced, and $h$ is the famous Plank’s constant:

$$h = 6.626 \times 10^{-34} \text{ J.s}$$

This revolutionary idea caused many scientists to do further experiments to define the accuracy of this theory. One of this scientist was the German physicist, Albert Einstein. He was born in 1879 and died in 1955. In 1905, he represented his photoelectric effect experiment. In this experiment, Einstein shot a beam of light with very low frequency onto a metal plate, and as the result, light emitted electrons from the metal plate. This is shown in figure number 3. It seemed like light was including thousands of tiny packets where each of them had enough energy to knock of one electron from the metal plate and give them exactly the same energy as the little packets of light.

The experiment was repeated with higher wavelength lights such as red light and nothing happened which shows that photoelectric effect only happens in really high frequencies of light.
He also proved that no matter how high the intensity of light would go, it would not change the result of the experiment. He called these little packets of energy as photons. He also proved that the formula for the electrons emitted is:

\[ KE = h.f - \phi \]

Which \( f \) is the light frequency, \( h \) is max plank’s constant, and \( \phi \) is the work function of the metal, which is the minimum energy needed to knock the electron loose. It is different for each metal. After all these experiments, scientists come to the conclusion that light sometimes acts like a photon and sometimes act like a wave. They call this wave particle duality.

In the mid-nineteenth century, scientist did many experiments related to the definition of an atomic model. One of the most significant ideas about atomic model was Bohr atomic model.

In 1913, Niels Bohr started to work on the model for atoms. He was a physicist from Denmark who was born in 1885 and died in 1962. What people believed at the time was Rutherford’s model, which believed that atoms are composed of a proton in center and electrons circling around it like planets orbiting the sun. But there were a few problems with this model; one of them was that Rutherford's planetary atom should have an extremely short lifetime. Electron is a charged particle, and like any charged particle, it gives off electromagnetic radiation as it moves, and it loses energy as it does that until it eventually loses all its energy, which does not happen. Also when taking a close look at hydrogen spectra, it has descriit units in it; it almost looks like its quantized! So Bohr put these all together and came up with the Bohr’s atomic model. This atomic model is shown in figure number 4.

Electrons can be on energy levels around the protons. They can only be on one energy level because they are quantized. Light emits when an electron travels from higher to lower energy level, and the frequency of light can be calculated based on the Max Plank formula.

Bohr’s model explained atomic model in terms of quantum concept; but what mainly explained was hydrogen which was what Bohr’s model was based on.

These experiments led to very interesting hypothesis which was made by a French scientist Louis de Broglie. He was born in 1892 and died in 1987. In 1924; he hypothesized that if waves can act as particles, probably particles can act as waves also. In other words, all kind of matter has dual particle - wave nature. Therefore, everything has a wavelength; me, you, even a hippopotamus has a wavelength! He also suggested a formula for particles wavelength:

\[ \text{Wavelength} = \frac{h}{m \times v} \]

which \( m \) is the particles mass, \( v \) being its velocity, and \( h \) is max planks constant.

So a crocodile has a mass of 60 kilograms and travels with speed of 5 meters per second. Its wavelength would be:

\[ \text{Wavelength} = \left(6.636 \times 10^{-34}\right) \]
\[ \text{Wavelength} = 2.212 \times 10^{-36} \]
This means that crocodile does have a wavelength but it is very small comparing to its size.

One of the most important discoveries in 19th century was the uncertainty principle discovered by heisenbug. He was a German physicist who was born in 1901 and died in 1976. He received the Nobel prize for this discovery. He discovered that certain complementary quantities in physics such as position and momentum or energy and time are linked by some sort of uncertainty. In other words, it is not possible to be precise about the measurement of both of these complementary quantities at the same time. He based his idea on the theory of wave/particle duality of the matter. Imagine trying to observe a particle’s position and momentum very accurately using similar experiment as the Einstein’s photoelectric effect experiment. In order to knock the electron loose, a very high frequency photon is needed which is going to have very high energy, and it is going to give this energy to the electron. Therefore, the electron is not going to have a very precise momentum, but its position can accurately be measured!

Based on everything that was said; there are different probabilities of location of a particle at the same time. in the double slit experiment, light passes through the two slits and light waves interfere and the result is the light and dark pattern on the second screen. In fact, the photon does not split, the probability of its location will split.

Based on this interpretation, a German physicist Schrödinger made a very unusual hypothesize. He was born in 1887 and died in 1961. In 1935 he was working on a fairly unusual problem. Imagine a box with a cat in it. A radioactive material is placed within the box near a Geiger counter that would detect the radiation. There is a 50 percent possibility that the radioactive material starts to decay within an hour. When the Geiger counter detects the radiation from the material; it will break a glass of poison gas, so the cat will die.

According to the interpretation of quantum mechanics and uncertainty principle, after an hour the cat is dead and alive!

This is one the most confusing questions in quantum mechanics that leads one to the idea of parallel universes.

**Many Worlds Interpretation of Quantum Mechanics:**

There are many different explanations for this phenomena, one of the most famous ones is Many World Interpretation. Hugh Everest was the first person coming up with this idea in 1957. Many World Interpretation believes that there is an independent reality that is not dependent on the observer. In other words, the two possibilities will collapse because that is what the observer will see. What if there is no collapse between the possibilities; and every possibility just branches to different direction. So, when a cat is seen alive; it is just the possibility seen by a particular observer. The possibility of the cat being dead will also happen; it is just not seen by this observer. This theory states that every one of this possibilities happen in a different world. These different worlds are parallel to each other; so they won’t collapse.

Many World Interpretation has its own conflicts! One of them is; it completely denies the role of the observer. When an experiment is set up, the result of the experiment is based on the choice of
the observer. For example, imagine property A and B have been put in a system and the result is property C. If there were different inputs rather than B and C; there might have been different results rather than C. In double slit experiment, the observer chose what to put as the input and that is what causes the output of the experiment. In this case, it is hard to deny that reality acts independent of the observer’s choices. So there are not only many different worlds but also many different observers each with different minds; each mind making different choices which results in different outcomes.

One of the other problems with this theory was the Born’s rule stated by Max Born who was working with heisenbug on the theory of Schrödinger’s cat. He proved that the probability of one outcome happening over the other is actually measurable. Back to the Schrödinger experiment, the radioactive material can decay or not decay at the same time, which results in the cat being alive or dead at the same time. Many Worlds Interpretation stated that both possibility can happen at the same time just in different parallel worlds. Born proved that the probability will change over time and it can be calculated. So, if the probability of the radioactive material decaying or not decaying if equal after an hour; it might not be equal after two hours and it actually can be calculated!

**Conclusion:**

There were many remarkable discoveries that happened during the 19th and 20th century. The most important ones were dual functionality of particles, the uncertainty principle, and the possibility of the parallel universes. According to Schrödinger idea there is equal probabilities from something to happen or not happen. So when someone passes a red light, there is a 50 percent possibilities that an accident would happen and a 50 percent possibility that it would not happen. Based on Many World interpretation, both incidents will happen in two parallel worlds.

Regardless of all the problems with Many World Interpretation idea, so far it has offered one of the best explanations for the theory of quantum mechanics. This revolutionary idea has attracted a lot of attention during the past decades. Many movies have been made and many novels have been written based on this idea.

Imagine all significant moments in the life time of this planet, such as the ice age, the extintion of dinosaurs, and the mutation in the great African ape which resulted in *Homo sapiens* (Humans); what if it all did not happen in another parallel universe or did happen in another parallel universe just in a different time. For instance, imagine there is another world that the mutation in the great African ape happened at least a billion years earlier than our planet. Imagine how advanced they can be?! At least 1 billion years more advanced than earth! Or in a different universe Adolf Hitler would not born, maybe his parents never met! So, there is no World War II in that universe’s history and that economy is blooming! Or in another one, Germans actually did win the war and everyone is still living under the Nazi rules!

This revolutionary idea never has been really proven; in fact, as mentioned earlier there are pretty powerful arguments against it. But it is still one of the most exciting ideas of the modern physics, why is that?!
One of the reasons can be because of what Sigmund Freud calls defense mechanism. Everyone has painful memories and in order to deal with them people choose many different defense mechanisms. Some people completely forget them, some people direct it to other activities, some people become criminal, etc. How about considering the possibility that in a different universe those painful memories did not happen. Would not it be less painful?

The other reason can be regret of making the wrong choice. It is agonizing when thinking about other choices that could have been made. “I wish I would have taken school more seriously”, or “I wish I would have been nicer t my parents”, or “I wish I wouldn’t get married to this person” or something more severe such as “I wish I would have never used drugs”, or “I wish I would have never committed this crime”. Wouldn’t it have been nice thinking that there is another universe that only right choices have been made? A universe free of pain and regret?!

On the other hand, as mentioned before, so far multi universe theory has offered the best explanation for the phenomena of Schrödinger’s cat experiment. In fact, human knowledge still is very limited when it comes to solving problems such as quantum mechanics. There is not a certainty that what is seen is the actual reality. There might be many independent realities that men as observer cannot see or prove.

A German philosopher went to sleep and dreamt about being a butterfly. When he woke up, he asked himself: “am I a human dreaming about being a butterfly or am I a butterfly dreaming about being a human?”. So, maybe there is more time and technology needed to solve this important problem of physics and philosophy.
Figures:

Figure number 1: Double Slit Experiment with the machine gun (Anon. Double slit experiment. [Internet]. [cited 2016 Apr 20]. Available from: http://physics.mq.edu.au/~jcresser/phys301/chapters/twoslitexpt.pdf)

Figure number 2: Double Slit Experiment with light (http://ipodphysics.com/prop-of-light-youngs-double-slit.php)
Figure number 3: Photoelectric Effect Experiment (http://physics.tutorvista.com/modern-physics/photoelectric-effect.html)

Figure number 4: Bohr’s Atomic Model (http://chemistry.tutorvista.com/inorganic-chemistry/niel-bohr-atomic-theory.html)
References:


Lion ND. The Bohr atom. iun [Internet]. [cited 2016 Apr 20]. Available from: http://www.iun.edu/~cpanhd/c101webnotes/modern-atomic-theory/bohr-model.html


The Procedural Growth and Classification of Galaxies

Sabrina Krozel

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Dr. William Sherry
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Abstract

In today’s day and age, astronomers are now capable of determining the classification of galaxies. Edwin Hubble developed a galaxy classification scheme consisting of four types: elliptical, spiral, barred spiral, and irregular (Network, 2016). These categories of galaxies not only list out the specific traits that a galaxy may have, but also state the approximate stage of evolution that a galaxy is in. There are no definitive explanations on how galaxies were formed, but scientist have theorized different models on how they came to be (Martin, 2015). Though there is still more research and data needed to be collected to confirm how galaxies are formed.

Introduction

After the creation of the universe, it was a very hot, hostile, gassy environment. During the Radiation Dominated Era, the universe went under a subsequent cooling and fragmentation. Thus initiating the process of gravitational coalescence that dawned the creation of protogalaxies (Bothun, 1999). Over time, astronomers have systematically searched for rare cases of distance background quasars and a stream of primordial gases near a foreground of a galaxy with the use of cosmic web imaging and absorption line spectrum (Pössel, 2013). With the data gathered from these tests, it provided valuable information that astronomers used to interpret the chemical composition, density and temperature of the gas. “A team of astronomers led by Caltech has discovered a giant swirling disk of gas 10 billion light-years away—a galaxy-in-the-making that is actively being fed cool primordial gas tracing back to the Big Bang” (Martin D. C., 2015). This discovery was consider a “smoking gun” within the last couple decades. It helped scientist create detailed models of protogalaxies on how they gather enough gases to develop their disk and later on, produce stars. Since scientist have been debating on how galaxies formed and were unable to explain how they gather enough gas. In the standard model, the protogalaxy gathers the gasses from the intergalactic medium; which later on goes in a holding stage where it has a diffuse halo around the galaxy and falls back in. Another model, known as the cold flow model, gases from the intergalactic medium funnels-streams directly on to the galaxy (Martin D. C., 2015).

Galaxies come in many shapes and sizes and they can house to billions and even trillions of stars. Periodically over time, galaxies are known to collide with another galaxies within their local group. When this event occurs, stars are flung out of orbit, gasses from colliding galaxies births new stars. There are a wide variety of galaxies in our universe from elliptical, to bard spirals and many more. Around the 1920’s, Edwin Hubble devised the most commonly used system of classifying galaxies (Mastin, 2009). He categorized four types of galaxies as elliptical, spiral, barred spiral, and irregular. The classification chart is recognized as the “Hubble Tuning Fork” or “Edwin Hubble’s Classification Scheme”. The diagram, simplistic as it is, is divided into two parts: elliptical galaxies, and spiral galaxies. Although, Hubble noticed that some galaxies were difficult to put into context for the diagram. The chart does not include irregular galaxies which had odd shapes, dwarfs galaxies are too small, and giant elliptical galaxies which are very large elliptical galaxies that resides in the centers of some clusters of galaxies (ESA, n.d.).
Elliptical galaxies are nearly featureless, smooth, ellipsoidal shape. Most of the observable largest galaxies are giant elliptical galaxies since elliptical galaxies are the most abundant type of galaxies in the universe. They have a broader range of sizes than any galaxies, ranging from a small dwarf that can be ten percent less than the milky way, stretching ten million light years across (contain about over ten trillion stars) (Staff, 2013). It has little to no global angular momentum or structures, compared to the spiral galaxies. So stars are in somewhat in a strange orbit around the center (G, 2016). Not to mention, elliptical galaxies are composed of older, low mass stars with minimal star formation. In other words, elliptical galaxies are commonly found close to the center of galaxy clusters and are considered as “early type” galaxies due to their location (EG, wikipedia, 20164). The abbreviation for elliptical galaxies E0 (0, meaning the galaxy looks like a circle) through E7 (galaxy looking very long and thin) (EG, 2006).

Spiral galaxies primarily consist of a flat disc with an ellipsoidal form bulging center and a surrounding halo. Though the disks may not seem like it, they consist of a substantial amount of interstellar gas and dust and containing a significant portion of the stars within the galaxy. The materials and particles within the disk spin in one direction circling the galactic center moving at hundreds of miles per second forming a very noticeable and recognizable spiral formation. The bulge however can be found within the middle of the disk and contains a far more aged stellar population with a low concentration of interstellar matter. The more recent and younger stars within the disk are given the name of stellar population I, and the older bulge and halo stars as population II. The older bulge containing the massive amounts of older stars is assigned the name of being a Globular Cluster when the older stars are in a large clump. One other aspect of these galaxies to note is that the halo surrounding the disks of spiral galaxies are thought to contain dark matter. Though it is not just that it contains it, though more of the idea that it is believed that the halo contains a very large amount of dark matter. Dark matter being the unobservable matter within the universe which does act quite similar to “normal” matter, though it is extremely hard to detect leading us to know far more about what it is not, than what it truly is.

The spiral galaxies are classified through a quite straightforward system which focuses on observations made when looking at them through the Hubble scheme. The spiral galaxies with a quite noticeable bar structure going through the middle of them are called barred spirals and are given the abbreviation Sa or SBa. The Galaxies with very noticeable bulges or pronounced arms and spirals are given the classification of Sb, and others with less prominent features containing weaker arms, and a more acute bulge are given the classification of Sc. The arms of a spiral galaxy are extremely interesting as well when considering how they are formed. The arms are not permanently held structures but rather an ever flowing current of the density waves of the particles of dust, gas, and stars moving closer and closer, and then further and further away from the center. This ebb and flow the density of these materials are the cause of the beautiful and magnificent spiral arms from the spiral galaxies (Cain, 2014). Along with all of this, the core of some spiral galaxies contain a massive black hole that consumes the surrounding gasses. Though that is the same for what is believed of all galaxies, that they all in some way contain black holes at their center. These black holes suck material into themselves though some of it is expelled from it quite rapidly, sometimes even to the speed of light. “These jets have been observed most spectacularly from the
centers of nearby galaxies but also appear in micro quasars. They are quick, and have enormous energetic spurts and sputters, as if someone had taken a video of a quasar jet and pressed the fast forward button (Cornell.edu, 2016).”

Type one Irregular galaxies are any galaxy which does not fall into a specific pre-defined category of identifiable features such as being a Spiral or Elliptical. Their lack of uniqueness is not where the differences end though. The irregular galaxies are considered primitive in some regards due to their low levels of heavy elements. This is very much so “In contrast to galaxies like the Milky Way which are rich in these elements, which have been manufactured by stars in a process called nucleosynthesis” (Bannister, 2001). Irregular galaxies also have a quite high density of hydrogen which is the cause of their very noticeable glow forming the luminous nebulae when heated up by surrounding stars.

Type two irregular galaxies are awe inspiring and a sheer spectacle to look at. There are multiple ways which they can take form. A more usual way for them to form is through a gravitational interaction with other close by galaxies. This may actually seem like a statistical improbability due to the miniscule likelihood of such an event ever happening. Though this is through the assumption that every individual galaxy and cluster of starts are actually evenly spaced which they are very much so not. Galaxies however usually coexist within clusters of themselves with the most common spacing of these clusters being far closer. Because of this the chances of collision happening between separate galaxies is far more possible. These collisions then allow for the creation of the amazing and spectacular irregular galaxies which we are able to view through telescopes. “The Hubble Space Telescope imaged a very unusual galaxy known as the Cartwheel. This is also a product of galactic collision. In this case, a small galaxy (which may be one of the objects on the right of the ring) passed through the middle of the main spiral galaxy, causing the compression of gas and dust. The ‘wave’ produced then moved towards the outside edge of the galaxy, leaving newly formed stars in its wake. It is estimated that the billions of stars were created in this collision (Bannister, 2001).” There are also a couple irregular satellite galaxies within the Milky Way which are known as the Large and Small Magellanic clouds. These galaxies are so bright in fact that they can even be seen with just a person’s eyes from the southern hemisphere of the earth.

Conclusion

In the end even with all of the information that the scientific community has been able to muster, all of the data is still just not enough to be able to accurately assess the unique and specific process in which protogalaxies are actually formed. There have been extraordinary leaps and bounds covered on the topic by astronomers though it is still a rigorous and imposing process. However websites like “Galaxy Zoo” are able to outsource the initial stage of first looking at different galaxies which allow for scientists to speed up the procedure of categorizing all galaxies observed through the Hubble telescope. It includes the full spectrum of Irregular, Spiral, and Elliptical along with multiple different variations and options for classifying them more specifically than before. The future looks quite bright and promising with the advances in technology, the improvement of telescopes, and the overall quality of information that we have at our disposal at this point in time. Though in the end of this, we still need to figure out just how exactly this all takes place without any holes or important aspects of this process left out.
FIGURES
References


https://www.le.ac.uk/ph/faulkes/web/galaxies/r_ga_irregular.html

http://ned.ipac.caltech.edu/level5/ESSAYS/Bothun/bothun.html

http://www.universetoday.com/110929/why-do-galaxies-have-arms/


www.spacetelescope.org: https://www.spacetelescope.org/images/heic9902o/

http://curious.astro.cornell.edu/the-universe/galaxies

Martin, D. C. (2015, August 5). *A Giant Protogalactic Disk Linked to the Cosmic Web - C. Martin - August 2015*. Retrieved from youtube.com:
https://www.youtube.com/watch?v=XaXrYLI1oul


http://www.physicsoftheuniverse.com/scientists_hubble.html

https://lcogt.net/spacebook/galaxy-classification/

http://news.ucsc.edu/2013/10/cold-flow.html

http://www.space.com/22395-elliptical-galaxies.html
Abstract

The purpose of this paper is to explain how physics perfects the sport of skateboarding. History of the skateboard shows how the shape of the board takes shape. Understanding the history of the skateboard allows for the physics to represent the perfection of the board. One important factor was the invention of the polyurethane wheels, which allowed for the popularity of the sport to soar. The perfections of the board include factors of forces, speeds, and energies. These factors allow for the skater to perform tricks such as the “Ollie.” As history continues to progress so will the model of the skateboard, such as the concept of the hover board.

Introduction

History of Skateboards

Imagine a pair of metal skates that contain three wheels and an adjustable toe and heel clip to strap a foot in. Along with these skates are a pair of poles, added to push these skates along mimicking cross country skiing. This was the first known device (pic. 1) to resemble a skateboard and it dated back to the 1920’s. An interesting idea to say the least, but it lacked in steering and brakes. Not more than ten years later another device called the Scooter Skate (pic. 2) was on the market. This skateboard/scooter hybrid allows the skater to ride the Scooter Skate with a handle or without a handle. Once again this device stood on three wheels and lacked steering. Brakes however were applied by the skater’s foot, an improvement from the previous skateboard. Eventually the skateboard we know today started to take form in the 1940’s. This 1940’s four wheeled device was called the Skeeter Skate (pic. 3) and it was composed of aluminum. This skeeter skate came with a removable handle and pedal-car like wheels (similarly to the 1930’s metal skates). Unlike the previous skateboard inventions this Skeeter Skate had its first ever steering axles, or what we call trucks. This allowed the skater to turn their Skeeter Skate for the first time, which was not possible with the previous models. Within two decades the skateboard had much ground.

The 1950’s is when the idea of the skateboard really took flight. California surfers took their surfboards to the streets inventing skateboarding, or what the kids called it during their time, “sidewalk surfing”. These surfers found much pleasure in being able to “ride the streets” when the waves were just not right. Many helped invent this idea of “riding the streets,” creating wood boards with roller skate wheels attached to the bottom of the board (pic 4). Moving into the 1960’s, clay wheels were introduced to skateboarding (1959). Skateboarding was all the rage in the 1960’s and the popularity of the sport took off at its peak. Half way through this decade all the rage seemed to die overnight. The fad of skateboarding was seemingly over. There were a few who stayed true to their sport and continued to make homemade skateboards but as for the rest skateboarding became a thing of the past. Many speculate that the clay wheels on the boards were far too dangerous and caused many injuries during it time in fame, causing the popularity of the sport to die.

There were many small business owners that took an interest this idea of skateboards during the 1960’s. Those by the names of Val Surf, Larry Stevenson, Hobi Alter, Larry Gordon and Floyd Smith, Skip Engblom, Jeff Ho and Craig Stecyk, and many more were creators, promotors, and perfectionists of the popular skateboards in the 1960’s. These inventors created the version of the skateboard that introduced three inspiring components: Clay wheels, double action-adjustable trucks and board “that combines Bo-Tuff (a fiberglass reinforced epoxy) with a maple wood core to create the Fibreflex skateboard” (Martin 2015) or otherwise stated, the first laminated board created (pic 5). These models of the skateboards created in this decade were the models that resemble the skateboards we have come to know today. Few changes to the
skateboard were made as the years passed, but the single most important invention made the skateboard have a comeback. In 1975, this invention changed what was so dangerous about the sport, clay wheels, into something more reliable. Replacing the clay wheels was the invention of urethane wheels (pic 6). This invention allowed for fewer injuries. Another revolutionary event that changed skateboarding forever was something called the “Ollie.” In 1978 a man named Alan Gelfand (nicknamed Ollie) invented a skateboarder maneuver where “He would slam his back foot down on the back of his board and jump, thereby popping him and the board into the air” (Martin 2015).

Skateboards only continued to progress. This progression included the production of different types of boards pertaining to the stiffness and shape of the board. For those who loved the sport continued to promote it by building their own ramps and making vert skateboarding popular. Then in 1995 ESPN held their first ever X-Games and forever brought skateboarding into the mainstream light of the world. Now here in the 2000’s stake parks popped up everywhere and many companies emerged and took part in this new sport of skateboarding.

Discussion

How a Skateboard Works: The Basics

A skateboard is easily described as a curved wooden board which four wheels and two trucks attached to the underside of the board (as shown in Fig. 1 and 2). Each skateboard is set up with a fixed set of variables or constants (see Fig. 1 and 2). Let

- \( \theta_f \) be the angle the front-truck makes to the board.
- \( \theta_b \) be the angle the back-truck makes to the board.
- \( AB \) be the distance from \( A \) to \( B \).
- \( CA \) be the distance from \( C_A \) or \( CB \), this is the height of the trucks.
- \( 2w \) be the width of the truck. This is measured from the end of the wheel-axis.
- \( P \) be a point on the board where velocity is strictly in the forward direction of the board.

(Osterling 2004)

To find \( P \), an inertial coordinate system will be used. An inertial coordinate system is found on the board so that when the board is tilted and moved forward \( d \), it will rotate around the origin of the inertial coordinate system. To demonstrate this, \( P \) will be placed a distance \( P \) from \( A \) (near the front truck) and let’s say the origin is a distance from \( P \) to \( A \), making it the front turning point of the board. Looking at Fig. 3, the equations from the front turning point would be:

\[
\Rightarrow \frac{y_1}{P} = \frac{y_2}{AB - P} \\
\Rightarrow \frac{y_1}{y_2} = \frac{P}{AB - P} 
\]

(Eq. 1)

After looking at Fig. 3 and understanding that the front truck will have a \( dx \) in the x-direction and a \( dy_1 \) in the y-direction, and back truck will have a similar direction of \( dx \) and \( dy_2 \). Each set of distances for each truck will move slightly and similarly. Now that there is an understanding of where the distances will come from, the angle for these distances will be represented as:

\[
\Rightarrow dx = (dy_1 \tan \delta_f) \\
\text{and} \\
\Rightarrow dx = (dy_1 \tan \delta_b) 
\]

After equating these, they can be substituted into the Eq. 1 above to get:

\[
\frac{dy_1}{dy_2} = \frac{\tan \delta_f}{\tan \delta_b} = \frac{P}{AB - P} 
\]
\[ \Rightarrow P \tan \delta_b = (\overrightarrow{AB} - P) \tan \delta_f \]

Thus, 
\[ P = \frac{a \tan \delta_f}{\tan \delta_f + \tan \delta_r} \]

This equation allows point P to be found on the skateboard and a central line (C.L.) can then be placed on the trucks to allow movement when velocity is applied (fig. 1). This central line (C.L.) will be placed through points A, P, and B. Assuming that the skateboard uses the same trucks for the front and the back of the board the angles \( \theta_f = \theta_b \), thus \( P \) will equal \( \frac{\overrightarrow{AB}}{2} \). Now that the \( P \) and the central line has been found the pivot axle (p.a.) can be placed on the truck at points CD. The pivot axle of the truck will pivot when velocity is applied to the board at point P. This allows the board to turn and move at the slightest velocity applied to it. (Osterling 2004)

Polyurethane Wheels

Now that the basic of the skateboard has been explained, the wheels also play an important part in the physics of the board. The wheels of a skateboard have come a long way, starting out with petal-car like wheels, to clay wheels, and now to a plastic wheel called polyurethane (look at picture 7). The applications of these types of wheels allow the skateboard to have movement with ease while connected to the trucks of the board. The wheels simply move to the pivot of the trucks as explained above. The urethane wheels allow for abrasion resistance, impact and shock resistance, cut and tear resistance, flex fatigue resistance, high mass capacity, uses on multiple friction type surfaces, mold and mildew resistance, and many more applications.

This type of plastic or rubber allows for deflection under high stress impacts such as doing tricks on a skateboard. Even after high compression situations the wheel will recover its original shape when the compression is removed. This type of plastic or rubber meets the demands that skateboarding requires because skateboarding is a sport that puts the wheels under high stress situations and the wheels bounce back unlike their previous model made of clay which eventually would break or wear down. Overall the application of the polyurethane allows the skateboard to become more reliable and safer to though to ride it.

Forces Involved

Skateboarding involves many forces acting on the board or the skater of the board. A few forces will be explained in this section, such as gravitational force, a normal force, and a frictional force. A gravitational force is the force that gravity and the mass or weight (mass denoted as \( m \)) on top the board exert on the board. “The constant acceleration of a freely falling object is called the acceleration due to gravity…At or near the earth’s surface, the value of g is approximately 9.80 m/s², 980 cm/s² or 32.3 ft/s².” (Young 2012). Thus the force applied on the board due to gravity is the following equation:

\[ \Rightarrow \text{Force} = \text{mass} \times \text{gravity}. \ (F=mg) \]

The normal force can also because a contact force. The normal force is the force of the object at rest on the surface. In this case the normal force is the skater resting on top of the skateboard, exerting a perpendicular contact force. The usual denotation for the normal force is \( n \). The next force is called the frictional force. This force most often acts against the sliding or rolling of an object on a surface. In this case there are two the frictional forces. First frictional force is the skater on the skateboard. The shoes of the skater act to resist the grip layer of the skateboard, in turn this friction is helping the skater stay on the board. The second frictional force is the wheels acting on the surface it is rolling on. The frictional force acts in the opposite way the skateboard is accelerating, whether up a hill, down a hill, or on a straight path accelerating forward. Friction
is denoted by using $f$. In Fig. 4 an example of a skater on a ramp (in parts) will explain how these forces act on the skateboard and the surface it is on.

**Speeds Involved**

Skateboarding involves two main speeds: velocity and acceleration, fig. 4. Velocity is defined as distance traveled over time. Its units are meters/seconds (m/s). A skater will start at an initial velocity of zero and as the skater travels a certain distance over time they will reach a final velocity. Acceleration is defined as an increase rate of speed over time. Its units are m/s$^2$. As the skater increases its speed by pumping his foot on the ground or going down a hill, the velocity increases over time creates an accelerating skater and have a constant acceleration (assuming there is no air resistance on the skater). If air resistance is present acceleration will not be constant and velocity by pumping of the foot or going down a hill will be needed to increase the rate of the speed of the skateboard.

**Energy Involved**

There are two types of energy involved with skateboarding: Potential and Kinetic Energy. These two energies are present when the skater is going down or up an incline but not on a flat surface. Potential energy is defined as “The energy which an object possesses by virtue of its position…” (Farrow 2006). Thus when the skater takes their board to the top of a hill or ramp the skater and skateboard will have had work done on them and will have gained potential energy in the process, fig. 4. The equation for potential energy ($U_g$) is as follows:

$$\Rightarrow U_g = mgh \text{ (mass x gravity x height)}$$

Kinetic energy is defined as “energy which an object possess by virtue of its motion” (Farrow 2006). This means that when the skater goes down a hill or a ramp the skateboard is in a sense is released and begins to make its way down to the bottom of the hill or ramp, and acceleration increases by the force of the gravitational pull of the earth. As the skater reaches the bottom of the hill/ramp the kinetic energy has reached its maximum point and proportionally the potential energy has reached its minimum, fig. 4. The equation for kinetic energy is as follows:

$$\Rightarrow K = \frac{1}{2}mv^2 \text{ (mass x velocity)}$$

**Newton’s Laws of Motion**

Skateboarding is all about motion, without motion a skater could never land a jump. “Motion is how, where, and why something moves” (Greathouse 2009). In the sport of skateboarding Sir Isaac Newton’s Laws of Motion play a big role. There are 3 laws that support the idea of skateboarding.

The first law of motion states “that an object will keep doing what it is doing unless an outside force acts on it” (Greathouse 2009). In the case of skateboarding, a skateboard will not move unless a force such as a skater or another force like wind pushes the board. The second law of motion states “that when a force acts on an object, the object will speed up or slow down” (Greathouse 2009). If a skater pumps the skateboard along the ground with their foot they are exerting a force to speed up the acceleration and velocity of the board. If the skater uses his foot to stop the skateboard then they are also exerting a force to slow down the acceleration and velocity of the board. The weight of the skater also has an effect on the force exerted on the board. The heavier the weight the more force is need to be exerted on the board to make it move. The third law of motion states that “for every action, there is an equal and opposite reaction” (Greathouse 2009). The wheels on a skateboard exert a force on the road as they spin and will push the road in a backwards direction. The road has an equal reaction, so that the road pushes the wheels forward. This equal and opposite reaction of the road and wheels allows the skateboard to move forward when a force is exerted on the board.
The “Ollie” that Involves Physics

The “Ollie” was invented in the late 1970’s by a man named Alan “Ollie” Gelfand. The “Ollie” paved the way for more complicated tricks. The “Ollie” is the most basic trick in the sport of skateboarding. This trick allows skaters to hop a curb or an obstacle with style. The “Ollie” is performed by the skater pushing down on the board and jumping up, allowing the skateboard to follow the skater into the air, as if the board never left the skater’s feet. Although the skateboard does leave the skaters feet even though it does not look like it, that is the miracle of the “Ollie” trick. This trick’s secret is that it involves rotation around multiple axes, allowing the trick to be possible. There are many forces exerted on the board to perform an “Ollie.” Look at Fig. 5 to understand what an “Ollie” looks like step by step before continuing.

Just before the “Ollie” is performed there are 3 forces exerted on the board. The first picture in Fig. 5 shows three arrows being exerted on the board. The green arrows are the weight of the rider on the board. The blue arrow is the force created by gravity on the board itself. Lastly the red arrows are the force of the ground pushing up on the board. Since the trick has not been performed yet, the net forces of all these forces are equal to zero. This zero net force means that the skateboard is not accelerating, but is rolling at a constant speed. This step in the trick, the skater is in a crouching position to create a low center of mass, which will be important to get air to the jump.

In the second picture in Fig. 5, the skater is accelerating upwards. The skater will straighten the legs and raise the arms to create a more dynamic jump. The skater will use their back foot to exert a greater force downwards on the rear of the board than the front foot, allowing the front of the board to pop upwards and pivot in a counterclockwise motion. The third picture of Fig. 5 the rear of the skateboard hits the ground. This hit on the ground creates a large force upward on the rear resulting in the board to bounce up and pivot clockwise around the center of mass (the skater). The fourth picture of Fig.5 shows the skateboard completely in the air from the resulting force from the previous picture. While in the air the skater’s front foot slides forward and the friction between the skater’s foot and the grip of the surface of the board creates a more upward vertical angle even higher in the air.

In the fifth picture of Fig. 5 the skateboard begins to level out. This leveling out is created when the skaters front foot pushes down on the board raising the rear wheels of the board. While this leveling out is taking place the skater lifts his back leg and foot up to keep it out of the way of raising the rear of the board upwards. If the lift is performed correctly, the rear foot and the rear part of the board will move simultaneously making the board look as if the feet of the skater are stuck to the board. In the sixth picture of Fig. 5 the “Ollie” is now at its max height. Both feet are touching the board and gravity begins to take affect by pushing the skater and the skateboard back down to the ground together. Lastly in the seventh picture of Fig. 5 the full extent of gravity has taken affect and the skater and the skateboard hit the ground horizontally. To absorb the impact of the gravity taking place the skater bends their knees upon landing. The result is called an “Ollie.”

MegaRamp Used In X Games

There are many types of ramps used in skateboarding. Skate parks are full of ramps for a skater to use. It gives the skater a challenge besides just riding on a flat surface. In the X Games (competition for pro-skaters aired on TV) they use what is called a MegaRamp (look at Fig. 6). A MegaRamp is nine stories tall and longer than a football field. Before a skater starts they can either choose between two drops: a 65 or 80 foot drop off a platform. This type of drop is used by pro-skaters in competitions to gain enough speed to soar across a 60 to 70 foot gap, landing in
a 27 foot quarter pipe where these skaters can begin to do their tricks and be judged by their performance.

This type of ramp puts all the components about skateboarding (velocity, potential and kinetic energy, acceleration, gravity, etc.) previously talked about into play. On top of the platform before the skater drops potential energy is at its greatest because the work had been done to get the skateboard to the top of the platform, in turn kinetic energy is zero. As the skater drops off the platform the potential energy decrease and the kinetic energy increases proportionally. At the lowest portion of the ramp kinetic energy is at its greatest and the potential energy is at its lowest (or possible at zero). At this point acceleration is at its max and as the skater continues back upwards toward the end of the ramp the kinetic energy decreases and the potential energy again begins to increase, thus resulting in the acceleration of the skater slowing down in midair. As the skater soars across the 60 to 70 foot gap the potential energy is once again at its peak (max height the skater gets in midair) and the kinetic energy decrease (if not to zero). As the skater comes in for a landing in the quarter pipe the potential energy again decreases and the kinetic energy hits its max once the skater gets to the middle of the quarter pipe (that’s if the skater doesn’t lose control and executes the landing). As the skater comes to the end of the ramp and flies up to 50 feet into the air the potential energy is at its greatest once again and the kinetic energy is at its lowest. The skater then returns to the quarter pipe and the potential and kinetic energy proportionally increase and decrease as the skater skates back and forth on the quarter pipe. During this time the skater has enough speed and air to do tricks with their board.

Gravity is the main component in using the MegaRamp. Gravity accelerates the skater up to speeds of 44 miles per hour on the drop off the platform. Gravity allows the skater to reach max speeds necessary to execute a landing on the Mega Ramp. Although if the skater does not get enough speed to clear the 65 to 70 foot gap toward the quarter pipe can result in injury or a potential fatality, in this case gravity can hurt the skater’s potential to execute the landing within the quarter pipe. Thus gravity is a very important component for the skater to perform a perfect jump, if all the factors about the jump are executed correctly.

This MegaRamp is an extreme version of a ramp for non-pro skater. It represents the easiest example to explain all the factors that are present in skateboarding. Even though only a few factors were explained above, all factors will play a role in this sport of skateboarding.

Conclusion

This research paper was to explain all the forces, speeds, and energies that are put into the sport of skateboarding. First the history of skateboarding helped us understand how skateboarding came to be. Skateboarding took many forms over time, and it took trial and error to get the mechanism of skateboarding just right. An example of trial and error, the board endured changing of the material of the wheels on the board. The wheels started out with pedal-car like wheels, then to metal, then clay, and finally polyurethane wheels. The urethane wheels allowed for fewer injuries thus making the sport of skateboarding a great sport, especially among the younger crowd. Once the shape of the skateboard was perfected then all the factors came together to create an incredible sport enjoyed by all.

In this paper many factors came into play to make the skateboard possible. These factors included forces exerted on the board such as mass, gravitational force, and a normal force. These forces affected how the board moved with velocity and acceleration. Along with these forces, Newton’s Laws of Motion worked with the forces to explain how motion on the board is either created or destroyed. Lastly, potential and kinetic energy play a role when a skater uses
their board on a hill or a ramp. The potential and kinetic energy are proportional to each other, as one increases the other decreases and vice versa. These energies especially come into play when gravity, acceleration, and velocity work together to give the skater enough speed to do tricks just like the “Ollie.”

The skateboard has come a long way throughout history and I believe will continue to take new forms and shapes to create an even greater sport to love. The invention of the hover board has always been an idea to be a model of the future. It is modeled after the skateboard in a futuristic way. The idea of the hover board has begun to take shape into a real and touchable model but has not been perfected yet. The hover board is still a working progress but I believe, in theory, it will be the next step into the progression of a new model of the skateboard.
FIGURES

Picture 1
http://www.skateboardingmagazine.com/1920%E2%80%B2s-the-first-thing-that-resembled-a-skateboard/

Picture 2

Picture 3
http://wackyboards.blogspot.com/2011/03/1940s-jet-boy-scooter.html

Picture 4

Picture 5

Picture 6
http://www.tailtap.com/cadillacwheels.html
Initially

After A Short Time $dt$

Figure 1 and 2

Figure 3
Figure 4

Figure 5

http://exploratorium.edu/skateboarding/trick.html
Figure 6
http://www.personal.psu.edu/pai5010/cadproject.html
References


The Life Cycle of Large Mass Stars

Kelley Long

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Dr. William Sherry
The life of stars is something not many understand. Even though there are billions of stars in the galaxy, no one considers how they operate and go through life. In this paper, the life cycle of large mass stars will be described in detail. Starting with how any and all stars first get their start and how long it takes for them to develop into an actual star. The large mass stars will then spend most of their life living on the main sequence fusing together different forms of gases and emitting energy. Finally, a large mass star will reach the end of its lifespan and end dramatically before leaving behind remnants for future stars.

The bigger they are the harder they fall. The same thing can be said for large mass stars. Contrary to many beliefs, large mass stars have a shorter life span than that of low mass stars. It all has to do with the amount of matter contained in the nebula, which is a giant cloud of gas and dust, when the star is first being formed (1). The nebula contains gases such as hydrogen and helium and charged particles known as plasma (2). Often called “stellar nurseries”, the nebula is where any and all stars are formed. To first understand how a large mass star is formed, it is important to know how the nebula is formed. Nebulae are often extremely large in size and can span several hundred light years across (2). They form when the interstellar medium begins to endure gravitational collapse. Some nebulae are also formed when a supernova explodes thus leaving behind remnants of gas and dust that are later ionized from the energy of the explosion of the supernova.

Once the nebula is formed, the birthing process of a star can begin. Large mass stars and low mass stars often form the same way except the formation of large mass stars occurs much more rapidly than that of a low mass stars (3). The mass of a star also depends largely on how much matter is available inside the nebula. The first stage in the formation of stars is the force of gravity acting on the hydrogen gas, pulling it in and spinning it around. Once the gas gets spinning fast enough it begins to heat up and forms a protostar (1). A protostar looks very similar to that of a real star however its core is not yet hot enough to fuse together elements (5). When the protostar’s core reaches 10 million Kelvin, the fusion process is ready to begin. Both types of stars have a hydrogen fusion core but the hydrogen fusion in a large mass star occurs in the CNO cycle. The CNO cycle, which stands for carbon, nitrogen, and oxygen, is a fusion reaction where stars convert hydrogen to helium (3). This cycle is a more dominant source of energy only in larger mass stars because of the temperature at which large stars form. Once the hydrogen has been used up, a helium core forms with a hydrogen shell, then a carbon core with helium and hydrogen shells (3). After these cores and shells form, gravity comes into play and shrinks the core causing the temperature and pressure within the core to rise leading to the formation of hydrogen burning and a core that fuses helium into hydrogen (4). (This process can be seen in figure 1.) In a large mass star, when one form of fuel is used up, a new fusion process begins when the temperature and pressure are just right to start it. As each fusion process occurs, the time between becomes less and less. The fusion of helium into hydrogen can take up to several hundred thousand years and the next process could take only a couple hundred of years and so on. In the end, a large
mass star ends up being so large because as each layer is fused together, it forms heavier and heavier elements (see figure 2). Unlike low mass stars, a large mass star will glow much brighter because of the high core temperature and the rapid fusion rate (6).

Eventually a large mass star will move from the beginning stages to the main sequence stage. A star on the main sequence will spend most of its life in this stage, even though it takes millions of years to form, a star can remain on the main sequence for billions of years depending on its mass. During this stage, a star focuses most of its time turning hydrogen into helium inside of its core (7). A large mass star can stay in the main sequence stage for several million years before it uses up all of its energy. The size of a star will remain relatively the same size throughout the entire main sequence stage due to the function of its mass. While on the main sequence, large mass stars are found towards the top and glow a bright blue but don’t have as high of a temperature as the low mass stars found at the bottom of the main sequence and glow a brilliant red. This diagram can be seen in figure three. The bigger the star, the more the force of gravity acts on the star pulling the gases inward. As the gravitational pull increases, so does the amount of pressure in the core (7). The star is unable to release any pressure until it reaches the fusion stage. From there, the star is able to fuse hydrogen thus releasing energy that was built up inside the core. Energy is released through three different ways, radiation, convection, and conduction (8). When energy is released through radiation, what actually occurs are photons carrying the energy away from the core of the star (8). When energy is released through convection, gas that rises in temperature compared to the gas surrounding it, rises, expands, and then cools. It is rare for stars to release energy through conduction but when they do, it is when electrons collide together, thus releasing energy. For large mass stars, the main form of energy releasing is through convection. Eventually, a large mass star will exhaust its fuel and it will do so more quickly than low mass stars because the core of a large mass star is much hotter therefore it burns through its fuel supply quicker. Once the large mass star has used up its fuel, it begins to move off of the main sequence and enter into the “dying” phase.

The death of a large mass star could be considered rather dramatic compared to low mass stars. They tend to burn very bright and go out with a bang. Even after a large mass star has used up all the helium fuel in its core, it continues to go through the nuclear burning cycle (9). The core inside a large mass star will begin to heat up thus being able to burn carbon into oxygen, neon, silicon, sulfur, and iron (9). Once the core has reached the iron, it begins to collapse on itself. Eventually the core will resist further collapse and throw the matter back out causing a massive explosion of the star called a supernova. The explosion of a large mass star can last up to one month, making it the brightest star in the galaxy because as it explodes, it emits gases such as carbon, silicon, oxygen, and iron into space. See figure four for example of supernova example. Eventually these gases will be used to make newer generations of stars throughout the galaxy. When a large mass star explodes it can also leave behind neutron stars and black holes (10). A neutron star is considered very dense and having a high mass but with a diameter not exceeding 20 km (10). A black hole is a strong force of gravity where nothing can escape from it,
including light. Black holes often distort the space around them and in some cases act like a vacuum, sucking surrounding space matter into it.

In conclusion, the life cycle of a large mass star is a complicated and complex cycle. Even though large mass stars are considered to have a far shorter lifespan than that of low mass stars, they still remain in the galaxy for a couple billion years. Starting from birth, the mass of stars is determined from the start based on how much matter is inside the nebula. Gravity also plays a large role in the formation of all stars. Without the force of gravity acting on the gasses and elements in space, the formation of stars would not be possible.

Large mass stars also go through the fusion process much more rapidly than that of low mass stars, mainly due to their massive sizes. Where a low mass star could spend billions of years on the main sequence fusing together elements, a large mass star can only spend several million to a billion years on the main sequence fusing together elements. Because of the high fusion rate, a large mass star reaches the end of its life span with an enormous explosion that lights up the entire galaxy for a short period of time before it officially dies and either turns into a neutron star or a black hole but not before it leaves behind remnants of gases for future generations of stars.
Figure 3

Figure 4
References


Memristive Technology and how it is Shifting the Electronic Era

Author
By Mark A. Macluskie

Date
11/19/15

Class Name and Number
General Physics II (PHY112) 12834

Professor
Dr. Casey Durandet
Abstract

Since the discovery of the transistors, scientists have been constantly trying to improve this technology but all that they have really done is make transistors smaller and added more of them to improve performance. The path that the industry is currently on must change course or will come to an abrupt end. The memristor may help scientists save themselves from this fate. The memristor (which is short for memory resistor) was originally theorized back in the 70’s and has recently been successfully prototyped. Its current future is unknown but Hewlett-Packard is looking to completely reinvent the computer with this new found technology in a program called, “The Machine.”

Body

Humans have always strived to innovate, to create, and invent. It is a part of human nature to do so. It is what drove people like Isaac Newton to discover the universal laws of gravitation or Alexander Graham Bell to invent the telephone. It is what inspired William Shockley, John Bardeen, and Walter Brattain to invent the transistor. Humans have a tenacity and drive which propels the human race to accomplish amazing and outstanding inventions. It is these qualities that make science and technology so exciting, scientists are constantly discovering new things and improving older ideas.

One of those ideas, the transistor, is one of the reasons that the electronic devices like smartphones or computers exist today. The transistor was invented in 1947 in AT&T’s Bell Labs by William Shockley, John Bardeen, and Walter Brattain. They found that by touching two gold point contacts to a piece of germanium, a signal was produced with a greater output than input. The team continued to work on this observation and eventually were awarded the 1956 Nobel Prize in Physics, “For their research on semiconductors and their discovery of the transistor effect” (Nobel Foundation, 1). Today transistors serve two main purposes, amplify, and as a switch. As an amplifier, it is used in anything from mobile phones to televisions for sound reproduction. When used as a switch, it can give off an “on” or “off” signal which is useful in digital circuits with logic gates.

Transistors are a huge part of modern day electronics. Since the debut of transistors companies have constantly strived to improve them. But the only ways scientists have currently sought to improve transistors is to make them smaller and add more. This is fine but only up to a certain point. Transistors can only get so small and there is only so many they can add. In an article posted by Stanley Williams he said, “The semiconductor industry’s obsession with the shrinking of transistors and their commensurate steady doubling on a chip about every two years, has been the source of a 50-year technical and economic revolution. Whether this scaling paradigm lasts for five more years or 15, it will eventually come to an end.” These “improvements” cannot continue indefinitely and the only solution to this problem is to invent something new, something better. This is where the memristor come in.

Combined with transistors in a hybrid chip, memristors could radically improve the performance of digital circuits without shrinking transistors. Using transistors more efficiently could in turn give us another decade, at least, of Moore’s Law performance improvement, without requiring the costly and increasingly difficult doublings of
transistor density on chips. In the end, memristors might even become the cornerstone of new analog circuits that compute using an architecture much like that of the brain (Williams, 1).

In circuits, there are three main components known as the fundamental elements: Resistors, Capacitors, and Inductors. Resistors impair electrical current to regulate the flow of a circuit, capacitors store electrical energy in an electric field and like to maintain constant voltage, and inductors store electrical energy in a magnetic field and like to maintain a constant current. We have known about these devices since the early 1800s and they are very useful but have their limitations. One man, in 1971, named Leon Chua, theorized of a fourth fundamental element of a circuit: the memory resistor or “memristor” for short. His theory said that he could relate current and magnetic flux (Fig. 2) and create a new type of resistor. One that could not only regulate the flow of the current in the circuit, but could also be used for nonvolatile memory storage.

In the current era of electronics, there are two main types of storage: hard disk drives (HDD) and solid state drives (SDD) or flash memory. HDDs offer large storage capacities at low costs but are insufficiently slow. To solve this problem, computers often have a small amount of random access memory (RAM) which is very fast but is a volatile memory source meaning that once the power is turned off, anything stored on RAM is long gone. However HDDs and SSDs are nonvolatile memory storage meaning that they will always keep what is stored on them even without power. Computers use the two types of memory in conjunction where it uses RAM to store what it needs for now and in the near future, and stores long term files on the HDD. For example, when typing in a word document, the words that are being typed are being stored on RAM because it would take too long to save every letter one at a time on the HDD, but when you save is when it is stored in long term storage. SSDs offer an extremely fast storage option but at the same time, are still relatively new so they are expensive and come in low storage options compared to HDDs. SSDs are also smaller due to the fact that they don’t need a physical disk and needle to “write” the information down. Instead SSDs store information digitally without the need of mechanical moving parts. Most desktop and laptop computers come with an HDD and RAM whereas smaller electronics like smartphones and tablets come with SSDs. What memristors would do is add a fourth option to this list.

Unlike HDDs, SDDs, and RAM whose only purpose is to store information, memristors can store large amounts of data and also act as a standard resistor. “Theoretically, Memristors, a concatenation of “memory resistors”, are a type of passive circuit elements that maintain a relationship between the time integrals of current and voltage across a two terminal element. Thus, a memristors resistance varies according to a devices memristance function, allowing, via tiny read charges, access to a “history” of applied voltage” (memresistor, 1). Basically this means that the resistance of memristor changes as a function of the voltage applied to it and in what direction. The “mem” part of it, is that it remembers what the last voltage that was applied meaning that it offers a form of nonvolatile storage. Stanley R. Williams has a great analogy for this phenomenon:

Think of a resistor as a pipe through which water flows. The water is electric charge. The resistor’s obstruction of the flow of charge is comparable to the diameter of the pipe: the narrower the pipe, the greater the resistance. For the history of circuit design, resistors have had a fixed pipe diameter. But a memristor is a pipe that changes diameter
with the amount and direction of water that flows through it. If water flows through this pipe in one direction, it expands (becoming less resistive). But send the water in the opposite direction and the pipe shrinks (becoming more resistive). Further, the memristor remembers its diameter when water last went through. Turn off the flow and the diameter of the pipe “freezes” until the water is turned back on (Williams, 1).

That freezing is what’s important and why this is such a great idea.

This theory sounds like it would be revolutionary for our electronics industry so why is it not already implemented? Well, a main reason is that although Leon Chua’s theory predicted the existence of a memristor in 1971, it wasn’t until 2008 when one was actually made (Fig. 3). Leon Chua’s theory was harshly criticized as highly impractical since the effects were on such a small scale and might even be impossible, so no one was in a rush to try and prove a dead theory. “Few people had read it, fewer had understood it, and fewer still had cited it” (Williams, 1). It was not until 2008 when Hewlett-Packard Labs headed by Stanley Williams accidentally made one that this theory was brought back from the dead and finally respected.

Stanley says that he and his team were six years into their research before they discovered what they had built. A man by the name of Greg Snider brought Chua’s paper to his attention. When Stanley looked at it, he noticed that Chua’s graph was looking “suspiciously similar.” Apparently what had happened, was when they were designing a circuit, some of the elements in it like the titanium dioxide had combined with the oxygen and transformed into slight variations that changed what they had built into something that resembled what Chua predicted. “On 20 August 2006, I solved the two most important equations of my career—one equation detailing the relationship between current and voltage for this equivalent circuit, and another equation describing how the application of the voltage causes the vacancies to move—thereby writing down, for the first time, an equation for memristance in terms of the physical properties of a material” (Williams, 1).

William’s team worked day in and day out once they realized what they had and how it works to perfect it and actually make something with it. Within a month they had working devices and were able to manipulate the characteristics of the memristors. They measured the resistance of the devices with and without power (using a voltage so small it left the resistance unchanged) and the readings were consistent each time. They had achieved actual memristance. “We had coaxed Chua’s mythical memristor off the page and into being” (Williams, 1).

Now that memristors have been invented and are no longer just a theory, there are many exciting applications. The most practical of these applications will be used on the home computer. Imagine doing a large volume of work on the computer, and being able to then unplug it and plug it back in, and then have it turn on in the exact spot that it was left in. No boot-up, no loading, just right back where it started. It would also be a faster, harder and the new architecture of these computers would last significantly longer than the current life span of transistors. It is also a more volume conscious way to build computers since memristors are around the size of 4nm and can be stacked in a 3-D matrix, meaning that instead of having memristors laying out flat on a board they can also be stacked on-top of each other. Stanley Williams and Greg Snider showed that by replacing specific transistors with these types of
memristors, they could reduce the area by a factor of 10 and improve the circuits speed relative to power consumption performance.

Another door memristors open up, is the ability to create a circuit that can perform real-time analysis in a similar manner like that of the human brain. This task is very complex without memristors. Trying to create such a circuit with only transistors and the traditional 3 fundamental elements of a circuit would be the size of a small city and would require so much power that it is practically impossible. But building one with memristors is already being worked on by Greg Snider. “The neurons are implemented with transistors, the axons are the nanowires in the crossbar, and the synapses are the memristors at the cross points. A circuit like this could perform real-time data analysis for multiple sensors. Think about it: an intelligent physical infrastructure that could provide structural assessment monitoring for bridges. How much money—and how many lives—could be saved?” (Williams, 1).

This new memristor will revolutionize the world. Electric tools and heavy machinery could identify a hazardous situation and prevent the operator from a physical injury. Draw bridges could identify when it was appropriate to raise and lower the bridge without human involvement. Roadways could regulate traffic more efficiently if these circuits were integrated into intersection lights. There are numerous ways for this circuit to be imbedded into current technology to prevent physical injury. Medical robots are another industry that is ripe for opportunities to imbed this technology. Machines could control them and instead know exactly what to do and how to do it and how to deal with a situation if something were to go wrong. Self-driving cars and self-piloting airplanes could eliminate the need for professionals, eliminate accidents, and eliminate the need for air traffic control. Everything could be connected to a grid and communicate with each other and eliminate the need for human involvement. Android robots will do more than fall down stairs and try to run. They will be the servants, the servers. These robots will clean hotels, cook fast food and check people out of the retail store. This is the technology that movies have dreamed about, written about and inspired future scientists.

These new applications are inspiring and exciting however there is a lot of work that must be completed before we can move forward. The current design of the memristor is simplistic and only in the prototype phase. There is still so much more research and development that must be completed before this technology is ready to take over the world. It is hard to predict how long it will take to create these gains. Only time will tell.

This technology is great, but what about the current system already in place. Products would have to be redesigned around this radically new technology. The first thing that has to happen is to get memristors out there in the marketplace which is proving to be harder than it sounds. Hewlett-Packard planned on releasing electronics using circuits based of memristive architecture in 2013 but there have been large delays that have stopped this from happening. “It’s sad to say, but the science and technology are the easy part. Development costs at least 10 times as much as research, and commercialization costs 10 times as much as development. So in the end, research — which we think is the most important part — is only 1 percent of the effort” (Garling, 1).
The question then becomes, who will make the investment? Hewlett-Packard is not the only one exploring this new technology. Other companies like Samsung and IBM are keeping a close eye on this upcoming development to see how it could improve their products. Already a professor named Jennifer Rupp is working with IBM to build a memristor-based machine. One hopes that with competition to create the next best product companies will be inspired to invest in this new technology and be able to drive down the manufacturing costs quickly so that it will flood the market with new products and opportunities. The government may want to consider investing in this new technology for its military applications. The more demand for this new technology and the more players involved in the development of this technology will decrease the production costs and hopefully make it more affordable and mainstream in a short period of time.

Stanley said in his press release about his team’s discovery that, “I’m convinced that eventually the memristor will change circuit design in the 21st century as radically as the transistor changed it in the 20th. Don’t forget that the transistor was lounging around as a mainly academic curiosity for a decade until 1956, when a killer app—the hearing aid—brought it into the marketplace,” and right he was. Hewlett-Packard recently partnered with Hynix, a South Korean company, to help kick-start memristors by manufacturing circuits with them. Hewlett-Packard is also launching a program called, “The Machine” (Fig. 4). The Machine is not an object but rather a continuum. Its intent to rebuild and rethink computer architecture from the ground up. They are changing how a computer computes, how it communicates, and how it stores data. With data, they say they can use memristors to completely replace all the different forms of storage and combine the best parts of them, so instead of having a hard drive and RAM and some cache for the CPU, there will just be one storage chip that can do all of that and then some all thanks to the memristors. This new technology is future proof for the next few decades at least. In Hewlett-Packard’s conference “Discover”, Martin Fink talked about The Machine and said that you can access “a byte of information from one-hundred and sixty petabytes in under two-hundred and forty nanoseconds.” To put that into perspective, modern day hard drives on average are around the magnitude of one to two terabytes and are around one-thousand times slower. “The Machine is a hyper-dense collection of computing hardware that could be used in anything from a data center to a mobile device. It has terabytes of storage and a much smaller power draw than today’s computing devices—all because of memristor-based memory and optical interconnects” (Gallagher, 1).

Hewlett-Packard has done an amazing job by proving Leon Chua’s theory, even if it was unintentional. It may be one of the greatest scientific achievements of our time and The Machine could not exist without it. Stanley Williams and his team have worked tirelessly and dedicated their lives to making Chua’s theory into a reality. Chua should be proud to see his work come to fruition. After years of scrutiny and ridicule, it turned out his theory was right all along. He certainly was a head of his time. However, without Stanley Williams and his team, Chua’s theory might have never been proven. Together, they will be changing the face of technology as we know it.

When it comes to storing data and information in a computer, some believe the cloud to be the way forward, but I say memristive technology is without a doubt the future. It is faster, larger in capacity while at the same time smaller in physical size, and all around a better technology than anything else on the market right now. Being a nonvolatile form of storage, it is
also safer. It also provides for more security and may be useful to business and government alike in storing extremely sensitive and private material and keeping it safe from hackers by allowing the volume of information to be stored on site rather than on a cloud server. Nothing can compare to what they have to offer! This technology will revolutionize the world the same way the TV revolutionized the radio or how the cell phone revolutionized the telephone. It may be as big as putting a man on the moon. It may even be the technology that allows us to fly to Mars. I wonder if some time in the future I will have a robot cleaning my house and a car that drives itself. It is just going to take time to implement this new technology, but I know it will happen, and it will be well worth it. If there is one thing I can tell you for sure, it is that I will be on board with this new architecture and ready to help test this however I can. I believe this really is going to usher in a new era of computing, one with lower energy usage, faster data processing and transferring, and a new way to experience computers, and maybe life.
Figures

Fig. 1. A Nano scale picture of a memristor from Hewlett-Packard Labs.


Fig. 2. A diagram of a memristor from Hewlett-Packard Labs.


References


Modalities Used in Physical Therapy
Nicole Mallery
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General Physics II – PHY 112
Dr. Casey Durandet
ABSTRACT

The purpose of this research paper is to discuss the principles of physics in relation to the selection and application of various treatment methods utilized to rehabilitate individuals with bodily pain, injury or immobility. Physical therapists, occupational therapists, chiropractors and other such practitioners specialize in manual therapy, in which they use their hands and arms to manipulate joints and apply pressure to muscle tissue in order to alleviate pain; however, they are also trained to employ other assistive tools and devices called physical agent modalities.

Some of the most commonly applied physical agent modalities used in patient rehabilitation include: neuromuscular electrical stimulation, transcutaneous nerve stimulation, low-level laser therapy, and therapeutic ultrasound. Each of the listed treatments exists as a result of the marriage between the fields of physics and physiology, and the research provided delves deeper into the explanation of this relationship. More detail is discussed around the phenomena of each method – the principles of physics that it presents, its practical purposes and intended outcomes.

EARLY EXPLORATION OF ELECTROTHERAPY

For many years, researchers have been curious to observe how the human body reacts to induced stimuli, such as electricity, light and sound.

Luigi Galvani (1737-1798) was a renowned Italian physicist and anatomist whose work in the late 1700s provided clarification on a well-studied but mostly misunderstood marvel where frog legs contracted in response to electrical stimuli 1. Years later, English physicist and chemist Michael Faraday (1791-1876) invented Faradism, a type of treatment that involved his new invention: the transformer 1. With his transformer, Faraday was able to demonstrate neural movement in response to electrical current stimuli 2.

Discovered in the 1960s, lasers made their appearance in medical applications many years after Galvani and Faraday made their impact with electrical stimuli. Hungarian physician and professor Endre Mester (1903-1984), though not the first to discover the laser, is considered to be the pioneer of photobiomodulation, another name for low-level laser therapy, due to his experiments that proved laser energies as low as 1 Joule/cm² per day could stimulate hair growth and skin repair 3,4.

The application of ultrasound is widespread – bats utilize echolocation to navigate, ships use the hydrophone to detect underwater obstacles, and doctors and expectant parents rely on sonograms to view unborn children. Despite its usefulness across a number of fields, it was not until the 1930s that ultrasound made its way into the medical field. Within two decades, ultrasound had become widely adopted in physiotherapy due to its apparent ability to repair injured or destructed tissues and nerves including that caused by diseases such as Meniere’s and Parkinson’s 5.

Today, neuromuscular electrical stimulation, transcutaneous nerve stimulation, low-level laser therapy, therapeutic ultrasound and other such treatment methods are collectively called physical agent modalities (PAMS). PAMS are defined as “procedures and interventions that are systematically applied to modify specific client factors when neurological, musculoskeletal, or
skin conditions are present that may be limiting performance” ⁶. Assistance from electrical, thermal and mechanical energy allows practitioners to enhance their therapeutic impact in conjunction with manual manipulation so as to achieve more advanced physiological changes in their patients.

The following sections outline the physics behind the aforementioned physiotherapies as well as other pertinent details, including the functional purpose and desired impact of each. It is important to note that many more PAMS exist in the realm of patient rehabilitation than can be discussed within this research paper. Those that were selected, however, have the highest adoption rates among practitioners, and thus, are the most valuable PAMS of which to be aware. For information on other PAMS, additional reading and/or research should be conducted. The references listed are valuable, as they collectively provide in-depth histories, findings, explanations and resources for a number of PAMS that were either discussed or not discussed in this review.

**ELECTRICAL PHENOMENA**

As mentioned, neuromuscular electrical stimulation (NMES), transcutaneous nerve stimulation (TENS), low-level laser therapy (LLLT), and even therapeutic ultrasound are all called PAMS, but more specifically, they are electrotherapy modalities. It is necessary to first explore the fundamentals of electrical phenomena, so as to understand the behavior and effect of such practices.

All biological matter, humans included, is composed of atoms. When outside forces such as electricity, light or sound, alter the electrical state of an atom, it becomes an ion with an electrical charge measured in coulombs (C) ⁸. The electrons released by this process create a current flow, measured in amperes (A) or one coulomb per second, and tend to become densely concentrated in one place. These electrons will then flow from high to low concentration. The larger the difference is between these concentrations, the greater the electromotive force (EMF) or voltage (V), and the greater the potential for electron flow ⁸. Current flow may be opposed by resistance, measured in Ohms, in a circuit. In reference to the characteristics of a current, we must consider amplitude (the highest value of oscillation), frequency (the number of cycles per second), phase (level or synchronicity between two current waves) and impedance (resistance) ⁸. Charge, or the “electrical threshold for excitation of nerve and muscle” is the factor of stimulus amplitude and the duration ⁸. As displayed in the strength-duration curve in Figure 1, amplitude and duration are inversely proportionate in determining the threshold stimulus or charge needed to activate a muscle or nerve.

Despite the fact that these therapies are within the same PAMS subcategory, each is different than the next due to the various laws of physics that define them, how they are administered by practitioners to patients, the specific presentations and ailments they are intended to remedy, and even the level of consensus around their affectivity. One differentiator is the three unique current classifications of electrical stimulation, all of which are measured in Amperes (A): direct current (DC), alternating current (AC) and pulsatile current. A DC current, also called a Galvanic current, flows in one direction for one second or longer, and the direction of the flow is determined by the polarity (either negative or positive) selected on the machine by the
practitioner. An AC current, also called a sine wave or medium frequency waveform, has bidirectional flow; it initially flows in one direction for less than one second, then it reverses direction. Lastly, the most common is the pulsatile current, which flows for a few microseconds, then ceases to flow before starting again in either a unidirectional or bidirectional manner.

**SPECIFIC ELECTROTHERAPY MODALITIES**

**NMES & TENS**

Often, the topic of clinical electrotherapy is complicated by the terminology, as many methods are designated by minor differences and some therapies have multiple names. As an example, it can be difficult to differentiate between TENS and NMES, which is also referred to as transcutaneous muscle stimulants (TMS) or electrical stimulation (ES). When applied, both stimuli have the ability to arouse the peripheral motor nerves, which subsequently leads to contraction of the corresponding muscle(s). It must be noted that the muscle will contract in response to TENS only if the muscle is denervated, or without nerve supply; thus, TENS and NMES are only interchangeable terms under these specific conditions.

Another similarity between these two electrotherapies is that they are both pulsating currents that flow for a few microseconds before beginning to flow again, as was previously mentioned. As well, both NMES and TENS are administered through electrodes that are strategically placed onto the patient’s skin and whose surface must be adhesive. This ensures that the electrodes maintain their location as determined by their practitioner, which is important for proper treatment, and that they remain in close contact with the patient’s skin, which is crucial for providing a closed circuit through with the electrical current can flow. As evident in Figure 2, practitioners, or patients who have their own TENS units, may place the electrodes in a variety of patterns to ensure personal and optimal usage based on the particular ailment and/or the goals outlined in the patient’s treatment plan.

The primary difference between NMES and TMS is that NMES serves primarily to stimulate the muscles to contract, whereas the purpose of treatment that involves TENS is to encourage the nerves to contract. NMES promotes muscle strengthening in a similar manner that exercise does, and it has been used in sports medicine “for muscle strengthening, maintenance of muscle mass and strength during prolonged periods of immobilisation, selective muscle retraining, and the control of oedema.” In Figure 3, there is a visual of a subject amidst experimentation involving light exercise and NMSE stimulation. TENS, on the other hand, “is a method of electrical stimulation which primarily aims to provide a degree of symptomatic pain relief by exciting sensory nerves and thereby stimulating either the pain gate mechanism and/or the opioid system.”

**LLLT**

LLLT is also known as photobiomodulation, and because it is laser light, it will travel in a straight line and at a persistent speed. The probe used to administer laser treatment can provide an output that is continuous or it may be pulsed depending on the discretion of the practitioner.
and the needs of their patient. As seen in Figure 4, LLLT provides a great deal of benefits, including but not limited to: decreased inflammation or edema, nerve regeneration, and cartilage and collagen production.

ULTRASOUND

“Therapeutic ultrasound is one of the most widely and frequently used electrophysical agents,” and it has played a role in clinical practice for over 60 years; however, there is still debate about its effectiveness. By definition, ultrasound is the utilization of ultrasonic sound waves, which is represented by a sine wave or sinusoid that repeats its oscillation pattern in a smooth manner. Furthermore, ultrasonic waves are characterized by short wavelengths and high frequencies that, at 20kHz, are above the human auditory limit. In reference to the previous discussion about current classifications, ultrasonic waves are AC, thus it has a bidirectional flow with less than one second before reversing.

The mechanism to transfer ultrasound waves to a patient is called a transducer. The surface of the transducer that interacts with the skin is adorned with piezoelectric crystals. This means the crystals are made to vibrate by an electrical stimulation, which results in mechanical energy in the form of sound waves. A simple visual of a transducer adorned with electrical connections to results in the vibration of a piezoelectric crystal can be seen in Figure 5. In order to properly administer ultrasound therapy, the practitioner must maintain consistent and direct contact between the patient’s skin and the transducer. Ultrasound gel, cream or oil can serve as a “coupling media” and assist in minimizing surface friction so as to ensure successful transmittance of the ultrasonic waves to the tissue underneath the surface of the skin. The primary purpose of therapeutic ultrasound therapy is to relieve pain and stimulate regrowth or repair of tissue, with the greatest level of absorption occurring in tissues that boast high collagen content.

CONCLUSION

The prefix “phys-” means “nature” or “natural order,” and it is no wonder that this derivation is present in both physics and physiology. The research presented here indicates a clear relationship between physics and physiology, not only in the day-to-day movement and maintenance of the human body, but also in the alleviation of various pains, ailments and limitations. Even before the 1900, physicists, physicians, anatomists and other trailblazers were exploring the way in which external stimuli affect involuntary human movement. Their findings, in addition to later experimenters who expanded and further refined the previously discovered outcomes, have aided in providing healing therapies to people all around the world – from athletes to the elderly, from the able-bodied to the ailing.

Manual manipulation provided by specialists in various physiological fields has been enhanced with the use of PAMS, and more specifically, electrotherapy. With so many methods at their disposal, practitioners can rely on their expertise, experience and critical thinking to best diagnosis their patient’s needs in order to create a treatment plan best suited for each individual. Many studies have shown that NMES, TENS, LLLT and therapeutic ultrasound are influential in a number of beneficial outcomes, such as pain alleviation, tissue repair, blood circulation, and
more. It is important that practitioners be knowledgeable in the physics behind each electrotherapy, as both the administration and settings of the treatment are critical in not only injury remediation, but also in avoiding any unnecessary risks that could be brought about by not utilizing proper care.
**FIGURES**

**Figure 1:** As displayed in this strength-duration curve, amplitude and duration are inversely proportionate in determining the threshold stimulus or charge needed to activate a muscle or nerve.

[Image: https://classconnection.s3.amazonaws.com/432/flashcards/2223432/png/3-1431CB1BF9A0AD1EB08.png]

**Figure 2:** Practitioners, or patients who have their own TENS units, may place the electrodes in a variety of patterns to ensure personal and optimal usage based on the particular ailment and/or the goals outlined in the patient’s treatment plan.

**Figure 3:** The individual in the photo is the subject of an NMES conducted at the University of Florida. In the photo on the left, the subject is relaxed, and on the right, he is contracting the muscle of his left leg during NMSE stimulation.


**Figure 4:** This visual demonstrates the benefits of laser therapy, including but not limited to: decreased inflammation or edema, nerve regeneration, and cartilage and collagen production.

[CLINICAL EFFECTS OF LASER THERAPY](http://www.cruzchiropractic.com/wordpress/wp-content/uploads/clinical_effects.jpg)
**Figure 5:** This is a simple visual of an ultrasound transducer. The electrical connections on either side of the piezoelectric crystal will cause it to vibrate and give off ultrasonic waves. These waves serve as mechanical energy that will then engage the tissues and hopefully relieve pain or encourage repair.


**Figure 6:** Though protein and bone have high levels of protein, which typically indicates high absorbency of ultrasonic waves, their surface will likely reflect rather than take in the current. The visual indicates those tissues that will most successfully absorb ultrasonic waves.

REFERENCES


http://www.ncbi.nlm.nih.gov/pmc/articles/PMC3375668/pdf/yjbm_85_2_201.pdf


4 Gáspár L. Professor Endre Mester, the father of photobiomodulation. J Laser Dent [Internet]. 2009 [cited 20 Apr 2016]; 17(3):146-148. Available from: 


http://ajot.aota.org/article.aspx?articleid=1867124&resultClick=1


10 Watson T. Main Electrotherapy Modalities [Internet]. International Society for Electrophysical Agents in Physical Therapy (ISEAPT). 2015 Jun [cited 20 Apr 2016]; Available from:  
http://www.electrotherapy.org/modality/main-electrotherapy-modalities

http://ptjournal.apta.org/content/81/7/1339
How Efficient is Your Kitchen?

By: Connor Maloney
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Physics 112
Dr. Durandet
Abstract

Have you ever been in the kitchen preparing dinner and thought to yourself how does cooking food really work? What are the physics behind cooking food? The answer is heat transfer in the form of energy or thermodynamics. There are three main types of heat transfer that will be focused on: Conduction, Convection, and Radiation. There are various appliances in the kitchen that utilize these different types of heat transfer. Placing a skillet over a hot stove is an example of conduction, where the heat is directly transferred between the two objects. An example of convection can be seen when placing a pot of water on a hot stove. As the water heats up and gets hotter it rises forcing the colder water to sink, which is creating the movement of fluid producing convection currents. The third type of heat transfer is radiation, most commonly seen in the microwave. This type of heat transfer includes absorbing or giving off electromagnetic waves, which does not require direct contact between the two objects. This paper will touch upon which ways of heat transfer are most efficient in terms of cooking time and energy cost.

When discussing the three basic ways heat is transferred among objects it is important to understand how Thermodynamics work. Thermodynamics can be understood as the study of systems containing energy in the form of heat or work. There are three laws of thermodynamics. The first law of thermodynamics demonstrates that energy can exist in many different forms and can be transferred to other objects. Energy is always conserved and can not be created or destroyed (Farabee, 2006). Energy can not just be created out of thin air, it has to be transferred from another system. The Second law of thermodynamics tells us the available energy after a chemical reaction will always be less than that at the beginning of the reaction. Energy conservations are never 100% efficient, This is commonly referred to as entropy. Although energy can not be created nor destroyed it can be lost during reactions in other forms. When heat is transferred between two objects there will always be some energy lost. The third law of thermodynamics defines that the entropy of a perfect crystal is zero, when the temperature of that crystal approaches or is equal to the absolute value zero (Farabee, 2006). When examining the Three Laws of Thermodynamics it is seen that the first and second laws are most applicable to heat transfer in terms of conduction, convection and radiation.

The first type of heat transfer that will be focused on is conduction. When discussing conduction it is important to understand how the process works. Conduction is the transfer of thermal energy among molecules that are in direct contact with each other or between two surfaces that are touching. It can be seen that heat is transferred along temperature and energy gradients. When conduction occurs the material itself does not flow, rather, it is the transfer of heat internally. The heat transfer can be seen traveling along the molecules that are vibrating in place. The molecules or particles that are vibrating have high levels of energy. They will pass part of that energy off as they continually bump into each other. These molecules with more movement or higher energy will collide into the less energized molecules, and as each molecule touches another, thermal energy will be transferred.

The formula for conduction is \[ Q/t = \frac{K*A*(T2-T1)}{d} \] (Wood, 2013)
The rate at which conduction occurs depends on several factors, such as: what is the object made of? How much surface area do the two objects in contact have? What are the temperatures, and the thickness? From the formula we see that:

- \( \frac{Q}{t} \) is the rate of heat transfer, the unit measurement for this is joules per second or Watts.
- \( K \) is the thermal conductivity of any material. All materials have a specific \( K \) which is unique.
- \( T_2 \) and \( T_1 \) are the temperatures of the two objects in contact.
- \( \frac{1}{d} \) is divided by \( d \), which is the thickness of the material.

So the formula reads as the rate of heat transfer of any object is equal to the thermal conductivity multiplied by the surface area, then multiplied again by the difference in temperatures, all divided over the thickness.

Conduction can be conceptualized by imagining dominoes set up in straight line. When the first domino is pushed over, it hits the second dominoe and then the third and fourth and so on. Heat transfer in the form of conduction acts like the dominoes. Each dominoe has thermal energy and as it falls over and knocks the next one down it passes along that energy. Imagine placing a metal rod into a burning fire. As the molecules on the end of the rod in the fire get hot they begin to bump into each other or vibrate. Each molecule will pass its heat off to the next until all the molecules are hot. As a result the entire metal rod will become extremely hot. Conduction can also be seen if someone were to place their hand on that same metal rod, after it was completely hot, the heat energy would transfer into their hand and burn them. This is also conduction because there is direct contact between two surfaces. (Figure 1).

Materials that transfer heat efficiently are called conductors and materials that do not transfer heat efficiently are called insulators. If a material is a conductor it will allow electrons to move freely from particle to particle. In contrast, insulators will impede the free flow of electrons among the molecules and atoms. For example, metal is a great material for conduction while glass and plastic are not and would be considered insulators. This is why pots and pans are made out of metals such as iron and copper because they need to be able to transfer the thermal energy produced by the stove top too the food. We can also see examples of insulators in the kitchen, such as plastic or glass cups which are great for insulating ice water.

The primary appliance in the kitchen that utilizes conduction cooking is the electric stove top. These stove tops use large metal coils that convert electricity to heat. Then the heat is transferred to the pan and into the food through conduction. Unlike electrical stove tops induction cookers transfer electrical energy through induction into a material which needs to be electrically conductive and ferromagnetic. This process takes place by heating the metal through electromagnetic induction. Electromagnetic induction occurs when a circuit that has an alternating current (AC) flowing through produces current in other objects with magnetic fields when placed close by. This is because every electrical current has a magnetic field surrounding it, and those alternating currents have fluctuating magnetic fields, referenced in (Vinetarie, 2014), (Figure 4). This will cause currents to flow in conductors due to fluctuating magnetic fields, which is referred to as Faraday’s Law. When a pot which has its own magnetic field is placed near an alternating current with its own magnetic field, it will produce what is called an eddy current. Eddy currents lead to resistance and produce heat in the conductor or metal pan. The pan itself is the object producing the heat. Although induction does not require conduction to transfer the heat between the stove and pan, conduction is used to transfer the heat produced by the pan to the food. This type of cooking is safer for all ages because the stove top remains
cool while the pan is the only object heating up. When looking at efficiency levels in terms of energy loss from heat transfer, there is less thermal energy lost with induction cooking than there is with electrical stoves. When cooking with electrical stoves conduction is required twice, in that energy is transferred from the coils into the pan and from the pan into the food. As mentioned earlier from the second law of thermodynamics energy conservations are never 100% efficient. This means that since electrical ovens require conduction to take place twice, there will be more energy loss, while induction cookers only transfer heat once through conduction. That is from the pan to the food directly. Induction cooking is easier to clean which save stime and saving time means being efficient. Based on what we have learned from thermodynamics induction cookers are more efficient and are the future in stove top cooking.

The transfer of heat through convection according to (Nave, 2012) can be defined as the bulk movement of particles caused within a fluid by the tendency of hotter and therefore less dense material to rise, and colder, denser material to sink under the influence of gravity, which consequently results in transfer of heat. Liquids and gasses have the makeup of loose random particles that move freely from location to location. These particles carry thermal energy. The particles with higher thermal energy move more quickly than those with less thermal energy. This flow of particles will cause a current within the contained system, and this is referred to as convection currents. As air heats up it expands and becomes less dense, therefore rising, which then gives way for colder more dense air to drop and replace the warmer air. The moving fluid carries energy and this process creates a flow and allows the thermal energy to be transferred within the fluid.

The formula for thermal convection is:
\[ q = (h)(A)(Ts-Ta) \] (Bramble, 2009)

- \( q \) is equal to the heat transferred per unit time, which is in Joules per sec or Watts.
- \( h \) in units of (W/(m^2 K)) stands for the convective heat transfer coefficient of the reaction taking place. This coefficient will depend on the physical properties of the fluid.
- \( A \) in units of (m^2) is the area of the surface in which the heat transfer takes place.
- \( (Ts-Ta) \) in units of (F) is the temperature of the surface minus the temperature of the air. There are many factors that can affect the rate of heat transfer, fluid velocity, fluid viscosity, heat flux, surface texture, and type of flow (Bramble, 2009).

Convection is the primary method of heat transfer in fluids such as liquids and gasses according to (Nave, 2012). For example, have you ever wonder how a small column radiator in the corner of the room does such a great job of heating up the entire room? It utilizes the basic principle of transferring thermal energy through convection. As the radiator produces heat it warms the immediate surrounding air. As this air heats up it expands, which causes it to become less dense and allows it to rise to the top of the room. As the warm air rises it pushes the colder air from the top towards the bottom and the colder more dense air replaces the warmer air at the bottom. The colder less dense air begins to heat up as it approaches the radiator and begins to rise as this process continues it creates an air flow in the room. The air travels along these currents carrying thermal energy from the radiator. Convection can also be seen at work through fluid water. For example, when placing a pot of water on a hot stove the water is heated up by convection, but at first the heat is transferred to the pot through conduction. As the pot becomes
hot it conducts thermal energy to the water. Once the water near the perimeter of the pot becomes warm it expands and becomes less dense. This causes the water to rise towards the top. Once this occurs the colder water begins to sink towards the bottom because it is more dense. The differences in the water density from the top to the bottom begin to form a steady current of hot and cold water. This current flow provides a way for the heated water to transfer the thermal energy throughout the pot. Eventually the entire pot of water will be hot and ready for cooking.

There are two primary methods of convection that can be applied when transferring heat. The first method of convection is called natural convection. This is caused by a natural difference of densities in the fluid, caused by temperature differences that result in convection currents. As the fluid is heated up at the source location, it will rise or move away to an opposing location allowing the colder fluid to take its place. This is a natural process that will steadily increase as the fluid heats up. This form of convection can be seen in nature, when the earth’s oceans and atmosphere use natural convection to heat up the earth’s surface.

The second method of convection is called forced convection. This method implements the use of fans or pumps that push the heated fluid from one location to another. While the air is heated up at the source a fan blows the hot air to a location of cooler temperature giving way for the colder air to move towards the heat source. This will create the convection currents artificially. Houses generally use this method of convection. The furnace heats up the air and then blows it through the ductwork which is then released through the vents into the colder rooms. These two methods may pose the question, which one is more efficient? The mode of forced convection has been shown to be faster for heat transfer due to an increase in flow velocity according to (Malhorta, 2014). For example, if there is an increase in the flow velocity it will create a faster flow of current, as this will take the heated up air or fluid and transport it more quickly to the cooler locations. This will then increase the time it takes to create convection currents and therefore transfers heat more efficiently. In our earlier example, it would take much less time to heat up a house using this method.

When standing outside on a windy day it will feel colder when the wind blows on you versus when the wind is not blowing even though the temperature remains the same. The same concept applies with convection. The thermal energy will be transferred to your skin more quickly than if there were no wind. The wind in this case can be considered an external fan. That is why when looking up weather conditions it is always important to also look at the wind chill, because if there is wind it will be a lot colder than what the temperature really is.

What makes a convection oven more efficient than a traditional oven? The answer is forced convection. As previously stated, forced convection is the movement of air or fluids using an external fan or pump to create continuous convection currents. A convection oven will circulate hot air through the oven cavity using an external fan. When hot air begins to blow on the food versus surrounding it, it will transfer the thermal energy faster, which speeds up cooking time. The heat transferred by convection is performed through the bulk movement of air or fluids of different temperatures which result in varying densities to different locations. The moving air will speed up the rate at which the thermal energy is transferred into the food. This will provide an equal temperature throughout the oven cavity and will prevent the food from being cooked unevenly because there is a continuous flow of air around the food, versus when air of two varying temperatures collide, which is natural convection. The conventional oven only has two heating elements which come from the top and bottom. The top and the bottom both radiate heat towards the middle of the oven, this uneven distribution of heat can sometimes lead to hot and cold spots within the oven. Another disadvantage to using a conventional oven is
the limitation of the amount of food that can be put in at once. Because the heat is being radiated from the top and bottom, the middle rack will not have access to the heat. In convectional ovens this is not the case because there is a flow of hot air through out the entire oven cavity, allowing more to be cooked at once. When cooking with a convection oven the time it takes to cook a meal is shorter which correlates to less time using electricity which saves money. This type of cooking is the most efficient way to prepare meals.

Heat radiation is a form of energy transfer by the release of electromagnetic waves which carry energy away from the emitting object, in part because of temperature differences, and the speed at which this energy is transferred occurs at the speed of light, according to (Nave, 2012). The word radiate itself means to send out or spread something from an original source. Thermal energy that is transferred by electromagnetic waves does not require a medium or movement of fluids such as convection or direct contact of two objects as in conduction. Thermal radiation which is also referred to as infrared radiation is a type of electromagnetic radiation and is a method of heat transfer through space or a vacuum. All objects emit and absorb electromagnetic radiation. The energy of electromagnetic radiation varies as the wavelength and/or the color of an object changes. The smaller the wavelength an object has corresponds to a higher frequency which means it has higher energy and is generally a brighter color. The formula that shows this relationship is: \[ E = \frac{h \cdot \nu}{\lambda} \] (Oswieler, 2015).

- E represents the energy
- \( \nu \) corresponds with the frequency
- \( \lambda \) is the wavelength
- \( h \) is a constant

The temperature of the emitting object also affects the wavelength and frequency and the rate at which the radiated waves occur. As the temperature of an object increases, the wavelengths of the emitted radiation decrease according to (Oswieler, 2015). This simply means the hotter the object is the wavelengths tend to be shorter, which relates to higher frequency radiation and brighter color. An example of this would be with a toaster. As the coils heat up they become bright red. The wave lengths that the toaster is emitting are short which means they have a high frequency and that relates to a higher temperature and color because we are able to see the red coils and that tells us they are hot.

The formula that demonstrates how to find the heat transferred by radiation is called the Stefan-Boltzmann law of radiation and is shown as:

\[ q = e \cdot \sigma \cdot A \cdot \left( \frac{\Delta T}{4} \right)^4 \] (Oswieler, 2015).

- The symbol \( \sigma \) stands for the Stefan-Boltzmann constant which is \( 5.67 \times 10^{-8} \) J s\(^{-1}\) m\(^{-2}\) K\(^{-4}\).
- A represents the surface area of the object,
- K stands for the absolute value of temperature in Kelvin, and the rate at which radiation occurs is directly proportional to the absolute value of the temperature.
- q stands for the heat transferred.
- e represents the emissivity of the object. The emissivity is the measure of how well or how poor an object radiates. For example, when \( e = 1 \) the object is an ideal radiator when \( e = 0 \) the object is considered a reflector.

As mentioned earlier all objects in the universe absorb and emit electromagnetic radiation even though it may not be visible to the naked eye. The rate at which objects transfer heat through radiation is based on the color of the object or in other words its emissivity value (e value). The color black is considered to be the most effective at transferring heat through
radiation and has an e value of 1, while in contrast the color white or gray is the least effective, having an e value of 0. Black objects will absorb large amounts of the incident radiant energy and reflect very little. Once this object retains the radiation energy, it will strongly begin to radiate the energy. These objects are considered to be good absorbers as well as good radiators (Figure 3). The grey or white objects will absorb minimal amounts of the incident radiant energy and reflect the majority of it. This means that less of the energy is retained and therefore less of the energy will be emitted. These objects are considered poor absorbers and poor radiators (figure 3). An example of this can be seen when sitting outside on a warm day in Phoenix Arizona. If a person decides to eat ice cream outside on a hot day, sitting in a chair on the black top of the parking lot, his ice cream will most likely melt more quickly because the black parking lot is emitting more radiation and will be considerably hotter. Data shows that the color black is an ideal absorber and an ideal radiator (Oswieler, 2015). If that same person was smart and wanted to enjoy his ice cream a while longer he should sit in a chair on the grey sidewalk. The color grey is a poor absorber and a poor radiator and will in turn radiate less energy. This scenario will be considerably cooler than the parking lot.

The primary appliance in the kitchen that utilizes radiation for cooking is the microwave. The microwave uses electromagnetic radiation to cook food without using a medium like conduction in stoves and convection in ovens. Microwaves use the same frequencies as radios and fall into the radio frequency band of electromagnetic radiation. For example, this means they are not as strong as x-rays or gamma rays, which can be dangerous at high levels. The microwaves themselves are produced by an electron tube called a magnetron which is inside the microwave. The magnetron produces these microwaves and as they reflect off the metal inside they get absorbed by the food. Once these microwaves are absorbed in the food they begin to cause the water molecules to vibrate which produces heat that cooks the food, according to (FDA, 2014). Although the heat is produced from within the food, the food does not cook “inside out”. Microwaves use conduction, which transfer the heat from the outside layers to the middle inside layers. That is why it is important to let the food cook half way through, taking it out to stir and then placing it back into the microwave to finish cooking. Microwaves are more efficient in the way of saving electricity and time spent cooking. They do not need to heat up the entire compartment because the heat is produced within the food. This cuts down on money spent on electricity and time in the kitchen. It is not safe to put aluminum or metal in the microwave because the microwaves do not pass through those materials and will be reflected back and can cause damage to the microwave, again according to (FDA, 2014). It is ok to put glass, paper, and/or plastic containers in because the microwaves can safely pass through those materials.

To summarize each concept, while thinking about the physics behind cooking, I will tell you that between conduction, convection, and radiation, that cooking using a convection oven would probably produce the best results, however, since I am a bachelor and a starving student, using radiation in the form of a microwave for my cooking is probably more likely. I like the speed of the magnetron producing short microwaves as they bounce around the metal inside and cook my food quickly. One downside is that I have to periodically stir the food during the total cooking time to help disburse the heat because the microwaves don’t heat the food as evenly as the convection oven would.

If I am using an electric stovetop to heat up my macaroni and cheese, I will have to wait for the conduction process to happen twice. First the electric coils will transfer the heat energy to the metal pan and then the heat passes from the pan to my dinner. According to the second
law of thermodynamics this is not as energy efficient as finding a more effective way (induction) to have that thermal energy passed directly to my food. I don’t have to touch the hot stove twice to learn this concept!

Forced convection cooking is more effective for cooking our Thanksgiving turkey. By creating a faster flow of current transporting the heat more quickly to the colder locations, this will result in heating the bird more evenly and efficiently. It could also cut down on my electric bill. By having forced convection in the oven, I can throw the traditional green bean casserole on another rack and the flow of hot air throughout the entire oven cavity will allow both the turkey and the casserole to be cooked at the same time.

I think convection and microwave cooking will continue to be more efficient than conduction cooking. Many ovens today have both options of traditional conduction or more effective and efficient convection baking and roasting. Convection cooks more evenly and uses less energy. For our fast paced lives, radiation cooking through microwaves will continue to have a place in our society. This form uses the least amount of energy.
Figures

Figure 1

This image shows how a metal rod that is placed in direct contact with a flame will perform heat transfer through conduction. It demonstrates that the molecules within the rod are “bumping” into each other and creating a flow of thermal energy down the rod.

Figure 2

This figure shows the heating of a room using a radiator which utilizes heat transfer through convection. As the radiator heats up the surrounding air it becomes less dense and rises which gives way for the colder more dense air to sink. This creates convection currents which carry the heat energy throughout the room.
In this figure we see that black objects are good absorbers and good radiators because they absorb the majority of the incoming radiant energy and strongly emit as they retain little of that energy. The silver or grey objects are poor absorbers and poor reflectors, they reflect the majority of the incident radiant energy and retain the little energy they do absorb because they are poor radiators.

This Figure shows how heat is produced by Eddy currents using induction heating. When there is a magnetic field and an alternating current surrounding it, it will produce heat in the form of E
References


Figure 1
https://www.google.com/search?q=radiation+frequency+chart&espv=2&biw=1280&bih=616&tbm=isch&imgil=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523Bhttps%25252F%25252Fwww.ekh.washington.edu%25252Frsotrain%25252Fsealedsources%25252Fraditationproperties.shtm&source=iu&pf=m&fir=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523B&usg=__UMzXTiizgnmQ7q4Tc-Jn_4KH6zA%3D&ved=0CCYQyjdQFQoTCPCt3O6xmMkCFQ8xiAodHYsLXw&ei=eZpLVrCNGY_ioASdlq74BQ#tbm=isch&q=conduction+heat+transfer&imgdii=emkzQkOo1pPdEM%3A%3BemkzQkOo1pPdEM%3A%3Bk1MUtzCZXj3jpBM%3A&imgrc=emkzQkOo1pPdEM%3A

Figure 2
https://www.google.com/search?q=radiation+frequency+chart&espv=2&biw=1280&bih=616&tbm=isch&imgil=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523Bhttps%25252F%25252Fwww.ekh.washington.edu%25252Frsotrain%25252Fsealedsources%25252Fraditationproperties.shtm&source=iu&pf=m&fir=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523B&usg=__UMzXTiizgnmQ7q4Tc-Jn_4KH6zA%3D&ved=0CCYQyjdQFQoTCPCt3O6xmMkCFQ8xiAodHYsLXw&ei=eZpLVrCNGY_ioASdlq74BQ#tbm=isch&q=covection+heat+transfer&imgdii=emkzQkOo1pPdEM%3A%3BemkzQkOo1pPdEM%3A%3Bk1MUtzCZXj3jpBM%3A&imgrc=emkzQkOo1pPdEM%3A

Figure 3
http://cnx.org/contents/031da8d3-b525-429c-80cf-6c8ed997733a@8.13:107/Radiation

Figure 4
https://www.google.com/search?q=radiation+frequency+chart&espv=2&biw=1280&bih=616&tbm=isch&imgil=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523Bhttps%25252F%25252Fwww.ekh.washington.edu%25252Frsotrain%25252Fsealedsources%25252Fraditationproperties.shtm&source=iu&pf=m&fir=AlktEjS9mFaAcM%2523A%2523BB5gvVfqXUC-DdM%2523B&usg=__UMzXTiizgnmQ7q4Tc-Jn_4KH6zA%3D&ved=0CCYQyjdQFQoTCPCt3O6xmMkCFQ8xiAodHYsLXw&ei=eZpLVrCNGY_ioASdlq74BQ#tbm=isch&q=eddy+currents&imgdii=rB7XVPlprwssLM%3A%3Br7XVPlprwssLM%3A&imgrc=rB7XVPlprwssLM%3A
Nuclear Energy
Karen Martinez
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Dr. Durandet
Nuclear energy is a newfound source of energy that has proved to bring benefits as well as advantages. The production of nuclear energy is a process that involves uranium and reactors. Uranium is a natural product that is molded into pellets through machines (1). Uranium goes through a conversion process to become uranium-235 in order for it to fission better and work with the neutrons that it will be exposed to. The reactors take a huge role in the conversion of nuclear fuel to power. Although there are multiple reactors every reactor uses uranium and plutonium as their main ingredients to produce energy. Just like any form of energy providing process it will always creates waste. Though nuclear energy can be less costly to manufacture the disposal of it can be tricky and dangerous due to the radioactivity it contains (2). There are benefits of using nuclear energy as well as downs falls but this type of energy may lead us to a better and cleaner type of energy.

As long as one can remember the use of energy was needed. Energy allowed for new breaks through and allowed the 21st century to have the advanced technology they have now. Nuclear energy is one of the controversial ways of producing energy using natural products and the waste of people. It has been the cleanest process in creating energy compared to green houses and coal burning (3). This type of energy producing process generates 7 billion kilowatt hours of electricity a year (4) Most are skeptical of nuclear energy because it will automatically be relate it to radioactivity, which affects many lands fills. In the recent years many discoveries have been made for the use of nuclear energy and power. We will look closely at the process of conversion from natural products until nuclear waste. We will discuss how the natural product is contained on transformed and the process it has in the reactor that converts the fuel to energy. Also the different types reactors there are available and some pros and cons that nuclear energy has. First we will beginner with the most important natural product.

Uranium plays a crucial role in nuclear energy. This metal can be found in the earths crust such as rocks, soil, rivers and sea water (4). Uranium is a natural product that can be extracted through mining, which is done through opening pits on the surface and underground mining that can be 120m deep with the help of in situ recovery (4). The less invasive type of mining is underground mining due less material being removed in order to get the ore needed (4). The disadvantage of this method is it can be dangerous (4). The newest method of mining can be seen through the use of situ leach (5). This method doesn’t remove the ore it recovers the minerals through dissolving the ore and pumping its solution to the surface (5). Oxygenated groundwater is circulated through the orebody in order to recover the minerals by leaching (5). In the United States in situ mines use alkali leach due to the significant amount of acid consuming minerals that can be found in the solution (5). Mining and conversion of uranium is a crucial step into the process of nuclear fuel because uranium give us Uranium fuel or also know as uranium oxide (UO2) which is loaded onto the reactors and allows for the process of conversion to begin (4). A better visual of the conversion process can be seen in Figure 1 through the nuclear fuel cycle (4). Only a small percent of natural uranium is capable of undergoing fission and in order to become fissible it has to be uranium-235 isotope (4). In order to get uranium-235 the natural source of uranium must go through a process. The conversion of uranium to uranium-235 starts with Uranium oxide going through a conversion in order to get the final product of fuel ceramic pellets, which are used for the nuclear energy (1). Uranium oxide is processed and converted into uranium hexaflouride through the process of heating so it may become gas and ultimately drained into 14 ton cylinders where it is left to cool and solidify (1&4). Fuel fabricators convert the solidified uranium hexafluoride into uranium dioxide power and it is pressed into ceramic fuel pellets, these pellets are then loaded into tubes made of zirconium alloy (5). These single
fuel pellets can produce energy as much as 17,000 cubic feet in natural gas, 5,000 pounds of wood, 149 gallons of oil, and 1,780 pounds of coal (5). A reactor that has an output of 1000 megawatts will have 75 tons of low enriched uranium (4). The heat in the reactors is used to produce steam that drives a turbine and electric generator, which generates 7 billion kilowatt hours of electricity a year and 44 million kilowatt hours of electricity, is produced through one toe of natural uranium (4). There are a couple things one must know before getting into the process of the conversion of energy.

In order to understand this process we must know that everything is made up of atoms that are charged whether it has a positive electrical charge know as a proton, negative electrical charge know as an electron, or a neutral charge know as an electron. In nuclear energy protons and neutrons are important, they are important because they allow nuclear energy to be created by combining individual nucleons know as fusion or by splitting into large combination known as fission (6). This is possible because both neutrons and protons whether put together or individually, are massive (6). Through fusion or fission a small mass is released and through the Einstein equation E=mc^2 it can be seen that the mass can be an amount of energy (7). Nuclear fission can be one of the most important steps in the process of energy due to the physics part of it. The energy mainly comes from the fission of uranium-235 because when it fission the fragments have a less mass than the uranium nucleus (7). The binding energy curvy can be seen in figure 2 (7). This curve gives a better understanding that if the mass of the fragments are equal or greater than iron at the peak of the curve then the nuclear particle will be tightly bound therefore it decreases the mass (7). There are different types of fissions that may occur depending on the number of neutrons (8). Even number neutrons fission will occur if incident neutrons have energy above 1 million electron volts (8). In odd number neutrons fission becomes larger in thermal energies where there are slow neutrons (8). Through a neutron cross section that is seen in figure 3, it is seen the fission increase as the neutrons velocity decrease form 20,000km/s to 2km/s making the interaction greater. Neutrons its self have thermal energy that occur when it slows down to thermal equilibrium according to its surrounding (8). This is why it is used in thermal reactor along with water to slow down the neutron (8). Around 85% of the energy obtained is initially kinetic energy of the fragments (8). The reactor, also known as chain reacting system, is in balance when the neutrons produced in fission are constant (8). The nuclear chain reaction occurs when two large atomic nucleus absorb neutrons and fission therefore allowing the heavy nucleus to split which then releases kinetic energy free neutrons, and gamma radiation (10). Neutrons play a crucial role in the balance of the chain reacting system because in order to raise or lower the power the neutrons need to me decreased or increased (8). The way to increase of decrease neutron can occur by using neutron poison or moderators (10). Now that we know the physics part of nuclear energy we can begin to look at nuclear waste.

Nuclear fuel is where we get our nuclear waste. Nuclear waste is obtained when a collection of nuclides are left over after the reactor has extracted the energy out of the nuclear fuel (2). The way we get our nuclear waste, also know as radioactive waste, is produced through the fission and conversion of nuclear fuel (web 1). Nuclides are made of neutrons that that do not fission but turn into isotopes (2). These nuclides may absorb neutrons and become heavy, they are referred to as transuranic (2). If nuclear fuel is used in a reactor, reactors are found in nuclear power plant, it is converted into nuclear waste and then it becomes dangerously radioactive (2). The process of converting nuclear fuel to waste depends on what is put into a reactor and how all the charges interact (2). In figure 4 we are able to see what type of waste is put into the reactors
Once all the energy is taken out and the fuel is no longer able to be used it goes through to one of two places, it can go through a reprocessing to be recycles or to a long-term storage to be disposed of. The choice of what to do all depends on policies a country may have, the waste that is left over, and its contents. The process of reprocessing and disposal can be a dangerous one.

The process of disposing nuclear fuel can be a dangerous process due to the high amount of radiation it contains, which is why disposal of nuclear waste is done with the help of water. Water acts as a shield and also cools down the reactor, once underwater the fuel will be removed and transferred into storage pools where they stay there up to 5 or more years depending on the radiation level. Once it’s been in the storage pools for 5 year it may be transferred into ventilated concrete container. The waste is left in the pools until the radioactivity begins to decay, a better understanding of the radioactivity decay can be seen in figure 5. The nuclear waste collected has different levels of classification, these levels will allow for the proper methods of storage to be chosen. High-level waste makes up about 3% of the total waste that comes from the generators but it contains 95% radioactive that comes from nuclear power. The most abundant type of waste in the total volume is low-level waste. This type of waste may consist of things such as tools clothes, and other items that are lightly contaminated. Low-level waste makes up about 90% in the total but it only has 1% radioactivity. Intermediate level waste will make up 7% of the total volume of waste but is only 4% radioactive. Low level nuclear waste is not very harmful to people, the type of radiation that one gets from this type of waste can be received through hospitals, clothing, filter, and tools. Low-level waste can sometime be disposed of as soon as it comes out of the radiator. Lower level waste doesn’t require further treatment this waste is similar to what can be found in municipal waste sites. The amount of radiation released in this level is shortly live because it makes up 1% or radioactivity. Intermediate level waste has a radioactivity of 4%, which make it important to be shielded. Materials that are in the intermediate level waste can be things such as chemical sludge’s, medal fuel cladding, and contaminated materials. Intermediate and low level waste can be disposed of close to the surface while high level waste is disposed in deep underground geological repositories. These methods are methods used to store the waste but it may also be recycled. Recycled or reprocessing fuel is a great benefit because it ensures that all the fuel is being used and nothing is being wasted. The reason to recycle is also because there is not set way to dispose of waste as well as limited storage sites.

The process of reprocessing starts with the separation of uranium and plutonium from the waste product and it is done by chopping the fuel rods and dissolving them in acid in order to separate the material. Uranium that comes from reprocessing contains higher concentration of Uranium 235 than natural uranium that was recovered in the begging of the nuclear cycle. About 95% of uranium is still in the used fuel, 1% of Uranium 235, 3% of waste, and 1% of plutonium. Plutonium is used to make mixed oxide fuel, which replace uranium oxide fuel that is added into the reactor during the first run of the nuclear cycle. Mixed oxide is a combination or uranium and plutonium oxides mixed together. There is much controversy that recycling doesn’t completely extract the waste from the chemical in essences leaving waste and have some sort of contamination. Plutonium and uranium reduction extraction allows for a better and pure form of extraction between substances. Reduction extraction has three cycles of purification. The first cycle tributylphosphate is dilute in a mixture of aliphatic hydrocarbon allowing en extraction of uranium and plutonium to occur from a 3-4M HNO3 solution. The extraction and chemical process can be seen on figure 7.
to the cycle is the partition of plutonium from uranium and it is done by reducing plutonium because tributylphosphate can not be extracted when it is in a higher oxidative state (12). The third step strips uranium from the tributylphosphate phase by dilute nitric acid (12). Hydroxylamine is used in the second and cycle to remove complete traces of uranium in order to get to step three where uranium and plutonium is completely removed (12). Though these three steps will allow for a better extraction of uranium and plutonium from waste we should begin to see how nuclear energy could benefit public.

There are multiple types or reactors that can be used figure 7 shows the types of reactors (10). The most common reactor that is used is the pressurized water reactor, which uses water as a coolant and moderator (10). A fuel cladding heats up the water of the coolant loop allowing for the water to heat up and pumped into the steam generator (10). The steam that was created in the steam generator is used for power generation (10). Coolant water is a moderator that lets neutron undergo collision with hydrogen and lose speed, when the water is denser there will be more collision occurring (10). This type of reactor uses about 150-250 ceramic rods that were previously mentions (10). The down fall of this reactor is that it requires water to be high pressurized in order to remain liquid at high temperature (10). A dilemma arises because the power plants run into the problem if making sure they have high strength piping in order to withstand the pressure and not have a leak, which can eventually lead to radioactivity getting out (10). Another reactor that can be used is a boiling water reactor, which uses water and steam (10). The water that is around the fuel element thermalizes the neutrons in order to reduce their kinetic energy allowing it to improve fission of the fissile fuel (10). With a boiling water reactor the steam occurs in the core instead of the steam generators, the steam then powers the electrical generator (10). In order to vary the thermal power the forced recirculation flow must be decrease or increase (10). This reactor always has to be shielded due to the water around the core always having contamination of radioactive nuclides (10). The reason why its radioactive is due to the cores continuous production of heat from radioactive decay after fission has stopped (10). Pressurized heavy water reactor is a reactor that has coolant under high pressure in order to raise the boiling point and avoid steam formation in the core (10). The difference between this reactor and the previously mention is that light water is not used, instead D2O is used because it absorbs fewer neutrons than light water would (10.) Moderators that have low temperature will reduce the likelihood of any change in the neutrons speed of collision (10). If the neutrons are at optimum speed in order to cause fission they have spectral purity and have a good range of neutron energies (10). Graphite moderated reactor which are also know as gas cooled reactor and advanced gas cooled reactors use carbon dioxides their coolant which carries heat to turbine and graphite moderator (10). Graphite moderator allow for natural uranium or slightly enriched uranium as fuel (10). Water-cooled light water phosphate moderated reactor is developed from plutonium production reactors (10). This reactor allow for excessive boiling to occur in order to decrease the cooling and absorption of neutrons without inhibiting fission (10). This reaction is all due to largely fixed graphite; graphite heat can make 5.5% of the thermal power (10). Fast breeder reactor use sodium or liquid metal as their coolant and has two methods of distribution (10). The liquid metal coolant is used to transfer heat from the core to steam that is used to power the electricity generators (10). There are multiple types of generators that may be used in order to produce energy. Most of the reactors have many similarities like the use or uranium (10). The only differences between these generators are the different types of coolants needed as well as the type of pressure it distributes. When thinking about the structure and the process of
conversion that these reactors do one must think about the financial cost it has to keep a power plant open and to have it working.

The downside of nuclear power is that it is not versatile there is still need of fossil fuel in production of transportation and pharmaceuticals (3). There is more abundance of coal that also hinders the need for nuclear power and the types of regulations that a nuclear plant needs to keep in order to provide safe energy is a disadvantage as well (3). In order to keep a power plant safe there is a lot of engineering needed to make sure that all the equipment can handle all the pressure and heat that may occur during the neutron chain reaction (3). If the engineering fails it can lead to a big problem because there will be release of radioactive waste and power plants may be shut down (3). Although fossil fuel has disadvantage that affect the atmosphere and ozone of the earth the disadvantage of nuclear power is the nuclear waste (3). The nuclear waste that comes along with this type of energy contains radiation that is highly dangerous (2). Compared to the waste fossil fuel will create, nuclear waste gives a big challenge when it comes to their disposal. The workers at these power plants have to make sure they do not come in contact with the radioactivity the reactors may have or they can die instantaneously (2). There is not a set system to get rid of the nuclear waste the only options are to store the waste until there is little to no radioactivity in it but that may take many years to occur and storage resources are scarce (10). The disposal of high level waste can be a hard to do because it requires the be buried in geologic repository that certain states may have but the people of those states do not want to live close or on radioactive waste (3). The most common disadvantage can be the cost to build and maintain power plants (3).

The benefits of using nuclear energy are that it reduces the use of green house gas emission and a reactor can make provide over a year of electricity to an average U.S household (3). Nuclear energy is used to provide electricity for factory, homes, and offices and this done by the use of fission breeder reactors that create steam which drive turbines that are attached to electrical generators (11). Nuclear energy provides about 20% of the united state electricity (3). There are many advantages to the use of nuclear energy such as less air pollution, low cost, and efficient use resources such as fuel and natural products (3). Nuclear energy doesn’t release gas emissions like coal burning would (3). Coal burning can carry different types of acid that can affect plant, lakes, river, aquatic life, and many man made structure (3). The way that the acid from coal burning can affect these things is because all the acid gets into the atmosphere but they return as acid rain back to the earth (3). Just like coal burning green houses release chemicals in to the atmosphere warming up the earths surface, which affect us in the long run (3). All these types of methods used to sustain energy release toxic and harmful chemicals into the atmosphere creating problem but nuclear power does not burn anything which makes it safe (3). All the steam that comes from nuclear power plant are not harmful pollutants (3). Although oil and natural gasses are used they aren’t a great source of energy they are used for small amounts of electricity and is produced in short periods (3). The use of nuclear power requires less quantities of material and the process of mining and transportation is simpler the reason being that 1kg of uranium is the same energy as 50,000 tons of coal was to be burned (3). Nuclear power leads to fewer deaths due to mining and transportation (3). Nuclear power is cheaper due to the need of uranium compared to coal.

Nuclear energy has been something that has been created and used recently. Power plants were granted license in 1989 to become fully functional (3). The process of obtaining nuclear energy begins with a natural product called uranium it goes through a process in order to become more fissile and have a better interaction with the neutrons (..). Plutonium is always a natural
product that is used and obtained. The retrieval of this natural product can be through a hole in the ground or mining using a technique called in situ leach (4). This technique makes it even less dangerous because nothing is removed only water is circulated in order to get the minerals from the orebody which is transported to the surface through a tube (5). Uranium if converted into uranium oxide and pressed into pellets that will be put into reactors (1). There are multiple types of reactors that have different coolants and may require different steps but the commonly used reactor is the pressurized water reactor that requires uranium fuel as well (10). These reactors convert nuclear fuel into energy; this fuel may stay in the reactors for years until there is no more use for the fuel (11). The process of removing what is left in the reactors is a dangerous process because whatever is left has radioactivity that is harmful (11). There are two options to do with the nuclear waste found in the reactors, it can either be recycled or reprocessed or it is disposed of and stored for long period of time (8). The recycling and reprocessing step can be beneficial because there is no waste of substances. The uranium that is recycled might not have to go through another process of conversion to uranium 235 because the recycled uranium has some of that uranium 235 in it (8). Recycling and reprocessing is easier because it reuses material as opposed to find places to store the waste. The process of storage and disposal is a dangerous one for the workers that do (11). The reactors contain radioactivity that can kill a worker if it is unshielded which is why they used water to cool down and transfer the waste (11).

There are many pros and cons to nuclear energy such as less toxic because there is no release of chemical pollutants into the atmosphere but compared to fossil fuel nuclear energy contain a lot of radioactivity and there is no way to properly dispose of it (3). Nuclear power plants are able to provide us with energy that doesn’t affect our earth because the steam that comes from power plants doesn’t contain all the pollutants that coal would (3). Not only is it less harming for the environment but also it’s less dangerous for those who are retrieving the natural products that are needed. Less people die in the retrieval of uranium because of in situ leach technique that doesn’t remove anything from the mines it just extracts the minerals (5). The cost to build these power plants and the maintenance may be too much but with correct supervision and engineering there are less accidents or errors to occur.

Nuclear waste is very controversial because many are afraid of the radioactivity the waste may carry but I believe it’s a way to a better and cleaner energy. The methods that are used now such as burning coal is slowly depleted our ozone and affecting the earth (3). I believe that in the next couple of years we will have no option but to turn to this form of energy because as the population increase so will the need of electricity and if we begin to burn more fossil fuel our environment will continue to deteriorate. I believe that through science and other engineering we are able to find a new way to dispose of the nuclear waste we have stored or find other uses for it. I understand that storing it underground is dangerous because if a leak occurs then radiation will leak out but I think that producing lower level waste can make it more beneficial. Lower-level waste will provide energy as well but it’s not so hard to get rid of as opposed to the high level waste that we bury. In order to find a fix we need to be able to experiment and I believe that if we were to shut down these nuclear power plants now we will not be able to find a solution and we would have gotten rid of a good source of energy. Through trial and error will there be a solution and I am positive that in the years to come their will slowly be a solution to the disposal of waste.
conversion process from uranium to fuel (4).

**Figure 1**

**Figure 2**

[Graph showing binding energy per particle and mass number]
This is the binding energy curve (7).

**Figure 3**

Neutron cross-section (8).

**Figure 4**

<table>
<thead>
<tr>
<th></th>
<th>Charge</th>
<th>Discharge</th>
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<tr>
<td>Uranium</td>
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<tr>
<td>Enrichment</td>
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<td>0.71%</td>
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<tr>
<td>Plutonium</td>
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<td>1.27%</td>
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<tr>
<td>Minor Actinides</td>
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<td>0.14%</td>
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<tr>
<td>Fission products</td>
<td>0.00%</td>
<td>5.15%</td>
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</table>

Type of waste is put into the reactors (2).
Figure 5

Radioactive decay (9).

Figure 6

\[
\begin{align*}
\text{UO}^{2+} \text{(aq)} + 2 \text{NO}_3^- \text{(aq)} + 2 \text{TBP (org)} &\rightarrow \text{UO}_2(\text{NO}_3^-)_2(\text{TBP})_2 \text{(org)} \\
\text{Pu}^{4+} \text{(aq)} + 4 \text{NO}_3^- \text{(aq)} + 2 \text{TBP (org)} &\rightarrow \text{Pu}(\text{NO}_3^-)_4(\text{TBP})_2 \text{(org)}
\end{align*}
\]

extraction of chemical process (12).
<table>
<thead>
<tr>
<th>Reactor type</th>
<th>Light water reactor (LWR)</th>
<th>Heavy water reactor (HWR)</th>
<th>Graphite moderated reactor (GMR)</th>
<th>Fast breeder reactor (FBR)</th>
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<td>Boiling water reactor (BWR)</td>
<td>Pressurized water reactor (PWR)</td>
<td>Gas cooled (GCR)</td>
<td>Water cooled</td>
<td>Electricity; plutonium production</td>
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<td>Heavy water (D₂O)</td>
<td>Water</td>
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<td>Gas (CO₂ or helium)</td>
<td>Molten, liquid sodium</td>
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<tr>
<td>Moderator type</td>
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<td>Heavy water</td>
<td>Graphite</td>
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<td>Fuel-chemical composition</td>
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<td>Uranium dioxide (UO₂)</td>
<td>Uranium dioxide (UO₂) or metal</td>
<td>Uranium dioxide (UO₂) (RBMK) or metal (N-reactor)</td>
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<td>Low-enriched</td>
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<td>Slightly-enriched natural uranium</td>
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<tr>
<td>Enrichment level</td>
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<td>Slightly-enriched</td>
<td>Various mixtures of (^{239}\text{Pu}) and (^{235}\text{U})</td>
<td></td>
</tr>
</tbody>
</table>

Type of reactors (12).
Reference:


7. Nuclear Fission [Internet]. [cited 2015 October 2]. Available from: http://hyperphysics.phy-astr.gsu.edu/hbase/nucene/fission.html


Where We’ve Been, How We Got There, and Where We’re Going

JAKE MATHER
ASTRONOMY 112-12666
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DR. SHERRY
In today’s society, it seems strange to believe that we are living in a world completely different from past generations. A mere century ago the idea of going outside of our atmosphere in a space ship would have been considered Science-Fiction. However, now there is an entire sector of every nation’s government strictly bound to the ventures of space travel. John F. Kennedy once said, “The exploration of space will go ahead, whether we join in it or not, and it is one of the great adventures of all time, and no nation which expects to be the leader of other nations can expect to stay behind in the race for space.” This is the final frontier for the human species, to begin the colonization of the stars. As one looks from the past to the present, one can only see improvements and endless opportunities awaiting us in the future. Not from a national point-of-view, but rather from a global perspective.

The true beginning of the Space age would be marked by the U.S.S.R.’s successful launch of the Sputnik 1 artificial satellite in 1958. Space programs around the world launched robots and animals into outer Space, in order to test the feasibility of one day putting humans aboard those ships. Not to mention began testing the hopeful candidates who wished to man these missions into Space. The Soviet Union launched a Terrier named Laika that they found on the streets of Moscow in Sputnik 2, while the American’s launched a chimpanzee by the name of Ham on one of their Mercury missions. Sadly Ham was the only one of these animals to return back to Earth. Laika died aboard Sputnik 2 in 1957, never able to see the Moscow streets again.

Many of the early manned missions were flown in wingless, cone-shaped capsules, which would detach from the launch rockets and fall back to Earth. These small, rudimentary capsules were designed to withstand the tremendous temperatures from re-entering the atmosphere as well as surviving the Ocean landing. And the true meaning of that landing would be a quite dramatic splash into the water, from the far reaches of our atmosphere with only parachutes to slow one’s decent. Definitely a huge difference from the landings of current space crafts.

On April 12th, 1961, the Soviets beat out the United State’s Space program NASA by making cosmonaut Yuri Gagarin the first human to orbit the Earth in the Vostok spacecraft for 108 minutes. However, a mere month later American astronaut Alan Shepard Jr. followed Yuri’s momentous lead with a 15 minute flight aboard the Freedom 7. Alan Shepard’s flight was watched on television on May 5th, 1961 by some estimated 45 million viewers. The first American to actually orbit the Earth was John Glenn occurring on February 20th, 1962. He orbited the Earth a stunning three times in his spacecraft named the Friendship 7. His flight lasted a total of 4 hours and 55 minutes.

Proving that humans were capable of venturing out beyond our atmosphere was only the beginning. Many other benefits came along with Space exploration. Satellites had been close to perfected and were starting to be harnessed for not only entertainment purposes but for connectivity as well. Before a manned mission even took place, un-manned spacecrafts were photographing and probing the moon in the early 1960’s. On July 20th, 1961, President John F. Kennedy set a national goal by declaring “Landing a man on the moon and returning him home safely within a decade.” This tremendous feat would essentially mark the start of both the Gemini and Apollo missions for NASA.
Over the next eight years, several Apollo missions were launched and carried out, however the national goal set by President Kennedy had yet to be achieved. This was until the Apollo 11 mission launched. Crewed by Neil Armstrong, Michael Collins, and Edwin “Buzz” Aldrin, they launched from Cape Kennedy on July 16th 1969, and their primary mission was to complete a lunar landing and return to Earth. The Apollo 11 Mission was a total success, and it truly was “a giant leap for mankind.” This mission didn’t only prove what people are capable of, but expressed the advancement of Space programs after only a decade of true Space exploration. Without a doubt the moon landing was one of the greatest accomplishments ever made by human-kind, and sparked the desire of Earthlings to venture out even further into the unknown.

In July 1976, NASA’s Viking probe 1 touched down on Mars. It was the first man made object that had a soft landing on the red planet. Though the Soviet Mars 2 and 3 probes touched down on the planet, they failed to work shortly thereafter. The Viking spacecraft sent back the first color pictures we have of the red planet. And actually has the longest running mission time on the Mar’s surface, lasting a total duration of 6 years and 116 days. Quite a bit longer than any mission length before.

One of the main missions that followed the lunar landing less than a decade later would be that of the Voyager and Pioneer missions. The Pioneers were sent to find out more about the gas giants we know as Jupiter and Saturn. Pioneer 11 took the first picture of Jupiter’s giant red spot, and when it arrived at Saturn it not only discovered a few moons that were unknown, but also an entirely unknown ring around the planet as well. Both of these probes however have stopped sending data and are currently flying out past our solar system, never to return again.

The overall goal over the Voyager mission was to extend the NASA exploration of the solar system beyond the Sun’s influence. The Voyagers 1 and 2 were launched in August and September of 1977 and have been sending back information ever since. The Voyager itself is made of three-axis stabilizing systems that utilize gyro referenced altitude control to maintain the antennas pointing towards Earth. Communication with these crafts are still documented due to the DSN or deep space network. The primary mission of Voyager was to explore Jupiter and Saturn, yet it was extended to explore Uranus and Neptune. To this day, it is still the only spacecraft to venture out and past the outer planets. As of 2012, Voyager 1 made it’s way outside of our solar system and as Dr. Ed Stone said, “Voyager is in interstellar space - the space between the stars.” Officially these crafts have gone further than anything or anyone in history.

After the fast paced manned missions of the 1960s and 1970s NASA geared more of their attention towards both preparing and launching the spacecraft that would change the way we view our capabilities in orbit, the space shuttle. The Columbia was the new craft that would carry astronauts outside of our atmosphere and carry out missions. The first mission was launched on April 12th, 1981. The Columbia stayed in orbit for 2 days before safely landing back on Earth at Edwards Air Force Base in California.

Within a few years of it being commonplace to have our capsules splashing into water as landing procedure, the scientific minds of NASA’s best employees devised a way to have a controlled liftoff and landing. The Columbia without a doubt was one of the biggest
improvements to Space exploration thus far, not only was the ship able to be re-used but flight complication became less and less by actually giving complete control to the pilots of these massive shuttles leaving and returning our planet.

Several other mission were ran during the mid 80s, including the Challenger missions. One of these missions even gave the crew the ability to capture and repair the SolarMax spacecraft, a device which monitored the sun during it’s solar peak, they were able to prolong its orbital lifetime. However, the nation was all reminded that nothing about Space exploration even down to the launch is safe. On January 28, 1986, the Challenger and it’s 7 crew members for the STS-51L mission. On a day just slightly above freezing temperatures, the crew and space center prepared for launch. A mere 73 seconds later a failure in the rocket booster O-ring caused a fireball to completely destroy the vehicle and the entire crew. Commander Francis Scobee, Pilot Michael Smith, Mission Specialists Judith Resnik, Ellison Onizuka and Ronald McNair, and Payload Specialists Gregory Jarvis and Sharon Christa McAuliffe, a teacher. All of these crew members lost their life on that cold January day, ever reminding the world the price that can be payed for knowledge and discovery.

Following John F. Kennedy’s presidential lead, Ronald Reagan set the proverbial stage for future missions by again presenting a national goal during his state of the union speech in late January of 1984 by challenging NASA to develop an internationally permanently crewed Space station. "A space station will permit quantum leaps in our research in science, communications, and in metals and lifesaving medicines which could be manufactured only in space," Reagan stated. "We want our friends to help us meet these challenges and share in their benefits. NASA will invite other countries to participate so we can strengthen peace, build prosperity, and expand freedom for all who share our goals.”

The Magellan mission was launched in 1989, headed for Venus to better map the surface of the planet and to determine the topographic relief. They wanted to know the overall makeup of the planet including it’s tectonics and chemical processes. The Magellan covered 98% of the planets total makeup before radio contact was lost in 1994.

Along with the Magellan, the Galileo was launched in October 1989 to study the planet of Jupiter. This mission was very successful, not only for the initial intentions but also because of the discoveries made along the way. The Galileo had the first recorded contact with an asteroid watching it collide with a planet. It is also solely responsible for the discovery of the saltwater ocean beneath the Jupiter’s moon Europa’s icy surface. Galileo plunged into the grips of Europa on September 21, 2003 ending it’s extensive and productive lifespan.

During the 1990s, the number of space shuttle missions nearly doubled that of the decade previous. This offered more data and attention to be payed to our scientific wonder. In 1990 the first great outer space observatory was lunched with the ever famous Hubble telescope. The impact of this launch is quite momentous in the view of what has been seen and discovered through this telescope. Especially the evidence provided to turn the theory of black holes into fact. The Hubble has been re-visited several times to provide maintenance and updates to the telescope.
NASA launched two other observatories as well, including the Compton Gamma Ray, and the Chanda X-Ray. The Gamma Ray observatory was used to better understand and investigate high radiation sources. Whereas the Chandra was designed to study black holes and supernovas. Even dark matter could be better explored in optics that have yet to be used.

Most of the focus throughout the 90s was building the ISS (International Space Station), the United States and Russia had an agreement for NASA’s astronauts to utilize the Russian Space Station Mir. This would assist the US with understanding and experiencing long duration missions.

In 2003 NASA’s Columbia exploded upon re-entry to Earth, and killed all 7 crew members. This accident once again took the world’s breath away and shut down shuttle missions for over two years, and completely halted the construction of the ISS.

Later in 2003 China joined the US and Russia with actually putting a human into space. Yang Liwei completed 14 orbits of the Earth, on October 14th, he joined a very small group of humans who can claim that they have been outside of our atmosphere.

One of the most iconic points of Space travel through the first decade of the 21st century would be the completion of the International Space Station. Upon completion in 2009, the ISS has had a constant crew ranging from 2 to 6 people constantly on the station throughout the decade. This is by far the most complex and in depth engineering process any international effort has completed. Not to mention the largest structure that humans have yet to put into Space, roughly covering the space of a football field and weighing nearly one million pounds.

According to NASA’s administrator Charles Bolden in 2015, he says that the new direction for exploration will be to send humans further from our planet than ever before. He speaks of sending people to an asteroid and Mars. They are developing the most advanced spacecraft and rocket aptly named the Orion. Bolden says "As a former astronaut and the current NASA Administrator, I'm here to tell you that American leadership in space will continue for at least the next half-century because we have laid the foundation for success - and failure is not an option."

Another truly exciting mission in the future of NASA, is the Asteroid Redirect Mission, which entails the first ever mission to capture and redirect a near Earth asteroid and direct it to a new stable orbit around the moon. Then have astronauts further explore it in the 2020s.

Every exciting thought to the future of Space travel, can be traced back to it’s roots. In order to have a fruitful future one must be well versed in the past. It is the curiosity and quest for knowledge that sparked this departure from our home planet. One can only be fully intoxicated with the prospect of what the future in Space travel holds, but when one looks back at the past and beginnings of Space exploration they see that there is no limit for what can be achieved as a global unit.
Works Cited


PHOTOGRAPHY: SAVING THE LIGHT

Neda Milešić

April 21, 2016

Physics 112

Professor Swingler
Abstract

Optics provide an important insight into the properties of light, such as reflection, bending, refraction, and diffraction. Light is an electromagnetic wave, which through different frequencies displays different characteristics that also include the visible light spectrum. Human eyes and cameras equally capture this light spectrum, with the final goal of retaining an image. Different types of camera lenses that follow specific optics laws help direct the light rays in the appropriate direction. Camera can be manipulated through varying the shutter speed, aperture and the sensor sensitivity to achieve a certain photographic look. Through manual control, this exposes the camera sensor to the light, which excites the cells to later store the excited charges in memory cards, later reproduced as final images.

“Let there be light”

The light has been celebrated since the dawn of time, through many civilizations, and in various religions. It has held the highest distinction in the human world, representing the good, the purest above all, and the life itself. Vision, as one of the most important human senses, has been a symbol of divinity and protection since the Egyptian and Greek times (Ancient Egypt Online 2010; Amvrazi A 2016). Early Greek philosophers observed and described many natural phenomena, including the light, thus paving the way to the study of modern physics. The studies were later expanded during the Renaissance era, giving scientists like Newton, Galilei, Copernicus, Kepler, and Snell, a chance to solidify the actual mathematics (Waldman 1983), thus providing scientific proof behind optics theories.

Per Newton, the light was initially considered to be corpuscle, a packet of matter, traveling through the air in a straight line and hitting a certain object without the ability to bend. (Waldman 1983). Given Newton’s reputation, this theory was widely accepted until Christian Huygens, a Dutch scientist, started his work in optics and formulized the idea of a light as a wavelike movement through the air (Davidson 2015). Even though both theories had good basic explanations, the wave theory was finally accepted by majority and it is generally used to this day.

The wave of light is known as a transverse wave (Serway and Vuille 2012), as it moves in both up and down motion, as well as in a perpendicular direction to this up/down plane (Figure 1). The up and down wave is the electric field radiated by certain charge, while the perpendicular wave is a magnetic field, which exists as a result of a moving charge. The sum of both, with the accelerating charge, finally results in electromagnetic wave, or the light. The upper peaks are called crests, while the lower are called troughs (Serway and Vuille 2012). The magnitude measured from the bottom of crests/troughs is known as the amplitude of a wave, while the distance between two crests/troughs is known as the wavelength.

It is this property that is one of the most identifying and most useful factors of light. The wavelength pinpoints the exact types of electromagnetic waves, such as radio waves, ultraviolet, gamma rays, or infrared waves, but it is the visible light that is the primary interest in the photography world. The components of the visible light are usually seen in rainbows, something Newton first experimented with, when placing a prism in front of a beam of light (Serway and Vuille 2012). This provided the opportunity for each color to exhibit their own characteristics through dispersion (Illustrated Magazine of Art, 1853). The visible light spectrum spans from 700 nm (longer waves) to 400 nm (shorter waves), and includes colors red, orange, yellow,
green, blue, and violet (Serway and Vuille 201) as indicated in Figure 2, and any combination in between that can be captured, modified, and molded with equipment, to achieve the final photographic look.

“Drawing with light”

The combination of all of the 6 colors, usually known as the white light (Serway and Vuille 2012), has interesting properties. It not only moves through the media in the fastest speed known to a man, but it can also reflect, bend, refract, and disperse. The ability to reflect, for instance, is what gives us the ability to distinguish between the colors, as the medium absorbs every single color, except the one reflected which results in what we see as a green bush or a red rose (Serway and Vuille 2012). What is reflected is what we see. When this light reflects, the angle of light impact to the normal on the plane (called the angle of incidence) will be equal to the angle of reflection (angle that results after the reflection).

On the other hand, when the light ray passes from one medium to another made up of a different substance, it refracts, and continues in a different path than the original, bending under the certain angle that can be determined using simple trigonometry of Snell’s Law (Kingslake 1992), which states the following:

\[ n_1 \sin \Theta_1 = n_2 \sin \Theta_2 \]

where \( n_1 \) and \( n_2 \) are specific indices of refraction of specific material, while the angles are of incidence and refraction respectively (Serway and Vuille 2012). Snell’s Law was an important contribution to the field of optics and it is today used in technology.

The reflection, refraction, and general movement of light, including the human fascination with the ability to process this through an eye, is what led to the first pinhole cameras, while the bending properties helped develop first photography techniques and with that, lenses.

Beginnings

Although much less sophisticated, cameras are recreations of a human eye. When eye, the most complex camera in existence, is exposed to light through the lens, this light passes to touch the sensor of the eye, or retina. Retina picks up the light with numerous receptors, or rods and cones, that further send action potentials through a series of dedicated neural networks, to finally leave an imprint in the occipital region of the brain (Clancy and McVicar 2002). This process creates a visual memory, a picture of what one sees in front of them, just as a camera would on a paper or a monitor.

First photographers attempted to replicate this method using cameras, films and light-sensitive silver halide emulsion (Introduction to photography: Differences between analogue and digital photography 2016). The light was projected onto the film, which was then transferred to the paper, after being processed in special dark rooms to avoid excess exposure. The final paper could be reproduced as a captured image. This process required special equipment, more room, and special chemicals, and later became known as the “analogue” photography due to the type of instruments used, as well as the proportional input/output signal.
that resulted from exposing the camera to the light. In other words, as more light hits the sensor (e.g. the light meter) this produces excitement in charges and electrical current as a response. This analogue signal is essentially what is produced when retina is stimulated in the eye. With cameras, the current is the “action potential” that occurs once the charge on the camera is exposed to the light and excited.

Analogue or film photography was used for several decades, until computerized systems were developed and perfected to a point where the first digital camera was created. This new era brought new possibilities and modern equipment, but the general idea behind photography remained the same. Whether the output is directly on the paper or the screen, the purpose of the camera is to freeze the picture by capturing the light.

**Optics**

The light is reflected or refracted in front of eyes, allowing us to see the object ahead. This recording, or perspective, is essentially a two dimensional representation of the three dimensional object (Kingslake 1992). The opening through which the light passes is called the center of perspective (Figure 3). When the light passes through the opening of this center, it falls on the recording plane behind. In the case of photography, this is a digital sensor or film. The size of the object will be proportional to the object recorded, but it will depend on the object position and distance from the lens opening (Serway and Vuille 2012). The perspective or the recording can thus be adjusted to appropriately depict the idea and proportions of the object ahead with respect to its surroundings, something which can be achieved using various types lenses or special types of cameras, as well as angle manipulations.

**Lenses**

Although every lens has unique characteristics, the basic idea is the same - lenses are sets of optical glasses that project the light to the designated sensor (film or digital), to capture the final picture. Each lens contains concave and convex glass lenses that can be manipulated to create a lighter or a darker image (National Geographic 2012). Lenses are categorized based on their coverage or “angle of view” that can be expressed through their focal length. Focal length represents the distance between the lens center to the focus on the sensor, and it is measured in millimeters. Based on this, the camera lenses are divided into: standard and fixed (50 mm), zoom (variable focal lengths), telephoto (70-300 mm), wide angle (shorter focal length), and macro (for close up shots).

Just as the eye accommodates its lens in order to properly focus on the closer or further image, the camera lens can be adjusted. Each has the near and far points, the closest and further points respectively, to which the lens can focus to create a sharp image on the sensor (Serway and Vuille 2012). The image sharpness is a result of the refracting light through the lens that further converges into a single point, also known as the focal point (Kingslake 1992). The distance between the lens and the focal point, known as the focal length, has the anterior side located in the front of the lens where the light hits the lens plane, and posterior being in the back, after the light refraction through the lens. When each ray of light parallel to the axial plane hits the lens, then refracts, it meets the other light rays at the focal point (Figure 4), where
the final picture will have the sharpest outcome. Any light falling outside of this area, leads to chromatic and spherical aberrations, which result in dispersion of color and lack of sharpness.

Chromatic aberration is a result of different colors refracting without intersecting at the same focal point, the effect of different indices of refraction for each color wavelength, which in photography translates to color fringes on the final picture (Miller et al. 2010). Spherical aberration follows the similar concept, except the image is unable to focus to full sharpness, as the light rays fall on separate areas of the axis. The correction can be made in camera lenses by using orthoscopic doublets (Hyperphysics: Spherical aberration 2000), a combination of two lenses that together rectify the focus issue. In biology, such issues result in poor eyesight and the need for corrective lenses or eyeglasses.

**Lens specifics**

There are 2 types (and 6 subtypes) of lenses: converging (biconvex, convex-concave, plano-convex) and diverging (biconcave, convex-concave, and plano-concave) (Serway and Vuille 2012). Special optical glass is used to manufacture photographic lenses, while different combinations of these types of thin lenses can be used to create different types of photographic lenses.

As indicated above, the light parallel with the optic axis will refract and pass through the focal point, but the light passing through the center of the lens will not be deflected and will meet the parallel light ray to produce the tip of the image (Figure 4). The series of formulas used to find the angles between these rays and axes lead to the equation for magnification of the lens (Serway and Vuille 2012):

\[ M = \frac{\text{Size of the image}}{\text{size of the object}} = -\left(\frac{\text{Distance of the image}}{\text{distance of the object}}\right) \]

The magnification of the real image is negative, as the final image is inverted on the sensor. As a specific example, if the size of the object is 10 meters, size of the image is 5 meters, and distance of the object is 50 meters, the final size of the image will be – 25 meters.

By combining the values and extracting the equations, the final thin-lens equation can be concluded as:

\[ \frac{1}{d} + \frac{1}{d'} = \frac{1}{f} \]

where \(d\) is a distance of the object, \(d'\) is distance of the image, and \(f\) is the focal length. For example, the distance of the object is 10 meters and distance of the image is 5 meters, the final focal length will be 3.33 meters.

Converging lens will always have a positive focal length, while the diverging lens will have a negative focal length (Serway and Vuille 2012). These concepts are very useful in photography, where the appropriate focal length can be chosen based on the distance of the photographed object, providing the photographer with a full control over the quality of the final image.

Finally, to create the thin lens, an index of refraction of the material, as well as the curvatures properties (i.e. whether it is biconcave, plano-convex, etc), have to be taken into consideration. For this purpose, the lens-maker equation is used:
$1/f = (n-1) \left( 1/R_1 - 1/R_2 \right)$.  

In this case, $n$ is the index of refraction of the specific material, $R_1$ is the radius of the front curvature of the lens, while $R_2$ is the radius of the back curvature. This helps determine the focal length, which later provides helpful information to manufacturers and users (Subrahmanyam et al 2008, Serway and Vuille 2012).

**Getting framed**

Once the light passes through the intricate maze of the lenses and focuses appropriately, it hits the reflex mirror of the camera. In order to see the image through the viewfinder, the light bounces off the pentaprism (five-sided prism that reflects the light), as indicated in Figure 5. (National Geographic 2012). This allows the photographer to define the composition and see the light, before making any adjustments. These adjustments are manually controlled by well versed and professional photographers, but most cameras today have built in program settings that allow amateur photographers to utilize what is typically known as a “point and shoot” method. Since the idea behind photography is to control the light, three factors are taken into consideration when making adjustments.

Aperture represents the opening of the lens that can be manipulated by closing and opening, in order to control the amount of light that touches the sensor (Serway and Vuille 2012). The sizes of aperture vary from $f/1.4$ to $f/22$ and are called the f-stop numbers. The f-stop values are calculated using the $f/d$ formula, where $f$ is the focal length and $d$ is the diameter of the aperture (National Geographic 2012). Large values represent small aperture, which translates into small diameter of the opening that further reduces the aberration and creates sharper images, as well as lots of depth in the image. This distinguishes the object of interest from its blurry surroundings and brings it into the visible spotlight on the final photograph. Consequently, the small f-stop value represents a large aperture that results in little depth of field and more overall image in focus. Controlling the depth of field by manipulating the aperture creates pleasing effects, from landscape photography, where large aperture is needed to bring the whole scene into focus, to macro photography, where a dew drop is isolated on a branch in the picture.

Shutter speed is the second factor involved in controlling the amount of light that reaches the sensor (Serway and Vuille 2012). As the term indicates, the speed of the shutter is manipulated. For faster freezing of the motion, methods often used in sports photography for instance, a fast shutter speed is needed. The amount of light that reaches the sensor is brief, but sufficient to capture the final image. On the other hand, in lower light conditions, slower shutter speed is required to let in more light. For instance, indoor or night photographers utilize this to capture their scenes. Shutter speed values vary from camera to camera, starting around several second exposures for slow shutter speeds to more than $1/1000$ of a second for faster counterparts (National Geographic 2012). Unfortunately, slower shutter speed is also very sensitive to movement and usually requires support such as tripod, since the light will not be falling on the same focus spot and out of focus images can easily be produced as a result.

However, besides adjusting the aperture in addition to shutter speed, in order to create the appropriate photographic effect, the final of the three light-controlling factors can also be manipulated - the sensor sensitivity or ISO. As already indicated, the sensor acts as retina in the eye, a receptor of the light rays that produces excited impulses, that are later translated into data.
of an image and stored in a memory card of a camera (analogue cameras used film for these purposes). Controlling the sensor sensitivity (ISO) in turn controls the exposure and how sensitive the camera sensor will be to light (National Geographic 2012). Film photography was limited to utilizing film of specific ISO values, which limited the user since they were forced to use the whole film before moving onto the next one. Digital photography is much more advanced in those terms and provides a range of options with only few button manipulations. ISO values typically begin at 100 and reach 100,000 in some cases (National Geographic 2012). The smaller values are appropriate for brighter light, sunny days. Moderate values that can be utilized indoors are around 400 - 800 ISO, with upper values typically being used in a very dark environment, artificial lighting, and night photography.

With these three components of the camera set in place, the depressed camera button can finally raise the mirror that was initially deflecting the light and the image to the viewfinder. This results in a shutter speed opening at a determined speed, aperture of the lens opening its designated diameter, and the light hitting the sensor at the appropriate strength, to capture and record the final image.

Conclusion

Digital age is the age of innovation and affordability. Analogue photography, although still in existence, has become a luxury and special interest. Films have become less frequent in photographic worlds, but to some, this is still a preferred method of capturing images. The special effects analogue cameras and different films produce, including specific colors, are still considered the best in some opinion. However, digital camera has officially become the most used photography equipment today. After stepping in place of analogue photography and creating revolutionary methods of capturing images on digital sensors, digital cameras brought new technological advances, more lens options, and more interplay with the light. The cameras have become so advanced, they even led to the evolution of smartphones, which along with the social media brought photography to the new level, making it accessible to anyone and everyone.

Photography equipment has never been more affordable, and more and more individuals are investing into great cameras or upgrading to better camera phones for both professional and personal use. Photography has become part of everyday life and modern documentation methods. Professional photography can still distinguish itself with more powerful models of cameras, but in essence, whether one is documenting the weddings, current events, or their daily meal, the idea of photography will never change.

As more experiments are done in the field of optics and light, this research will be extended to photography to better the lenses and camera fragments. This will also improve the user friendliness, sizing of cameras and lenses, and bring final images closer to the level of eye perfection. However, perfection itself will always remain straightforward - a beam of light, passing through the lenses of various shapes, adjusting to these special glasses, refracting to a certain point, focusing appropriately to excite the receptors on sensors, and to further send the signal that will create that one final desired light image, known as a photograph.
Figures

**Electromagnetic Waves**

![Diagram of an electromagnetic wave showing magnetic and electric fields, wavelength, and wave direction.](https://lcogt.net/)

Figure 1. Transverse light wave diagram. (Retrieved from https://lcogt.net/).

![Diagram of the electromagnetic spectrum showing wavelength in meters and frequencies in Hz.](http://www.ozone-hole.org.uk/02.php)

Figure 2. Electromagnetic spectrum (Retrieved from http://www.ozone-hole.org.uk/02.php).
Figure 1.1. The meaning of “center of perspective.”

Figure 3. Perspective and image capture (Kingslake 1992)

Figure 4. Anterior (object to lens) and posterior (lens to image) focal lengths (Retrieved from http://www.oocities.org/rjwarren_stm/Physics_Notes/ThinLens.gif).

Figure 4. Anterior (object to lens) and posterior (lens to image) focal lengths (Retrieved from http://www.oocities.org/rjwarren_stm/Physics_Notes/ThinLens.gif).
Figure 5. Interior of digital camera and camera lens (Retrieved from https://aminna1998.files.wordpress.com/2013/10/dslr-diagram.jpg).
References

Amvrazi A. 2016. The eyes have it: the evil eye in Greece [Internet]. Available from http://www.athensguide.com/journalists/articles/evileye.htm


An Overview of Styrene Migration
From Polystyrene Containers into Microwaved Foods

Alee Monaco

Dr. Scott Massey
Chemistry 152
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Abstract

Polystyrene is a thermoplastic that is commonly used for food packaging. It can take on different material properties depending on the manufacturing process. There has been sufficient evidence that styrene monomers and oligomers can migrate from polystyrene containers into the food they contain, especially when undergoing a heating process. This study was conducted to determine if the type of polystyrene container affects the quantity of monomers and oligomers that migrate from food containers into food simulants when microwaved. The H-NMR that was used to test this was not sensitive enough to detect the concentrations of styrene that may have migrated. Small peaks did appear in the same area that styrene is expressed, however the noise to signal ratio was too high to conclude a result. It is possible that a more sensitive H-NMR could detect such concentrations in a replication study.

Introduction

History of Polystyrene

Polystyrene is a widely used thermoplastic polymer that is composed of styrene monomers. It was originally discovered on accident by the German apothecary Eduard Simon in 1839 when he distilled an oily substance from the resin of the Turkish sweetgum tree. He originally thought it was a monomer, so he named it after the tree’s resin, storax, calling it ‘styrol.’ He later found that it had thickened into a jelly-like substance and assumed that it had oxidized, to which he called the substance ‘styrol oxide.’ It wasn’t until 1922 that German chemist Hermann Staudinger identified the material as a polymer and recognized its thermoplastic properties which allow it to be easily molded into a rigid shape. This discovery led to commercial manufacture in Germany in 1930, and it eventually made its way to U.S. markets in 1937.

Polystyrene is now one of the most commonly used plastics in the world. It is cheap, lightweight, and versatile; therefore, it is used in several industries such as construction, medicine, electronics, domestic appliances and packaging. Food packaging alone accounts for thirty-seven percent of its production, as much as the other four industries combined. It is also used for several miscellaneous purposes which do not fall into any of these categories. (See Figure 1.)

Chemical Properties and Formation of Polystyrene

The monomer styrene is an organic hydrocarbon molecule that occurs naturally as a colorless or yellow, oily liquid. Its molecular formula is C₈H₈, and its structure can be represented by the chemical formula C₆H₅CH=CH₂. (See Figure 2.) This structure is essentially a benzene ring with a dehydrogenated ethylene and can be considered dehydrogenated ethylbenzene. It is hydrophobic and insoluble in water. Other names for the molecule are phenylethylene, phenylethen, ethenylbenzene, cinnamene, vinylbenzene, and styrol.

Polystyrene is most commonly formed by the process of free radical polymerization of styrene monomers. This process requires a free radical which is a molecule with an unpaired electron in its outer valence shell. The free radical, which is highly reactive, initiates a polymer chain by pairing with a single electron from styrene’s carbon-carbon double bond. This satisfies the octet of the free radical, but leaves an unpaired electron on the styrene monomer, thus forming another free radical. This new free radical with the initiating molecule will pair with yet another styrene monomer in the same way, and the pairing process will continue creating a chain
of styrene monomers. (See Figure 3.) The addition of these monomers to create a chain is called propagation. The chain finally ends at the termination stage when the last free radical finds another growing polymer chain’s free radical to pair with.\(^5\) For production purposes, terminator compounds can be added to prevent the formation of overly long chains as they may not melt properly.\(^6\)

There are different industrial methods of free radical polymerization that can affect factors such as viscosity, contamination, molecular weight, and thermal control of the reaction. The most common methods are bulk, solution, emulsion, and suspension polymerization. Bulk and suspension polymerization seem to be used most commonly for styrene polymers that are not in conjunction with other monomer types. Bulk polymerization produces the least amount of contamination as it only uses an initiator molecule, styrene monomers, and heat. It is easy and cost-effective, yielding large quantities of product. Its disadvantage is that the viscosity of the reaction mass increases drastically, preventing free radical ends from finding each other and in turn making severely long polymer chains. This makes the weight distribution of the polymers uneven and the overall reaction difficult to control. Suspension polymerization employs a water solution and water soluble compounds to initiate the free radicals. It reduces viscosity, so it is easier to control and yields similar molecular weights. It produces small, spherical, bead-like polymers that are useful for certain products as will be discussed later. The disadvantage to suspension polymerization is that it has a high chance of contamination and cannot be used for many plastic polymers due to temperature constraints.\(^7\)

**Residual Monomers**

When polymers of non-gaseous monomers are made, some residual monomers get trapped between polymer chains. Residual monomers are unreacted with the polymer product and are suspended in the macromolecule. The type of polymerization the molecules undergo can affect the amount of residual monomers left unreacted.\(^8\)

Residual monomers can also result from depolymerization, which is the decomposition of bonds between monomers. Every polymer will undergo depolymerization at a sufficiently high temperature. In the case of thermoplastics, polymer chains are entangled, not straight, and are held together by weak Van der Waals forces.\(^9\) Heat can break these intermolecular bonds, then further break the bonds between monomers themselves leaving partial polymer chains and residual monomers.\(^10\) One study suggested that more free-standing monomers were generated by depolymerization than were formed by the original polymerization process of polystyrene.\(^11\)

**Types of Polystyrene and Their Uses**

There are two general categories of polystyrene: rigid and foamed. Both have thermoplastic properties which allow it to undergo the melting and cooling process innumerable times. They can be reshaped and recycled almost indefinitely without severe damage.\(^1\)

Unmodified polystyrene, otherwise known as general purpose polystyrene (GPPS) is hard and brittle at cool temperatures, but can be heated to a liquid and molded into any shape.\(^12\) These shapes can serve as petri dishes, CD cases, TV housing, disposable cutlery and service ware, and countless other items. It is considered to be one of the rigid forms.\(^3\) Polystyrene can also be polymerized with a rubber such as polybutadiene to form a stronger, more flexible plastic called high impact polystyrene (HIPS). This rigid form of polystyrene has a higher impact strength and can assume varying stiffness based on the polystyrene to rubber ratio. It is commonly used in toys, signs, and yogurt containers.\(^13\)
Polystyrene can also assume a foam consistency which is lighter and more compressible than the standard rigid form. The foamed form has excellent insulative properties due to the process by which it’s made. This process begins with the formation of polystyrene balls through suspension polymerization. These balls are about the size of sugar granules. They are injected with a hydrocarbon blowing agent such as pentane. They are then steamed which expands the hydrocarbon gas inside the balls and increases the volume of each ball to 40 times its original capacity. As it cools, the polystyrene hardens at its new, increased size and the final product is a larger gas-and-air filled ball with a thin polystyrene coating. (See Figure 4)

Gases, including air, are poor thermal conductors. Their molecules are not as close together as those in solids or liquids, so it is less likely that other molecules will collide with them and produce or consume energy. This makes foamed polystyrene a great insulator as its gases prevent hot and cold molecules from reacting and changing temperature. Additionally, the long, tangled polystyrene chains create an elaborate barrier to the surrounding temperature, trapping in the heat or cold of the product it contains.

There are two types of polystyrene foam: expanded polystyrene (EPS) and extruded polystyrene (XPS). They are both commonly referred to as ‘styrofoam,’ but the term is widely misattributed. Styrofoam specifically refers to extruded polystyrene and is a registered trademark of Dow Chemical Company. Extruded polystyrene is much more dense and compacted than expanded polystyrene, leaving less space between the foamed polystyrene balls. It is primarily used for building insulation, shipping containers, and crafting materials. On the contrary, expanded polystyrene is more permeable and light. It is more commonly used for consumer goods such as disposable coolers, coffee cups, and take-home food containers. (See Figure 5.) The chemical processes by which EPS and XPS are formed are considerably different. While they both begin with the same blowing agent process, they undergo different finishing processes, thus they have different material properties.

**Health Hazards of Polystyrene**

Both rigid polystyrene and foamed polystyrene are used for food packaging. High impact polystyrene is mostly used for dairy products such as yogurt and margarine containers, general purpose polystyrene is used for disposable service ware, and the expanded foam form is used for hot beverages, take-out containers, and egg and meat packaging. The safety of polystyrene as food storage material has generated controversy since the early 1970’s. The effects of inhalation of styrene in plastic-factory workers has been studied extensively and has shown that the monomer can cause several neurological and lung problems. Some reported neurological problems include decreased reaction time, trouble concentrating, and poor motor skills particularly in relation to balance. Reported lung problems include decreased volume capacity and reduced flow rate. Furthermore, the Department of Health and Human Services (DHHS), National Toxicology Program (NTP) and The International Agency for Research on Cancer (IARC) have suggested that styrene is a “possible carcinogen.”

In contrast, there has been limited research on the ingestion of the monomer. One study was done on a population in Spain whose water had been contaminated with high concentrations of styrene. The population experienced acute irritation of the nose, throat, eyes, and skin. Most studies showing more long-term effects have been performed on non-human animal models. Styrene ingestion in these animal models disrupted immune and nervous systems and caused liver and kidney damage. There is some evidence that styrene-oxide, which is the product of
styrene metabolism, binds covalently with cellular proteins such as hemoglobin in mice.\textsuperscript{23} Despite these results, the Agency for Toxic Substances and Disease concluded that, “Based on the animal data...the oral toxicity of styrene in humans would be expected to be low to moderate.”\textsuperscript{20} 

In recent years, the use of expanded polystyrene foam as disposable food containers has been banned by nearly 100 cities in the United States and other countries across the globe.\textsuperscript{24} However, these bans were adopted primarily because polystyrene is non-biodegradable and cannot be efficiently recycled. Some cities recognize the health risks as further cause for the ban, though the health risks noted are due to inhalation of production fumes; the suspicion of adverse health effects via ingestion are unsubstantiated thus far.\textsuperscript{25}

\textbf{Migration to Food Products}

Studies have shown that polystyrene containers can leach styrene monomers and oligomers (extremely short polymer chains) into the food they store. The process by which this occurs is called migration and it is extremely complex. A comprehensive review on the topic concludes, “The migration process can be divided into 4 major steps: diffusion of chemical compounds through the polymers, desorption of the diffused molecules from the polymer surface, sorption of the compounds at the plastic–food interface, and desorption of the compounds in the food.”\textsuperscript{12} To explain, diffusion is the movement of molecules or compounds from an area of high concentration to low concentration. Desorption is the release of a substance from another substance, and sorption is the intake of a substance by another substance. In the case of polystyrene food containers, residual monomers are more abundant in the container than they are in food, so they will diffuse to the inner surface of the container that is in contact with food. The styrene is then released from the container to which it is absorbed by the polystyrene–food interface, which is the surface that divides two distinct phases of matter. The interface then releases the styrene into the food and the migration process is complete.\textsuperscript{12}

Several factors can affect the rate, quantity, and incidence of migration. Such factors include temperature, duration of food-container contact, concentration of residual monomers in packaging, and physical properties of the food. For instance, as temperatures rise, molecules move more quickly, so the diffusion process of styrene increases leading to more migration. Extreme fluctuations in temperature also yield more migration, as in the case of frozen meals which go from freezing to cooking temperatures within a few minutes, though this process is quite complex. Physical properties that can affect migration are the phase of the substance, fat content, water content, and other properties of the food. Food whose surface area is largely in contact with the container will leach more styrene than food whose surface area is in less contact with the container. Styrene is hydrophobic and insoluble in water, so it has been shown that styrene migrates more readily to foods with high fat concentrations and less readily to aqueous foods.\textsuperscript{12}

A few studies have been conducted to compare the migration rate between GPPS, HIPS, and EPS. These studies assume different procedures and report contrary data. Some studies do not include both rigid and foamed forms of polystyrene packaging in their comparisons.\textsuperscript{26,27,28,12} Expanded polystyrene foam seems to be a likely candidate for the most migration, as the individual polymer beads are not as close together as HIPS or GPPS. This may mean that the forces holding the beads together are not as strong and can be dismantled more readily when heated.
**Materials and Methods**

**Apparatus**

An Anasazi Eft-60 MHz pulsed Fourier transform NMR spectrometer was used for this experiment. It has a permanent Varian EM360 magnet and allows collection of 1-D and 2-D proton and carbon-13 spectra.

**Food Simulants**

Pure olive oil, whole milk, and 2% milk were chosen to simulate food products of varying fat contents. Milk and olive oil had been used commonly in related studies. To decrease potential migration from other plastic polymers, only milk in cardboard containers was used. However, the cardboard cartons are lined with a polystyrene coat. Plans to account for migration due to this lining were made. Half-gallons of Simple Truth Organic whole and 2% milk were purchased at the same time selecting for the same sell-by date. All samples came from the same carton of each type of milk as milk is a natural product and can vary from carton to carton. Pure olive oil could only be found in polyethylene terephthalate (PET) bottles. Market Pantry 100% olive oil was purchased in a 750mL PET bottle, and plans to account for this variable were established. All food simulants were purchased from the same store at the same time and transported to a refrigerator whose internal temperature was 2.9°C. Products remained in the refrigerator for one hour in their original containers before being utilized.

**Polystyrene and Control Containers**

Two types of polystyrene cups were chosen to serve as rigid and foamed polystyrene food containers. One type was an up & up™ brand, 12oz expanded polystyrene cup with a bottom diameter of 5cm, top diameter of 8.5cm, and a height of 11.2cm. The other was an up & up™ brand 9oz high impact polystyrene cup with a bottom diameter of 4.7cm, top diameter of 7.5cm, and a height of 9.5cm. The control containers were 8oz Ball™ brand, glass mason jars that were bisphenol A (BPA) free. These jars were also used to hold and transport the samples once they had been microwaved and left to cool in their polystyrene containers.

**Sample Preparation**

All glass jars were rinsed thoroughly with 100% acetone then air-dried. Samples for each food simulant would be prepared for the following testing variables: HIPS: microwaved, EPS: microwaved, HIPS: non-microwaved, EPS: non-microwaved, glass: microwaved, and glass: non-microwaved. (See Table 1) These variables were chosen to determine whether the type of polystyrene container and/or the heating process makes a difference in migration, and to verify that the microwaving process is not solely responsible for styrene monomers to occur in food as would be seen by microwaving food simulants in non-polystyrene glass containers.

Six samples of each food simulant were weighed out to 120 grams into appropriate containers and placed in the refrigerator at 2.9°C for a total of 48 hours. Whole milk and 2% milk microwave samples were microwaved for 4 minutes at 30% power in a 700 watt Sunbeam® microwave oven, and left to stand in the microwave for 3 minutes. They were then allowed to cool to room temperature outside of the microwave for about 130 minutes. Olive oil followed this same process, but at 10% power instead, however the olive oil food simulant had to be discarded from the study for reasons that will be discussed later. These criteria were calculated to achieve a temperature of about 74°C in equal amounts of time without boiling the simulants at
any point. The temperature was chosen based on the safe, minimum cooking temperatures for a variety of foods. The non-microwaved samples were transferred to glass jars at the same time to ensure all samples got the same amount of exposure time to polystyrene with the exception of the controls.

Results
Olive oil has been commonly used in styrene migration studies, however it completely dissolved both the high impact polystyrene and the expanded polystyrene cups that were microwaved once the temperature of the solution reached about 46°C. For this reason it could no longer be tested.

The H-NMR that was used was not sensitive enough for this study’s purposes and the noise to signal ratio was too high. In some cases there were incredibly small peaks in the same region that styrene would appear, however it was inconclusive whether the peaks were in fact styrene.

Discussion
Most studies of this kind have been done using a gas chromatograph mass spectrometer (GC-MS) or an ultraviolet-visible spectrophotometer (UV-Vis), however the equipment available to me did not include these. I decided proton nuclear magnetic resonance spectroscopy (H-NMR) could work, given that the benzene ring on styrene is expressed in a different range than the other hydrogen components of my solvents. The H-NMR that I used, however, was slightly outdated despite being research grade, and was not sensitive enough for my samples. Based on a few highly concentrated samples that were run through the H-NMR, I maintain that it is possible to measure styrene concentrations in a study such as this if the machine is more sensitive.

Surprisingly, the aqueous nature of the milk samples did not overwhelm the H-NMR which was a concern. This implies that aqueous solutions can be tested in an H-NMR. This information is useful as other lab equipment such as GC-MS cannot analyze samples containing water, making H-NMR a possible option for other experiments.

The olive oil likely dissolved the containers due to solubility properties of like substances. At a sufficiently high temperature it was able to react with the polystyrene polymers and dissolve them. After a further look at the methods of other studies it appears that those which used olive oil used high impact polystyrene cups, which are polymerized with rubber. I hypothesize that the HIPS containers they used were a different ratio of polystyrene to rubber, and thus required a much higher temperature to dissolve. This could also be why there were no studies that heated olive oil in expanded polystyrene containers as they are pure polystyrene.

Conclusion
Several studies show that styrene can migrate from polystyrene containers to food. Foods with higher fat contents are more susceptible to migration. It is possible that the type of polystyrene has an effect on migration, however studies that attempt to show this have contradicting results, and thus further testing is needed.

The health effects of styrene inhalation can be severe, however there is not sufficient data to support or deny major health issues from ingestion. More research should be conducted in this area to decide if typical styrene ingestion is safe.
Figures

Figure 1
Polystyrene Uses. CIEC. [Internet] [cited 2016 Apr 20]. Available from: http://www.essentialchemicalindustry.org/polymers/polyphenylethene.html

Figure 2

Figure 3
Styrene polymerization. University of South Carolina Upstate. [Internet] [cited 2016 Apr 20]. Available from: http://faculty.uscupstate.edu/llever/polym er resources/synthesis.

Figure 4
Figure 5

Molecular XPS vs. EPS. Extruded Polystyrene Foam Association. [Internet] [cited 2016 Apr 20]. Available from: http://ww3.owenscorning.com/content/docs/xpsa tech talk 07-02-2013 revised.pdf
References


13 Processing styrene polymers and copolymers by injecting molding [Internet]. 2001. BASF; [cited 2016 April 20]. Available from: http://www2.basf.us//PLASTICSWEB/displayanyfile?id=0901a5e180005b3e


Nuclear Magnetic Resonance Technology: An Overview

Alee Monaco

Physics 112
Professor Mike Swingler
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Abstract

Proton nuclear magnetic resonance (H NMR) spectroscopy is a commonly used method of analyzing compound structures containing hydrogen. It is a non-damaging method that allows the sample to be used after analysis, and thus it is appealing for many analytical purposes. It utilizes the basic nuclear principles of nuclear spin states to detect resonance after transmitting radio frequencies within a magnetic field. It can measure the amount of hydrogen types, the ratio of these types, and give some insight as to the chemical environment surrounding the different hydrogens. It has some limitations in that it is only relative data rather than absolute, and solvents cannot contain hydrogen, though the hydrogen ion, deuterium can be used. The advances of H NMR technologies have increased sensitivity and accuracy of analysis, as well as improved general usability for other substances.

Introduction

There are many ways to analyze chemical compounds and the various features of them. Different forms of lab equipment can measure variables such as mass, isotopes, light wavelength absorbed, unique constituents, and many other characteristics of organic compounds. These characteristics often aid in determining the exact compound in a sample, or exact properties of the compound in a sample. Analytical chemists use these tools every day to help analyze things such as water contaminants, drug quality, food quality and contaminants, and hundreds more.¹

One type of analytical technique is Nuclear Magnetic Resonance Spectroscopy (NMR Spectroscopy). It is a technique used to determine an organic compound’s exact chemical structure. Chemist William Reusch states, “Of all the spectroscopic methods, it is the only one for which a complete analysis and interpretation of the entire spectrum is normally expected.”² It is a non-damaging type of analysis, which means the sample is unharmed after testing, and can technically be used for other purposes once analyzed. This is dissimilar to other methods such as gas chromatography mass spectrometry (GC-MS), which separates all constituents of a sample in order to analyze the atomic masses of the components, resulting in a non-reusable sample.²

Physical Principles of NMR Technology

In order to effectively use and understand NMR Spectroscopy, one must understand the physical principles by which the technique is founded on. Every atom has a nucleus containing protons and neutrons. Protons are positively charged, while neutrons have no charge, so the nucleus has a net positive charge. Isotopes that have an odd number of protons and neutrons have a nuclear “spin” known as an angular moment (I). The spinning of the charged nucleus results in a magnetic moment (µ) which generates magnetic interactions with the particles and matter around it. Isotopes can have integral spins, fractional spins, or no spin based on the number of protons and neutrons they have. If the number of protons and neutrons are both odd the spin will be an integer (1, 2, 3…). If the number of protons is odd and the number of neutrons is even or vice versa the spin will be a half-integer (½, ¾…). If the number of protons and neutrons are both even, there will be no spin.² (See Figure 1.)
Nuclei with half-integer spins have a spherical charge distribution, so their behavior is easiest to interpret. Hydrogen’s spin is half-integer \( \frac{1}{2} \), so a vast majority of NMR’s are tuned to hydrogen nuclei, though other isotopes such as \(^{13}\text{C}, ^{19}\text{F}\) and \(^{31}\text{P}\) also have half-integer spins. When an NMR is tuned to hydrogen it is called a proton NMR (H-NMR). For the purposes of this paper, the following explanations will assume the use of an H-NMR.\(^2\)

When an external magnetic field \((B_x)\) is applied to the charged nuclei, the energy level of the nuclei are split and equal but opposite spin states occur as \(+\frac{1}{2}\) and \(-\frac{1}{2}\). Before a magnetic field is applied, nuclei spin randomly and unpredictably. After it is applied the nuclei align in a predictable fashion. (See Figure 2). The proton nuclei that line up with the magnetic field \((\alpha)\) have less energy and those that line up against the magnetic field \((\beta)\) have more energy. More nuclei line up along the magnetic field than oppose it, so more alpha nuclei spin states exist.\(^3\) As the field gets stronger the difference in energy \((\Delta E)\) gets larger. This can be expressed by the following equation: \(\Delta E = \frac{\mu B_x}{I}\) (See Figure 3.) This energy difference is important to what the NMR will ultimately measure which will be discussed later.\(^4\)

**The Parts of an NMR**

While there are different types of NMR’s, there are a few components that every NMR requires. Arguably one of the most important parts are the strong magnets that generate a homogenous magnetic field. They are often coil wires that allow a current to pass through. They are also typically superconductors, meaning at a sufficiently cool temperature, they have no resistance and can run forever, making the NMR quite efficient. Maintaining the superconducting qualities of the wire is somewhat expensive, however, as it requires a liquid helium bath to surround the wire. Liquid helium is costly and can boil off as it is only a liquid at extremely low temperatures (approximately -270°C), so it is generally surrounded by a heat shield that prevents the precious helium from going to waste. The heat shield is commonly liquid nitrogen, which is relatively cheap. This poses a problem to the samples, though, as such low temperatures can freeze many solutions, so a special compartment tube set for normal room temperatures passes through the magnet for the samples to be measured.\(^5\)

The variables measured by NMR are incredibly sensitive, so it is crucial that the magnetic field is completely homogenous. Therefore, “shim” coils are placed between the superconducting coils. Shim coils generate small magnetic fields that correct the imperfections of the superconductor’s field. There can be tens of shim coils in modern NMR spectrometers that effectively balance out the main magnetic field.\(^5\)

The tube that holds the sample in the magnetic field is called the “probe.” It has circuitry that sends radio frequencies to the sample and detects the response. Probe types can vary based on the nuclei they are tuned to. Some probes, called broadband probes, can be tuned to any nuclei. This is helpful for labs that measure compounds with different nuclei, however some accuracy is lost by having multiple options. Dual probes, on the other hand, can only be tuned to one type of nucleus. They are much more accurate and sensitive than probes that are tuned to multiple nuclei. Many labs prefer to use H NMR, and only need the equipment to be tuned to proton nuclei, so they use dual probes for better accuracy.\(^6\)
Another crucial NMR component is the radiofrequency transmitter. The transmitter is often controlled by a computer interface. The computer can dictate the specific frequency and allow a pulsing transmittance rather than a consistent flow. The pulse specifications can be adjusted for applied time and pulse intervals.5

The signals produced by the NMR must also be received and displayed for the user to interpret. The signals from NMR are rather small, so they need to be amplified to adequate levels. For that reason, pre-amplifiers are placed near the probe to detect the small signals and send them to the spectrometer console. The console can then express the data on the computer for the user to interpret.5

Most NMRs come with a specific software that displays the data in the form of a graph. Different software can be used, and is largely a matter of preference.5

To see an NMR diagram, see figure 4.

How an H NMR Works

An H NMR works by a series of processes that take advantage of nuclear physical principles. The steps are as follows: a sample is prepared and put into a thin, glass NMR tube. The tube is inserted into the NMR and spun rapidly by air jets to homogenize the solution and disperse the hydrogens equally. At this point the nuclei are spinning in random and non-uniform directions. The magnet is then turned on which attracts the nuclei to align in either the alpha or beta spin state. More nuclei will align in the alpha spin state as they are in the naturally advantageous lower energy level.4

Once the nuclei are aligned, a fixed radio frequency is transmitted to the sample. The appropriate radio frequency with the appropriate magnetic field will pump the nuclei in the lower alpha energy state into the higher beta energy state. The radio frequency radiation energy must equal $\Delta E$ in order to do this. Not much radiation is needed as the energy difference between opposite spin states is relatively low. Radio frequencies have little energy, which is why they are used for NMR and also why NMR spectroscopy is such a non-invasive process.2

It is not the excitement of the nuclei that is measured, however. As the nuclei are allowed to relax back to their original alpha spin state, they emit electromagnetic signals by creating a fluctuating magnetic field. The process of the proton nuclei switching between the alpha and beta spin states is called resonance. This magnetic field generates a current that is measured by a receiver coil that surrounds the sample. This electromagnetic signal generates the results for the $y$-axis of the NMR data, which is the height of the peaks in terms of the signal frequency. The area under the peak also shows a relative value of the amount of chemically equivalent hydrogens.2

The height of the vertical peak is relative to the absolute quantity of hydrogens in a sample and is directly proportional. This means that diluting a sample to one thirds its original concentration would result in a peak one third of the original. This is important when analyzing equal molar concentrations of two different compounds such as benzene and cyclohexane, as benzene’s vertical peak will be half as intense as the peak for cyclohexane due to it having half as many hydrogens.6
Interestingly, NMR has an atypical x-axis that reads from right to left instead of left to right. The x-axis of the data is known as “chemical shift” and can be represented by parts per million (ppm). This represents the strength of the magnetic field required to generate resonance in comparison to a predetermined standard. This is because the predetermined standard should require the strongest magnetic field in comparison to the sample. The standard, therefore, lies to the right of the x-axis, while the magnetic field required for the sample lies to the left, indicating how many parts per million of the standard’s required field is needed. For example, an article explaining how NMR technology works explains, “A peak at a chemical shift of, say, 2.0 means that the hydrogen atoms which caused that peak need a magnetic field two millionths less than the field needed by TMS to produce resonance.” (See Figure 6.)

Different chemical environments around the proton nuclei will require different magnetic fields to produce resonance. This is why peaks show up in different regions along the x-axis depending on their environment. It is possible for a sample to require a larger magnetic field to induce resonance, however it is uncommon.

The Standard for Comparison

NMR can show relative values of various qualities, but the values are not absolute. For this reason, a standard for comparison is required. For H NMR, the standard is nearly always tetramethylsilane (TMS). TMS is a compound that contains twelve hydrogens, one silicone, and four carbons. (See Figure 5.) This great number of hydrogens is important, but what makes it useful is that the hydrogens are in exactly the same chemical environment and are drastically shielded from the magnetic field. This means that only one peak will show up and that the field needs to be quite strong in order to reach the hydrogens and generate resonance. This makes it a great standard as a great majority of samples will need a lesser magnetic field in order to produce resonance.

Some software shows the TMS spike to the right of the x-axis at the zero point, while others do not include it. If it is shown, the TMS spike is ignored when assessing data.

What can H NMR measure?

An H NMR can measure a multitude of characteristics of a sample. It can measure how many different types of hydrogens exist in the sample. These “types” are the different parts of a compound that contain hydrogen and are separated by the orientation of other elements around them. The other elements cause a difference in the hydrogen’s electron density due to different electronegativities. Each separate “type” of hydrogen will appear in a different region and create a peak.

H NMR can also measure the relative number of the types of protons. This measurement is not an exact number, but a ratio between two types. This is particularly useful for determining what an unknown sample is, as it shows the relative variation of types of hydrogens. For instance CH₃CH₃ would show a 2:3 ratio of hydrogens rather than just a total of five hydrogens, providing more information about the compound’s structure. The ratio is measured by comparing the areas underneath each of the peaks.
Expanding on this, the NMR can show the electron environment of the different hydrogens. The electron density of the hydrogens is what determines what region the peak will show up in. The more density surrounding the hydrogens, the lower the region, thus the relationship between electron density and region is inverse. For example, a simple hydrocarbon such as pentane would appear in the region between 0 and 3 because the electron density is entirely on the hydrogen as carbon is not electronegative. A hydrocarbon with electronegative atoms such as oxygen or fluorine would appear in higher regions as they pull the electron density away from the hydrogens.  

NMR can also detect how many hydrogen “neighbors” a hydrogen has. A hydrogen neighbor is a hydrogen that is paired closely to another hydrogen group that is expressing a signal. It is usually a carbon-hydrogen group, but there can be others. The hydrogens, which are tiny magnets, can add or subtract to the signal of the actual hydrogen group being measured. This results in hydrogen neighbors expressing an extra peak in a similar region as the main signaling hydrogen. There are often multiple extra peaks in these instances that make the total peak appear symmetrical.  

These various data are incredibly useful for deciding what unknown compounds are in a sample and the structure of these compounds. The information obtained from this can be applied in countless, useful ways.  

Solution Solvents for NMR Analysis

Often times, samples will be prepared by creating a solution of the substance that will be analyzed with a known solvent. It is important that the solvent does not interfere with the substance in question. For this reason, solvents that do not contain hydrogen are most appealing. Examples of these include Carbon Tetrachloride (CCl₄), Carbon Disulphide (CS₂), and Trichlorofluoromethane (CFCl₃). The absence of hydrogen means these solvents will not generate peaks on the NMR graph.  

It is also possible to use solvents that contain hydrogen if the hydrogens are replaced with their isotope, deuterium. Such examples include Chloroform-d (CDCl₃), Acetic Acid-d₄ (CD₃COOD), and Acetone-d₆ (CD₃COCD₃). While deuterium does have a magnetic field, it is not strong enough to interfere with the NMR and thus solvents containing deuterium are commonly used.  

Other Types and Advances in NMR Technology

Adjustments to NMR technology have been made in order to improve the usefulness, accuracy, and sensitivity of analysis. An article speaking to the advances in NMR technology from 2009 states, “The availability of high-field magnets, cryogenically cooled probes, and probably in the near future hyperpolarization techniques, the intrinsic NMR sensitivity has increased by at least one order of magnitude.”  

Additionally, solid-state NMR spectroscopy has been developed to analyze solid macromolecules. This has been useful for drug companies and various fields in which the solid substance cannot or should not be dissolved.
NMR has also been recently coupled with chromatography methods which allows more quantitative data to be generated. In particular it has been useful for biochemical analysis of complex body fluids and drug design.\textsuperscript{10}

**Conclusion**

NMR technology is quite useful for determining the characteristics of substances. It allows us to determine unknown substances, understand how compounds are structured, and make inferences on the chemical environment of atoms. Its gentle nature on samples allows it to be an easy method, as there is no worry of destroying a small sample if that is a concern.

What is most interesting is that the technology relies on basic, physical properties that are not complex. The brilliance behind NMR technology seems to be the simplicity of it all. For instance, acknowledging the opposing spin states of nuclei in order to further decide that energizing then relaxing them will omit signals that can glean such valuable information. Additionally, the use of magnets and radio waves seems almost comical because they are so weak and less destructive than many other forms of analysis, yet they are sufficient for these purposes.

I have personally gained a respect for the simplicity of this science, and the realization that such helpful tools are often the simplest ones.
### Figures

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**Figure 1.**

Nuclear spin quantum numbers. [Internet]. UC Davis. [cited 2016 Apr 21]. Available from: http://chemwiki.ucdavis.edu/Core/Physical_Chemistry/Spectroscopy/Magnetic_Resonance_Spectroscopies/Nuclear_Magnetic_Resonance/NMR%3A_Theory

![Random orientation outside of field](image)

**Figure 2.**

Figure 3.
Spin state energy differences. [Internet]. William Reusch. [cited 2016 Apr 21]. Available from: https://www2.chemistry.msu.edu/faculty/reusch/virttxtjml/spectrpy/nmr/nmr1.htm

Figure 4.
NMR diagram. [Internet]. William Reusch. [cited 2016 Apr 21]. Available from: https://www2.chemistry.msu.edu/faculty/reusch/virttxtjml/spectrpy/nmr/nmr1.htm
Figure 5.

Figure 6.
References


Abstract: This outline examines the power and beauty of thunderstorms. From weak to strong and powerful, these storms are filled with science and physics properties that are just waiting to be uncovered. This essay includes a description of the clouds that form to create powerful storms. The physics behind lightning including; polarization and static electricity. The development of the supercell and tornadic activity including; centripetal force and angular momentum. A depiction of weather technology including the science of Doppler radar. Finally, a look into different safety precautions including; lightning rods and general safety measures that can be utilized to prevent injury and loss of life.

Introduction

Thunderstorms are considered the epitome of severe weather phenomenon on earth. Storms can come in many different forms and range in exhibition and severity. From lightning and rainfall to massive hail and tornadic activity, thunderstorms are machines filled with massive amounts of power. Throughout this paper, an in-depth look will take place into the physics behind these storms. The facets examined will be; the science behind cloud development, lightning, and the development of the thunderstorm itself including; the formation of the ever-fearing supercell. Additionally, current weather technology will be examined along with some various protection measures that can be applied. By understanding the science and numerous safety practices behind these storms, better preventative measures can be established to impede severe structural damages as well as injury.

Foundation: Clouds

Clouds play a vital role in the advancement of the thunderstorm. Essentially, clouds lay the foundation. Clouds can come in all shapes and sizes and range in possible danger. To begin, it is important to understand what a cloud is. As basic as it sounds, clouds are actually filled with lots of scientific involvement. Clouds are comprised of thousands of tiny water droplets that are incredibly light and able to suspend in mid-air. Clouds form due to water vapor condensing into tiny water droplets usually compromised of a radius of 10 micrometers (University of Alaska Fairbanks, n.d). When the atmosphere is below 100% humidity, water is consistently evaporating and condensing into and out of the air. When the atmosphere reaches 100% humidity, it is said that the atmosphere has reached the maximum amount of moisture it can hold (University of Alaska Fairbanks, n.d). At this point, water vapor can only condense back into the tiny, liquid particle form; thus the formation of the cloud. A huge component that aids in the formation of a cloud would be the process of air rising. The two main processes that occur to facilitate this movement would be convection and forced ascent.

Convection and Forced Ascent:

Convection is characterized by the vertical circulation of air that occurs by means of the air getting heater below the ground or at the oceanic level via the sun (University of Alaska
Air particles, that are warmer than the surrounding air, expand and lighten and therefore rise. As these air particles rise, the energy that was used to expand the molecules is lost and causes the air to cool. At some point in time, the air will reach a cooling point by which the humidity level is 100%. This particular temperature is known as the dew point temperature (University of Alaska Fairbanks, n.d). This particular phase change from vapor into liquid releases more heat into the air. The cycle then begins again. This constant loop is what allows for a cloud to grow and stack upon itself exponentially (See Figure 1).

The second way in which cloud formation occurs is by means of forced ascent whereby a warm front moves closer and closer toward a cold front. As these two fronts approach each other, a point is reached where less dense warm air is forced upward in order to gradually slip over the cold front (University of Alaska Fairbanks, n.d). The amount of energy released in these clouds as the growth continues can become astronomical. Clouds used in the formation of thunderstorms have been said to be able to reach energy levels higher than an atomic bomb (University of Alaska Fairbanks, n.d) (See Figure 2). Clouds are the building blocks of thunderstorms. The next component of examination is going to focus on lightning.

**Lightning**

Lightning is one of the most common grounds of electrostatic electricity in nature. The clouds that were discussed above can reach a point whereby they get filled with loads of electrostatic energy (University of Alaska Fairbanks, n.d). The main foundation of such activity is that of polarization. In other words, the high points of the clouds are said to acquire more positive charge. The lower points of the clouds are said to acquire more negative charge. There are two different mechanisms by which lightning is formed within a storm cloud.

The first mechanism involves the millions of small, suspended water droplets. As water from the ground evaporates into the air, these particles stick together and form clusters. As these clusters move upward, a collision occurs between the clusters and the suspended water droplets causing electrons to get ripped off the ascending clusters of droplets. This causes a separation of positive and negative charge within the cloud. The electrons remain toward the bottom of the cloud as the positive charge of the ascending molecule moves upward within the cloud (The Physics Classroom, 2016). The second mechanism by which this polarization occurs would be the process of freezing. As the altitude increases, the rising moisture temperature gets cooler or decreases. The ascending water droplet clusters proceed to freeze. These frozen clusters tend to stick together forming central regions of the droplet clusters. The frozen section of these clusters of particles becomes more negatively charge while the outer droplets acquire a more positive charge. Air currents can rip the outer particles upward causing the frozen sections to drift downward therefore contributing to polarization (The Physics Classroom, 2016). This polarization has a grand effect on the surface of the earth.

**Lightning Strikes:**

Clouds produce an electric field on the earth. The electrons present on the earth’s surface are repelled by the electrons present at the bottom of the storm cloud. An opposite charge is formed between the bottom of the cloud and objects on the earth’s surface like; trees, buildings and care. These objects all receive a positive charge. A buildup of static charge within the storm cloud makes the electric field stronger. This electric field becomes ionized and therefore, more
conductive. The electrons from the outer shells of gas molecules are torn off from this ionizing movement. The air now becomes filled with free electrons and positive charge. This phenomenon opens a path that allows a charge transfer or lightning bolt, to make it from the cloud to the ground (The Physics Classroom, 2016).

For a lighting strike to occur, a step leader needs to form (See Figure 3). This is characterized by the movement of electrons at the bottom of the cloud through the conducting air at 60 miles per second (The Physics Classroom, 2016). The electrons make way toward the ground moving in zig zag formations in many different directions. This acts as the roadway for which a lightning strike will make way toward the ground. As this process continues to grow stronger, the amount of positive charge on the people, building and objects on earth will in turn grow stronger. These positive charges continue to grow in an upward fashion known as a streamer (See Figure 4). When the streamer and stair leader meet, there is a mapped out, conductive pathway and lightning begins. This lightning bolt can strike, moving at 50,000 miles per second. Upwards between a million or billion electrons can travel across this pathway in a millisecond (The Physics Classroom, 2016). Various discharges occur between the negative and positive charges that cause an electric strike (The Physics Classroom, 2016). This causes a rapid and violent expansion of air that eventually creates a large soundwave known as thunder.

**Supercells and Tornadic Activity**

By the far the most dangerous form of thunderstorm is the supercell. Supercells are incredible forces of nature that have the capability to produce one of the most dangerous weather phenomenon’s; tornadoes. The formation of the supercell is driven by a vertical wind shear. A vertical wind shear is described as a downdraft pushing away from an updraft. If enough rotation and wind shear is available, there is a great possibility of a mesocyclone forming. Tornadoes are formed by rapidly rotating supercells. Tornadoes rely on angular momentum and centripetal force to be able to continue growing in size, power and strength.

**Angular Momentum and Centripetal Force:**

Tornadoes have angular momentum as they move along the ground attached to the supercell. Angular momentum can be described as the total amount of kinetic energy in rotation. By pulling mass in towards the axis of rotation, the rate of the spin will increase because the angular momentum is being conserved (Indiana University Bloomington, 2016). An example of this can be best described as a figure skater pulling arms inward to increase the speed of the spin.

As an object moves along a curved path, there is a tangent force that tries to pull this object away from the curve. This is known as momentum (The Physics of Tornadoes and Hurricanes, n.d) (See Figure 6)

\[ L \text{ (Momentum)} = \text{mass} \times \text{velocity} \]

In order for these power houses to continue spinning, a force must act against this momentum trying to straighten and slow the path of the tornado activity. A centripetal force is the force that keeps an object moving in this circular direction (Hyper Physics, 2016). A tension is required within the tornado to keep the strength and speed moving. The center core of the
tornado creates its own tension that allow for the tornado to keep moving in a circular motion. If
the radius is to decrease, the velocity is the increase. These two are inversely proportional to one
another (Hyper Physics, 2016) since mass usually remains at a constant within a tornado (See Figure 5).

\[ F = \frac{mv^2}{r} \]

Storm Tracking Technology: Doppler Radar

Storm chasers rely heavily on Doppler radar for guidance into and out of storms. Meteorologists also utilize this information to warn citizens of severe weather via television. The science behind these instruments is quite fascinating. RADAR stands for radio detecting and ranging. This technology is rooted in radio waves. Radar sends out electromagnetic waves that can detect precipitation. When these electromagnetic waves collide with precipitation, parts of the energy scatter back into the radar (Australian Government: Bureau of Meteorology, 2016). The best way to understand this is imaging shouting into a canyon. You scream something at the top of your lungs and hear those same words as feedback. The Doppler radar works that same way and is able to track precipitation location and amount.

Components:

Most standard, basic radars have four different components. These components include; a transmitter, a transmit and receive switch, antenna and a receiver. The transmitter creates an energy pulse. The transmit and receive switch tells the antenna when to transmit and receive pulses. The antenna sends the signal out into the atmosphere and in turn receives the pulse or echo that comes back. The receiver transmits, receives and transforms the signal into a video or visual format that meteorologists can visualize on a screen. The output of the radar typically comes in two different formats; reflectivity and velocity. Reflectivity is defined as a specific measure of how much precipitation is occurring within a specific area. Velocity is comprised of both speed and direction of precipitation that comes toward or away from the radar technology. Doppler radars can measure both of these (Australian Government: Bureau of Meteorology, 2016). This is unique in that most radars can only measure the direction of the precipitation, not the speed at which it moves. (See Figure 6).

Reflectivity and Doppler:

Reflectivity stems from wave theory. This theory states that invisible electromagnetic waves radiate in certain electrical circuits with the speed of light. The speed of light is dependent upon the frequency and the wavelength. Likewise, these waves reflect in a similar way (Australian Government: Bureau of Meteorology, 2016).

\[ C = \lambda f \]

Doppler affect is the effect of sound waves changing in pitch when there are shifts in the frequency (See Figure 7). Doppler can be used to calculate how fast an object is moving based on the sound frequency. In weather terms, Doppler is used to determine the speed of the
precipitation in the atmosphere as the precipitation falls in conjunction with the wind speeds. Through this technology, meteorologists are able to calculate wind speeds and determine the magnitude of danger associated with the storms (Australian Government: Bureau of Meteorology, 2016).

Radar Images:

The images that are presented on the display screens are very useful for storm trackers because they are displayed in a map format with precipitation intensities. These intensities are color coded. Lighter, cooler colors are used to represent light intensities (i.e. light blue or white). Darker, warmer colors are used to represent heavier intensities (i.e. dark red or burgundy). These intensities are measured by the size of the particles that are transmitted by pulses received by the radar. It is from the size of the different particles that the state of the precipitation can be determined (i.e. rain or hail) (Australian Government: Bureau of Meteorology, 2016). These images are absolutely essential for storm chasers to make it through storms when the weather is acting against them. Storm chasers rely quite heavily on this technology especially when tornados are considered rain–wrapped. In other words, the tornado is hidden under an entire curtain of rain. Visibility to the naked eye is difficult. However with improved technology such as radars, storm chasers can chase storms safely and efficiently to be able to warn citizens to take cover.

Errors in Technology:

As with all technology, errors can occur. At the same time, storms can be highly unpredictable and can change course at the last minute. Radar cannot detect every last thing that is going on in the atmosphere. Likewise, images that come up on radar do not necessarily represent precipitation. If radar is in close range to the coast, it may pick up some of the sea as precipitation (Australian Government: Bureau of Meteorology, 2016).

Another dangerous concept for chasers would be mountains. Not only do mountains reduce visibility of the storm to the naked eye, but the mountains can also interrupt the radar signal and the pulse or echo cannot make it back to the receiver. Objects can be picked up by radar as well. Doppler radar is more likely to pick up these signals because it is highly sensitive (Australian Government: Bureau of Meteorology, 2016). Radar beams broaden as the distance increases between the radar and the target (See Figure 8). Sometimes the entire rain shaft cannot be detected, only the upper sections of it (Australian Government: Bureau of Meteorology, 2016).

Protection – Preventing Injury and Saving Lives

These storms can be incredibly dangerous and contain a lot of power. It is important to install protective equipment and understand basic safety precautions in the event of a severe weather outbreak. There are several protection measures that exist to help counteract the extreme power of these storms. The first protective measure to touch on would be lightning rods. There are many buildings that are equipped with lightning rods to help improve safety. The two safety theories that prevent violent lightning strikes are known as lightning dissipation and lightning dispersion.
Lightning Dissipation:

The principle behind the safety measure, proposed by Benjamin Franklin, states that the electric field around a pointed object is high. This pointed object or rod can be used to ionize the surrounding air and increase the conductive ability. This theory states that as a storm cloud approaches there is a conductive pathway that forms between the lightning rod and the storm cloud that is polarized. The static charge from the cloud is said to slowly move down this pathway toward the ground. The hope is that this slow movement prevents a violent discharge from occurring (The Physics Classroom, 2016).

Lightning Dispersion:

This theory states that there is a conductive pathway of the charge to the earth. Lightning rods are likely attached to a large, thick, copper cable that is buried under ground. The purpose of this mechanism is to maneuver the charge from the static cloud, through the pole and into the ground to counteract the violence of the discharge. This theory is the one measure that is typically used today in building (The Physics Classroom, 2016). Although these two methods provide some means to prevent the intensity of a lightning strike, there is no exact science that has been discovered to prove an actual prevention measure of a lightning strike (See Figure 9).

Faraday Cages:

The car is an example of a faraday cage and can protect one should they find themselves in a lightning storm outside. Michael Faraday invented these cages by creating a room coated in metal foil. Faraday proposed that charge only resided on the outside of a conductor. Exterior charge had no influence on anything enclosed within the conductor. The reason behind this would be that charges redistribute in a way that the interior fields due to them cancel one another out. During a lightning storm staying within your car is the smartest and safest thing to do (Rubin, 2013) (See Figure 10).

General Safety

There are also some basic guidelines of safety when it comes to storms as well such as remaining indoors and staying away from class windows during large wind gusts and tornado warnings. Children in schools are encouraged to hide within the hallways of the school during a tornado warning. At the same time, citizens are pleaded to hide in basements, bathtubs or in ditches if one finds themselves outside during the storm. Another really important safety measure would be to listen carefully to weather sirens that warn that severe weather is close. These small measures can be taken in order to help protect people during these dangerous storms.

Conclusion

Throughout this essay, a comprehensive look was taken into the physics behind thunderstorms. There are many components that factor into the power of these storms. A look was taken at the foundation of storms; the cloud. A deeper understanding of lightning and the formation of supercells as well as subsequent tornadoes was given. Another concept looked into
was weather technology and how the advancements are used to predict thunderstorm activity. Lastly a look into safety measures was taken. By better understanding the science behind these powerful storms, more improved safety precautions can be made as well as tools and technology to help make better predictions in regards to severe weather. At the same time, earlier weather alert systems can be made to provide quicker warning systems for citizens.
Figures

Figure 1

Figure 2

Figure 3

Figure 4

Figure 5

Figure 6
The effects of the curvature of the earth on weather radar.
References


A Broad Spectrum of Photovoltaics

Nicholas Nabours

Paradise Valley Community College
PHY 112 #12714
Professor Michael Swingler
April 21, 2016
Abstract:

The constant input of energy received by the sun often goes unnoticed in today’s society except for the daylight, which without would make society fall apart. Humans have developed a way to mimic plants ability to absorb the suns electromagnetic rays that wash over the earth to produce energy in the form of solar panels. Harnessing these rays, solar panels or photovoltaic cells have the ability to take photons produced by the sun and create an electrical charge to supplement the power of today’s electronics. The structure of these panels, most often made of doped silicon, along with their efficiency and ability to only absorb certain wavelengths of light will be discussed. Along with this, a few of the many different forms solar panels or photovoltaic cells come in will be examined. Finally, we will take a look at the future of solar technologies that are being researched.

To understand how solar panels or photovoltaic cells work the physical makeup of a basic commonplace solar panel can be used. Common solar cells consist of a p/n junction, where a p-doped semiconductor and a n-doped semiconductor meet, and electrodes connecting the two opposite sides of these layers with an electrical load in between. An n-doped semiconductor, like phosphorous-doped silicon, and a p-doped semiconductor, like boron-doped silicon, would be great examples for the two layers. N-doped semiconductors contain atoms with excess or five electrons in their valence shell, allowing one unbounded electron to jump when excited, while p-doped semiconductors contain atoms with only three valence electrons or missing electrons creating holes where the missing electron should be. Where these two layers meet they form the form an n/p junction. These opposite charges neutralize each other only at the n/p junction that act as a wall for negative electrons seeking positive holes. This divide is crucial to the formation of charge when photons strike the solar cell and ionize electrons. Electrons will follow the path of least resistance and while it may seem that we are creating this energy, which is not possible due to the second law of thermodynamics, it is just being converted from electromagnetic energy to electrical energy. When electrons are ionized by photons as seen in Figure 1, in the n-doped silicon they are forced to travel through an external circuit due to there being less resistance than trying to pass through the n/p junction in the center. This is where the current is generated. The voltage comes from the force provided by the difference in charge at the n/p junction and with current flowing, a direct current circuit is formed. This circuit is possible only due to the photons exciting electrons in the n-doped silicon, but due to the many different wavelengths or energies of photons not all of these are able to be absorbed.

Solar cells, being reliant on the sun for their energy, would need to be able to absorb photons in wavelengths between approximately 1400nm and 250nm and these are inversely proportional meaning more energy for shorter wavelengths. Using these wavelengths and the equation in Figure 2, the energy of visible light spectrum can be calculated to be from 0.88 electron volts (eV) to 4.95eV. Silicon has a band gap of 1.1eV, which means it absorbs 1.1eV or greater, photons that strike its surface. This means that any photons outside of silicon’s bad gap cannot be absorbed, pass right through, reflect off, or are converted to heat. Knowing this information, solar cells with larger band gaps or absorb higher power (eV) photons and can be used to absorb different wavelengths of light which is now being done with multijunction photovoltaic cells, these will be examined later on. The reduced ability of photovoltaic cells to absorb photons leads to reduced efficiencies, the major shortcoming of solar panels to date. The scientific community has been striving to increase the efficiency of solar cells since their creation. Currently the highest efficiency rates for commercially available solar panels range from 15-20% efficient, which means these solar panels are also more space efficient needed
smaller areas to produce their power. The efficiencies are determined under AM1.5 conditions shown in Figure 3. The total efficiency under AM1.5 conditions is then calculated using the formula in Figure 4 where Eta represents the efficiency percentage and the Fill Factor is determined by differentiating the power of a photoelectric cell with respect to voltage. Increasing the voltage output of the cell or increasing the current produced would both greatly increase efficiency of the cell as a whole, but this is much easier said than done. Panels that are only 15% efficient would only produce 150 watts for every 1000 watts of photons they receive. According to PVEducation.org the total solar radiance can reach a, “maximum of about 1 kW/m²” each day, or 1000 watts/m², meaning that most of the time solar panels will end up producing much less power than they receive. The amount of power is also dependent on the position of the sun in the sky as well. When the photons from the sun strike the panels surface at a 90° angle, they are able to transmit a greater amount of their power to doing the work of freeing up electrons to create an electrical current. In Figure 5 we can see the relation of tilting a panel to create the perpendicular surface the photons need to produce the most energy. The intensity of sunlight would be equal to the \( \cos(\alpha+\beta) \) where when equal to 90° the light will strike with the most force, and when at 45° the light would strike at approximately 70% its original intensity. Achieving the highest possible value for the angle \( \alpha+\beta \) will achieve the most absorption of the insolation and produce the most energy. Solar arrays made up of solar cells that can track the sun at this angle were made to maximize the power output by maintaining this angle throughout the day. Many solar farms today, which are made up of hundreds to thousands of arrays, will use these sun tracking arrays that track the sun according to the time of day, and month of the year. Compared to a 2-gigawatt coal or nuclear power plant, solar farms appear modest only being able to produce around 1% as much power. When compared this way solar cells seem rather pointless, but they allow power production to be shifted to the point of use at a person’s home, thus removing the loss of power from converting electricity and sending it long distances. This gives solar cells a small edge so they aren’t entirely counted out.

Next, a few of the many different types of photovoltaic cells can be surveyed. Starting with crystalline silicon, two different types are available, polycrystalline and monocrystalline silicon solar cells. Polycrystalline silicon solar cells, first introduced in 1981, are “Raw silicon [that] is melted and poured into a square mold, which is cooled and cut into perfectly square wafers.” These cells will appear speckled blue coloring due to impurities in the silicon. Being very similar to the production of many older computer components, this mode of manufacture process is both simple and more cost effective in producing solar panels, reducing their price per watt produced. Polycrystalline solar panels are less efficient though, due to reduced silicon purity, averaging 13-16% efficiency, and are less tolerant of heat that reduces their ability to produce a constant voltage. Using the low end of their efficiencies, if a polycrystalline solar panel received 1000 watts/m² of energy from the sun at a maximum, it would only produce 130 watts/m² at a maximum each day. This reduced efficiency translates to an increased panel size so that enough power will be produced by the solar panels to supply their electrical load. Monocrystalline silicon solar cells, on the other hand, have a much higher grade of purity, and will often appear with a uniform external even coloring. These Monocrystalline solar cells are made from silicon ingots that are cylindrical in shape, much how modern computer transistors electronics are made from, and have 4 edges cut off to give them a more square shape and improve performance. Due to their high purity the crystal lattice transfers electrons through the crystal more easily when excited by photons, allowing for more power to be produced by the solar cell. Monocrystalline solar panels usually have efficiency rating of 15-20% which is a
significant jump from the Polycrystalline solar panels. Using the same 1000 watts/m² maximum insolation as with Polycrystalline panels, and their minimum efficiency, Monocrystalline cells would produce roughly 150 watts/m². This is at minimum 20 watts more than Polycrystalline solar panels. Monocrystalline solar panels are also available to purchase commercially at their 20% maximum efficiency making them a viable option for those who wish to go ‘off the grid’. This increased efficiency will allow you to produce more power using less space than Polycrystalline solar panels. Along with this increased efficiency Monocrystalline solar panels perform better in low light conditions and have a roughly 40% longer lifespan, only losing 0.36% output each year while polycrystalline solar panels will lose 0.64% each year. The drawbacks to Monocrystalline solar panels would include their higher cost from a more expensive manufacturing process. Silicon of higher purities is not an easy product to make, and when you have to cut away the 4 edges of the silicon crystal cylinder to increase the surface area of the solar cell itself to produce more wattage, this cut away silicon would only end up as waste to be sold off or disposed of. Following these there are thin film solar cells. Thin film solar cells are relatively easy to make, consisting of “depositing one or several thin layers of photovoltaic material onto a substrate” and can be mass-produced relatively cheaply by these means. Common material used for thin film solar panels cans consist of amorphous silicon, cadmium telluride, copper indium gallium selenide(CIS/CIGS), and organic or organometallic photovoltaic cells. These solar panels usually have and efficiency rating of 9% for production models. Due to this low efficiency, thin film solar panels only are effective when applied on a large scale where land is plentiful. Their large footprint is just not suitable for residential application. These thin film solar panels also degrade more rapidly losing almost 1% of output each year. Despite these drawbacks, thin film solar panels come with some special properties that can increase their spectrum for applications. The amorphous silicon, the CIGS, and organic/organometallic solar panels have the ability to be flexible. Their flexibility allows them to be placed on curved surfaces, like the contour of a building or window or car, not limiting them to the standard flat planar format of the hard crystalline silicon solar panels. The amorphous silicon has been used in pocket calculators for years, and requires much less silicon than other manufacturing processes while also being lighter. “The cadmium telluride solar panels have surpassed the cost efficiency of crystalline silicon,” according to Energyinformative.org for large-scale solar farms. These systems would have to have solar arrays that track the sun to increase the amount of power generated. All of these previous solar cells and panels have been single junction photovoltaic cells that efficiently absorb a small bandwidth of the electromagnetic spectrum. Multijunction solar cells are multiple layers of photovoltaic material that respond to different frequencies or eV values of photons. The higher the eV value of a photon the more energy it has to transfer, but with lower eV values photons have longer wavelengths and may just pass right through a material and not ionize any electrons to produce an electrical current. The stacking of layers of photovoltaic cells that have the ability to stop high frequency or higher eV energy photons at the surface and allows you to absorb photons with longer wavelengths on each subsequent and lower level causing the lower energy photons to be absorbed the further the light passes into the photovoltaic cells. In Figure 6 we can see an example of a three-junction solar cell cross-section. Photons with a 1.93 eV or higher would absorb into the first layer, the second layer would absorb photons less than 1.93 eV and greater than 1.39 eV, while the bottom layer would absorb anything less than 1.39 eV and greater than 0.94 eV. This would allow you to absorb a much broader band of the total electromagnetic spectrum, making for a more efficient, and higher power producing solar cell. This is easier said
than done though, and as Phys.org notes researchers working with multijunction solar cells have a lot to account for such as, “lattice constant, thickness, dielectric constant, electron affinity, band gap, effective conduction and valence band densities, electron and hole mobilities, the doping concentration of shallow acceptors and donors, the thermal velocity of electrons and holes, the alloy density, Auger recombination for electrons and holes, direct band-to-band recombination, and how many photons with a specific wavelength are absorbed and reflected by each layer based on its dielectric properties” just to name a few. These multijunction solar panels are currently available today but are not cheap and often they do not work alone, but Multijunction solar cells are perfect for concentrating photovoltaics. Concentrating photovoltaics involves altering a large area of the sun’s rays and focusing them onto the solar panel. This is achieved in several ways such as use of Fresnel lenses, parabolic mirrors, reflectors, and luminescent concentrators. Starting with Fresnel lenses, named after French physicist Augustin-Jean Fresnel, are “[comprised of] several sections with different angles, thus reducing weight and thickness in comparison to a standard lens.” This also allows lenses to focus closer and direct a larger area of sunlight. Fresnel lenses are often used in lighthouses to direct beams of light long distances for ships, so it makes perfect sense that they would be use to direct photons from the sun towards a solar panel. In Figure 7 we can see an example of a cross section of a circular or row shaped Fresnel concentrator. Fresnel lenses can be made circular in shape or a cylinder to create, either a focus point for the light or a row of concentrated light respectively. Multijunction solar cells can be used to receive the roughly 500 times concentrated light from circular Fresnel lenses and high efficiency silicon is more common for the rows of concentrated light from Fresnel lenses, but these concentrator add heat so a cooling system will be necessary. The increased concentration of light produces an astounding effect of increasing the efficiency of solar panels allowing a jump to above 30% or much more in laboratory studies. Next we can look at using parabolic mirrors to concentrate the sun’s rays onto a solar panel. In Figure 8, two of the properties of parabolic mirrors are being used to manipulate and concentrate the incoming light from the sun. Light from the sun will be traveling close enough to perfectly parallel from the sun by the time it reaches the earth. When this parallel traveling light strikes the surface of a parabolic mirror the light will reflect off in three dimensions towards a certain focal length equal to one half the radius of curvature of the mirror. Figure 8 shows this but the addition of a second smaller parabolic mirror with its focus placed at the same focus point as the larger mirror can be seen. This second smaller mirror has the light rays from the larger mirror passing though its focus point and striking the mirror. When light rays originate at the focus of a parabolic mirror the travel outward in all directions to strike the mirror and will reflect off the mirror parallel. This is what allows the second mirror to redirect the light rays from a larger surface area onto a smaller point opposite the mirror, thus concentrating the sun’s rays up to 500 time normal. Typically, parabolic mirror concentrator require arrays to track the sun in two dimension requiring more expensive stands and computer to control this tracking. These types of arrays also will generate a significant amount of heat on the solar panel itself, due to solar concentration, and require a method to cool the panel. Multijunction solar cells would also be optimal for these situations, due to the high degree of solar concentration making good use of their wide eV band. A more simplified version of this would be reflectors. Reflectors use mirrors do direct sunlight so it strikes the solar cell and concentrate a small amount of sunlight. As seen in Figure 9, light is reflected at the same angle to the normal as which it strikes the surface allowing light that would not normally strike the solar panel to be used to create energy. No cooling would be necessary for reflector due to their
low levels of concentration of typically 1.5-2.5 times the normal. Lastly we have luminescent concentrators. For luminescent concentrators, “light is refracted in a luminescent film, and then [is] channelled towards the photovoltaic material.”\(^1\) As seen in Figure 10, light from the sun passes through the film between the two pieces of glass and is bent past its critical angle. When light strikes a substance at an angle larger than its critical angle it reflects off the surface rather than passing through. After the light reflects it continues on to reach the solar cell so that even diffused light can be used to power the solar cell eliminating the need to track the sun causing it to concentrate about 3 times the normal.\(^1\) These cutting edge, solar technologies are viable options in this day and age of photovoltaic power production.

Looking to the future of photovoltaic cells a plethora of possibilities seem viable. Combining these technologies offers ways to improve performance and generate photovoltaic cells more efficient than though possible, but this will be a real test of human ingenuity. At the forefront currently would be multijunction solar cells. Their ability to absorb more and more of the electromagnetic spectrum emitted by the sun is increasing in laboratory test and the current record for a 3-junction solar cell is 43.5%.\(^6\) These cells have a much higher theoretical limiting efficiency when compared to crystalline silicon. The theoretical limiting efficiency of silicon-based cell is roughly 34%, while multijunction cells are almost 87%.\(^6\) This means there is a greater room for improvement than with the current silicon cells. Due to a much greater eV band for multijunction cells to absorb, they can ultimately produce much more power from the same light.

The main hurdles multijunction solar cells have to overcome would be their cost and their ability to evenly produce power in each junction. Right now the cost of multijunction prevents them from being made into large panels and often requires sun tracking arrays and concentrators that must be precisely aligned for them to achieve significant results, while simple polycrystalline panels can be purchased for a few hundred dollars online and set on a roof. Multijunction solar cells must also be matched to where each junction produces the same amount of power because, “if one subcell has less photocurrent it will limit the current generated by the entire device.”\(^6\) This matching process will require advances in semiconductor materials and the manufacturing process as well to provide a uniform crystalline structure for the highest efficiencies.\(^6\) Alternatively, the rise of organic/organometallic photovoltaics could offer a whole new spectrum for the materials used to make the junctions. Taking a lesson from plants, organic photovoltaic are, “carbon-rich compounds with a structure tailored to optimize a particular function, such as responsiveness to a particular range of visible light.”\(^7\) This would seem very similar to the chlorophylls that plants use to absorb photons and ultimately produce glucose for energy. There are many advantages to organic photovoltaics including, “rapid, ultra-low-cost manufacturing to extremely thin, lightweight, and flexible form factors” and organic photovoltaics have, “no restriction on the size and shape of OPV devices, and every conceivable shape and form can be envisioned, with only human synthetic capability as the limiting factor.”\(^7\)

Being able to take any shape would be useful for anything from airplanes to electric cars. The surface of an object could be coated to become a solar panel; it’s a novel idea. Building could have their sun facing exteriors become solar panels and supplement much of their energy consumption during the day. Organic photovoltaics could be the answer to the desired roll-to-roll manufacturing process that would allow cheap production, similar to that of printing a newspaper. We could also see nanotechnology playing a roll in increasing the efficiency of solar panels as a whole. Scientists in the UK have printed rows of aluminum cylinders, as seen in Figure 11, on the surface of solar panels and have increased their rate of solar absorption by trapping light rays in the n-doped layer where absorption takes place.\(^9\) The aluminum works by,
“[bending] and [scattering] light as it travels past them into the solar cells.”⁹ Because the light is bent along the path of the solar panels surface or into the solar panel itself it travels a longer path in the absorptive substrate giving it a higher probability of exciting an electron. If crystal lattices of solar panels could be printed with the same precision as these aluminum cylinders on a large scale, the crystal lattice could be flawless and provide the highest opportunity for achieving efficient panels. Many of the technologies of photovoltaic of the future will only continue to increase efficiency and reduce the total cost of solar to the point where it may be the only thing necessary for our power needs.

In conclusion, solar power is in it’s infancy still, but we can, and have been rapidly improving how solar panels function and produce power efficiently. Modern solar technology started out as crude photoelectric cells of silicon and has become a much more complex and viable option for producing power in our everyday lives. With each new generation of solar panels, their efficiency is improved allowing them to harness more power from the sun. Our ability to concentrate light, using parabolic mirrors and Fresnel lenses, to up to 500 times that panel or cell would normally absorb is just one of the many examples of how older technologies build and alter the new ones we create.

My prediction for the future would ultimately be continued increases in solar panel efficiency. I would also include that multijunction solar cell would replace the current simple silicon single junction cells that are most prevalent to date. Multijunction cells will most likely surpass any silicon solar cells in the future with higher efficiency ratings and lower costs. Organic photovoltaics or thin film photovoltaics will likely allow is to harness the wasted exteriors of buildings, roads, and even cars all for the purpose of producing electricity. Concentration photovoltaics will likely become our main means of producing electricity due to the ability to harness much larger areas of light for the purpose of energy creation. The possibilities can only be limited by our design.
FIGURES

Figure 1

Photons strike the surface of the n-doped silicon and transfer their ionized electrons through the electrical load.

Figure 2

\[ E(eV) = \frac{1.24}{\lambda(\mu m)} \]

electron-volt (eV)

The exact value of \( 1 \times 10^6(hc/q) \) is 1.2398

To convert wavelengths of light to electron volts the above equation is used. Electron volts are more commonly used than joules for light.
http://www.pveducation.org/equations/photon-energy-ev

Figure 3

The specified atmospheric conditions are:

a) the 1976 U.S. Standard Atmosphere \(^3\) with temperature, pressure, aerosol density (total aerosol loading), air density, molecular species density specified in 33 layers
b) an absolute air mass of 1.5 (solar zenith angle 48.19°)

\( c \) Angstrom turbidity (base c) at 500 nm of 0.084 \(^8\)

d) total column water vapor equivalent of 1.42 cm
c) total column ozone equivalent of 0.34 cm

f) surface spectral albedo (reflectivity) of Light Soil as documented in the Jet Propulsion Laboratory ASTER Spectral Reflectance Database (http://speclab.jpl.nasa.gov/)

Standard atmospheric conditions used to test the efficiency of solar panel.
http://rredc.nrel.gov/solar/spectra/am1.5/
Figure 4

\[ \eta = \frac{V_{oc} I_{sc} FF}{P_{in}} \]

where \( V_{oc} \) is the open-circuit voltage;
where \( I_{sc} \) is the short-circuit current; and
where \( FF \) is the fill factor
where \( \eta \) is the efficiency.

Formula used to calculate the efficiency of a solar panel.
http://www.pveducation.org/pvcdrom/solar-cell-operation/efficiency

Figure 5

Trigonometric relation of the sun's location to a solar panel's angle of tilt toward the sun. The goal is to have \( \cos(\alpha + \beta) = 1 \).
http://www.pveducation.org/sites/default/files/PVCDROM/Properties-of-Sunlight/Images/TILTARR.gif

Figure 6

Simple structural makeup of a multijunction solar panel that absorbs a broad eV band of 0.94 and up.
Figure 7

Cross-section of a Fresnel lens, that shows its ability to bend light to focus on a solar cell.
http://www.greenrhinoenergy.com/solar/technologies/images/Concentrating%20PV-01.jpg

Figure 8

Diagram of a parabolic mirror concentrator, showing the ability of mirrors to collect and concentrate large amounts of light onto a solar cell.

Figure 9

Simple cross-section of reflector concentrator for a solar cell. Small amounts of light are reflected onto the surface of the solar cell.
http://www.greenrhinoenergy.com/solar/technologies/images/Concentrating%20PV-03.jpg
This diagram shows a luminescent concentrator refracting light into the solar cell utilizing critical angles of the light and indices of refraction for the film to channel light toward the solar cell.


Figure 11

Rows of aluminum cylinders on the surface of solar panel scatter light into the absorptive surface of the solar panel.

References


You Are What You Protein: Elementary Physics of Proteins Interactions

Siavosh Naji-Talakar

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Dr. Casey Durandet
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Abstract

Proteins are complex molecules that provide vital and critical functions for cells to sustain and operate. Understanding the processes that dictate the creation and modification of proteins can lead to further advances in diagnosis of medical ailments, treatment of those, and better applications for the treatments. The field of biophysics studies any biological process with a predictable pattern and assigns mathematical models to them. Understanding the smaller patterns that govern the basic primary structures of proteins, to their complicated quaternary structures, allows for biophysicists to begin piecing together the puzzle needed to create full scale models that can predict more medically applicable functions of protein modeling. These applications may range from medical equipment such as better MRI machines to medical treatment such as vaccinations. Biophysics with an emphasis on protein physics is at the forefront of bridging the gap and merging our biology based medical universe with physics.

Discussion

Proteins are complex molecules assembled as a chain of amino acids in varying length which play critical roles in the systems of all living creatures. They are very functional in cells and are involved in cell structure as well as helping aid the regulation of larger living systems such as organs or tissues. Homo sapiens' genome contains coding for 20,000 to 25,000 different proteins made from 34 megabases out of the 2.85 billion nucleotides that make up the entirety of the genome; equating to about 1.2% of the total genome coding for proteins. However, the importance of protein interactions cannot be understated even if they seemingly take up small amount of space on the human genome. Proteins play vital roles in the human body ranging from but not limited to enzymes that catalyze or facilitate biochemical reactions such as pepsin in the stomach, act as antibodies to help the immune system protect against attacks, regulate gene structure or gene expression with histones, help in the movement of muscles in the form of myosin or actin, provide structural support in the form of collagen, coordinate body functions in the form of hormones such as insulin, or transporting oxygen through our bodies in the form of hemoglobin. Understanding how important proteins are for life allows us to learn more about the molecular biology that effects proteins, their structures and their formation. Protein interactions can be described and predicted in the emerging field called biophysics. Biophysics allows the comprehension of protein interactions as described with physics based equations and mathematical models; if there are principles that can describe a pattern then a detailed prediction can be made and tested.

Biophysicists study all of biology and placing a mathematical model on any pattern is fair game. Applying this science specifically to proteins has provided amazing lessons and advances for humanity. Modeling protein machines, even though they are a million times smaller than the machines we are used to, shows that they work on similar principles. Kinesin, a motor protein, is a remarkable cellular machine powered by adenosine triphosphate hydrolysis allowing it to “walk” along microtubules in a cell allowing it to supports many cellular functions such as cellular transportation to mitosis. Biophysics is driven based on what society needs, such as aiding in creating vaccines, describing metabolic disease like diabetes, or understanding growth diseases like cancer. From these driving forces technologies such as sonograms for diagnosing cancers, MRI, PET scans, and CAT scans have been developed. Understanding the molecular interactions of proteins on a microscale transcends into novel advances on a much larger scale to benefit all of society.
General Properties

Proteins have an enormous variety of functions and those functions are based on the high specificity for target molecules that they interact with. This relationship is referred to as a lock and key approach. The need for such specificity demands a rigid spatial three-dimensional structure of the given lock and key proteins. Due to the way proteins interact it is important to first understand the physics protein structure before being able to delve into the physics of protein function or their ability to self-organize. Discussing each and every aspect of protein physics is beyond the scope available for this report; as such the discussions will be topical as they pertain to important background information and critical aspects of protein physics such as structures and folding.

Proteins, polymers, are made of amino acids that are linked into varying lengths called peptide chains with each chain having a chemically regular backbone with various side chains; these bonds are referred to as peptide bonds. The human body has a pool of twenty amino acids which are positioned and encoded by genes determining the protein polymer. Proteins are divided into three groups based on their environmental conditions and general structure. Fibrous proteins form large water-deficient aggregates with a high hydrogen-bonded structure; they are highly regular and are maintained by interactions in between chains of proteins. Membrane proteins also are in water-deficient environments; the intramembrane portion is highly regular and hydrogen-bonded with their size restricted by the membrane thickness. Water-soluble globular proteins are not as regular; their structures are kept with interactions of the protein chain with itself. These classifications are not fully adhered to by all proteins because some can have a globular head portions with fibrous tails.

Understanding of the 3-Dimensional structure of proteins has been limited mostly to water-soluble proteins. Water-soluble proteins can be more easily separated as molecules to have their structures established with X-ray crystallography and by NMR. Most discussions and figures refer to water-soluble protein regularities for this reason. The covalent bonds holding a sequence of monomers create the primary structure of a protein. Secondary structures are α-helix and β-structure; α-helices are represented by helical ribbons and extended β-structural regions are represented by flat sheets sticking together. When the secondary structure is packed into a globule it is referred to as the tertiary structure; and when several of the protein chains are made into a very large globule they form the quaternary structure. Proteins are remarkable in the fact that they can form these structures independent of cell machinery. Their spatial structures are determined by the amino acid sequence only allowing the protein to self-organize based upon the interactions determined by the genetic code which created it. It is important to note that this mechanism is not fully functional for very large proteins due to aggregation and many larger proteins require help in folding in the form of chaperones, beyond the limits of this review.

Proteins have the innate ability to spontaneously fold, given that they were not subject to strong post-translational modification processes. Post-translational processes take place after a protein has been synthesized by ribosomes and has undergone initial folding. An example of a protein losing the ability to self fold is insulin which ends up losing half of its amino acid chain after it has been folded in vivo and thus cannot refold. Proteins undergo a number of different post-translational modifications such as cleavage of the chain or proteolysis, acetylation, glycosylation, lipid binding, phosphorylation, and even splicing.
Two critical topics are the formation of S-S disulfide bonds between two cysteine amino acids and the cyclic structure of proline. Disulfide bonds are particularly strong because the side chain of cysteine contains a reactive sulfhydryl group (-SH)\textsuperscript{9,\textit{h}}. On release of a proton (H\textsuperscript{+}), a sulfhydryl is converted into a thiolate anion (S\textsuperscript{-}). Each of the two adjacent sulfhydryl groups are oxidized to form a covalent disulfide bond (-S-S-)\textsuperscript{9,\textit{h}}. Proline also plays a critical role because its structure allows for specific sharp turns in the polypeptide chain\textsuperscript{8,\textit{h}}. There are side factors to consider throughout the understanding of protein physics. Proteins are hard chains of amino acids tightly packed with atoms very close to one another; many depictions or illustrations show open space between each amino acid\textsuperscript{5}. Proteins usually also have co-factors like ions, small molecules, sugars, nucleotides, or fragments of nucleic acids\textsuperscript{5}. All of these non-peptide co-factors are involved in the function of the protein and in the formation. Proteins, in a solid state, act like a non-periodic crystal, they are firm at one environmental condition but may immediately melt given a change in that environment\textsuperscript{5,8}. This means they immediately transition from one state to another and do not “melt” but act more like a switch from ‘on’ to ‘off’.

Most typical proteins consist of around 300 amino acids, the size of a folded protein globule is 5nm, and these proteins fold into their native states within a tenth of a second\textsuperscript{8}. As discussed proteins are understood to be the molecular machines, or actuators, that carry out the vital tasks necessary for a cell. It is further understood that proteins are designed then to carry out regulatory functions such as binding and dissociating within a closed volume cell compartment of (V ~ 1um\textsuperscript{3})\textsuperscript{8}. This volume may contain anywhere from ten to one thousand copies of a regulatory protein. Requirements of concentrations at order of 1 molecule/(um)\textsuperscript{3} ≈ 10\textsuperscript{-9} M the binding reaction between the two proteins can be stated \( A + B \leftrightarrow AB \) in that neither concentration is shifted totally to the left or to the right \([A] \approx [B] \approx [AB] \sim 10^{-9} \text{ M} \)\textsuperscript{8}. We may further apply a dissociation constant \( K \) of the order \( K = \frac{[A][B]}{[AB]} \sim 10^{-9} \text{ M} \) that can be corresponded to the binding free energy equation \( \Delta G \sim -10 \text{ to } -15 \text{ kcal/mol} \)\textsuperscript{8}. This number is understood to be fundamentally universal for all protein interactions and also seems to be the difference in the free energy found between proteins of a single-domain in their folded to unfolded states\textsuperscript{8}.

The relationship is not a coincidence, if the \( \Delta G \) of protein stability was larger, then it would prove to be difficult to modify proteins with weaker protein-protein interactions. This relationship between macromolecular interactions and the cell volume is two-way. Looking at the converse relationship it can be determined that if 10 kcal/mol is found to be the typical protein-DNA or protein-protein interaction then cell compartment or volume is limited to ~ 1-10um\textsuperscript{3} \textsuperscript{8}. If this was not true then a much higher copy of proteins would be needed and changes to the chemical composition of the environment would be much more difficult.

**Structures of Proteins**

As previously mentioned, proteins contain a one-dimensional backbone synthesized from amino acids. This primary structure acts as a non-periodic crystal with the relative positions of each monomer being fixed giving the “native” functional state of the protein and held together by non-covalent interactions between each amino acid when in an aqueous environment\textsuperscript{8}. The structure of the amino group is centered on an alpha carbon, \( C_\alpha \), with an amino group head and a carboxyl group tail\textsuperscript{7,\textit{h}}. The difference between each amino acid is the side chain ‘R’ attached to the \( C_\alpha \). These side chains can be hydrophobic, e.g. alanine –CH\textsubscript{3}, to hydrophilic (charged), e.g.
aspartic acid (-CH$_2$-COO$^-$), or polar, e.g. serine (-CH$_2$-OH)$^{7,8,9}$. These properties play critical roles in the interactions of each amino acid in a polymer chain. Polar molecules have a permanent dipole moment giving them a charge and thus considered hydrophilic (water loving) due to their charge-dipole interactions with polar water molecules; hydrophilic molecules typically form hydrogen bonds$^{5,8,9}$. Non-polar molecules cannot form hydrogen bonds with water and are labelled hydrophobic (dislike water); when hydrophobic molecules are dissolved in water they go through a hydrophobic interactions, tending to cluster and aggregate to minimize their hydrophobic surface interaction with water$^{5,8,9}$.

Charged amino acids are located on the surface of globular proteins and are assumed to be important in defining or limiting the interactions between proteins and other proteins and/or DNA$^8$. Charged amino acids can carry either a positive (+) or negative (-) charge, with oppositely charged molecules having the ability to bind to one another through a salt bridge$^6$. The extent of the ionization on these charged amino acids depends on the pH of the solution. In lower pH the basic (negative) side chains are ionized and at higher pH the acidic (positive) side chains are ionized resulting in the overall charge of the protein being dependent on the pH of the environment$^8$.

The interactions of the 20 amino acids with each unique property was investigated by translating relative occurrences of contacting points between molecules in a real protein sample into two-body effective interaction strengths$^{10}$. The interaction strength between amino acids of type ‘i’ with type ‘j’ can be deduced by reviewing the overall frequencies of their interaction as neighbors in prior known protein structures$^{8,10}$. The equation $P_{ij} = \kappa P_i P_j \exp(-M_{ij}/k_B T)$ where $P_{ij}$ is the probability of ‘i’ and ‘j’ being neighbors, $M_{ij}$ is the effective interaction potential matrix that describes in parameters the nearest-neighbor interactions, and $P_i, P_j$ are the frequencies of the individual amino acid occurring$^{8,10}$. $K$ is a normalization constant that shows a constant can be added to all $M_{ij}$ and maintain the validity of the fit$^{8,10}$. Amino acids have an average of ~3 neighbors and when completely inside a protein it has 6 ± 2 neighbors, excluding the two that are on the polypeptide chain with them$^8$. This number is independent of the size of the amino acid but does vary with each structure. The average distance from the center of mass of the amino acids in contact with one another is $\sim 6–7$ Å$^8,10$.

Typical single chains of proteins are between 100 and 1000 amino acids and are in the primary structure. The peptide bond in the polypeptide acts as a partial double bond hindering the ability of the protein to fold. This structural feature requires the carbonyl carbon and the amide nitrogen and the atoms bonded to them to be in a fixed plane$^{7,3}$. No rotation is possible around the peptide bond itself. As a result of this the only flexibility in the chain backbone, to allow the protein to fold, is the rotation of the fixed planes on peptide bonds adjacent to one eachother around two bonds$^{7,3}$. The two bonds involved are the C$_\alpha$– amino nitrogen bond with the rotational angle referred as ‘φ’ and the C$_\alpha$–carbonyl carbon bond with the rotational angle referred as ‘ψ’$^{7,8,3}$. Only a small number of angles may be possible. If the atoms are too close the conformation would not be possible$^8$. The energy for φ or ψ rotation has three minima that are small barriers, $\Delta E < 5k_B T$, allowing the flip to occur quickly$^8$. The energy delta between the minima is also small $\Delta e \sim 2k_B T$ allowing for the persistence length of the random coil to be small, meaning the chain is flexible$^8$. The equation, $l_p \sim 0.38 nm \times \exp\left(\frac{e}{k_B T}\right) \sim 2 - 3 nm$, describes the energy landscape for ψ rotations where 0.38nm is the length per amino acid residue$^{8,3}$. The persistence length has been found to correspond to ~5 amino acids, very close to a typical single domain
proteins diameter\(^1\). This estimate has not been verified or considered certain. Contemporary measurements for the rate of contact formation for the ends of small peptide chains have suggested that \(l_p \approx 0.8\text{nm}\) \(^1^2\). In summary it can be seen that every amino acid adds entropy associated to each possible orientation. A protein that has 100 residues can be calculated to have a total local energy minima number of \(-6^{100} \approx 10^{77}\) if we take into account that \(\varphi\) and \(\psi\) can have 6 conformational minima per amino acid.

Polypeptide chains in an aqueous environment that want to fold into their secondary structure conformation face a problem. In each peptide the oxygen linked to the carbon and the hydrogen linked to the nitrogen can form hydrogen bonds\(^5\). In the swollen state of the polypeptide chain the hydrogen bonds would typically be satisfied with the water around them\(^8,13,k\). This interaction makes it not favored to fold the compact structure due to the exclusion of water\(^8,13\). The protein overcomes this issue by deciding to use compact structures where different residues along its chain can H-bond to one another\(^8,13\). Hydrogen bonding has limited directionality which places a constraint on the potential structure of the protein\(^8\). This is how the two regular conformations of polypeptides are derived: the \(\alpha\)-helix and the \(\beta\)-sheet\(^8,13,d,e,f\).

In the right handed \(\alpha\)-helix each of the N-H hydrogen in the backbone can form an H-bond with the carbonyl oxygen of a residue which is four residues away\(^7,8,e\). This allows for two H-bonds per residue resulting in a pitch of the helix at 5.4 Å, corresponding to 3.6 residues per turn\(^7,8,e\). The side chains in this conformation stick outwards which confers on the \(\alpha\)-helix its specific chemical identity; that is based on the pattern the \(\alpha\)-helix is globally amphiphilic since the polar residues will be on one side and the hydrophobic residues on the other\(^8,2\). \(\beta\)-sheets, the other regular element of secondary structures, are formed by parallel \(\beta\)-strands that form H-bonds with one another through the extended conformation of the protein\(^8,5,f\). The oxygen of the backbone hydrogen here also bonds with the N-H hydrogen on the adjacent strand\(^8,5,f\). The result is the side chain amino acids protruding above and below the plane of the sheet\(^f\).

These secondary interactions pack together and lead to the tertiary structure. The secondary structure elements such as \(\alpha\)-helices or \(\beta\)-strands go across the protein molecule from one end to the other and end up being connected by circular loops at the surface of the protein\(^8,l\). The predictions on \(\alpha\)-helices formation is much easier done that that of \(\beta\)-strands. Since \(\alpha\)-helices are separated by four residues along the chain the prediction is easier rather than the possible large separation of interacting amino residues on a \(\beta\)-strands\(^8,l\).

The fourth level of protein structure is the quaternary structure. Quaternary structure alludes to the spatial arrangement of each subunit and the way they interact\(^7\). Subunits are defined as an individual polypeptide chain\(^7\). Among quaternary structures the dimer is the simplest as it is made of two identical subunits\(^m\). Of course, quaternary structures may be much more complicated as more subunits are involved. The human protein hemoglobin is a quaternary structure made of two \(\alpha\) subunits and two \(\beta\) subunits, referred to as a \(\alpha_2\beta_2\) tetramer\(^n\).

### Folding of Proteins

Proteins undergo many forces but their stability is determined by residue-residue and residue-water interactions. All forces end as electrostatic in origin but each force that a protein deals with can still be sub-classified into van der Waals interactions, electrostatic interaction between charged groups, hydrogen bonding, hydrophobic forces, and disulfide bridges\(^8\). From a cumulative result of all of the mentioned interactions it can be determined that each amino acid
to amino acid contact can contribute a free energy order of $\Delta G \sim (-5 \text{to} +1) \text{kcal/mol}$ consisting of contributions from $H$ and $\Delta S$. When proteins fold their contact energy between residues has to counteract the entropy of the polymer folding. If we assume the effective possible orientations for every amino and take into account its side chain, $g=8$, and we may define $\Delta G_{\text{polymer}} \approx k_B T \Delta S \sim k_B T \ln g \sim 2k_B T$ per residue. The effective potentials between proteins do vary. As such it is best to only discuss further hydrogen bonds and hydrophobic interactions as they pertain to folding of proteins.

The secondary structure formation in proteins is hydrogen bonding. The potential found in hydrogen bond is very specific and has a steep angular dependence. The functional form $E_{\text{HB}} = D_0 \left[ 5 \left( \frac{R_0}{R} \right)^{12} - 6 \left( \frac{R_0}{R} \right)^{10} \right] \cos^4 \theta$ where the angle $\theta$ is counting the deviation from the ideal alignment between the donor and acceptor with $D_0 \approx 8 \text{kcal/mol}$. Hydrogen bonds are very strong and directional; this accounts for the forces that derive the $\alpha$-helix or $\beta$-sheet secondary structures.

Hydrophobic interactions are the other primary force in proteins that stabilize them. As discussed a tendency for a molecule to be hydrophobic is because it cannot form hydrogen bonds with the surrounding aqueous environment and as such reduces its solubility. The reported difference in free energy that can be associated with the exposure of hydrophobic groups to water is 4.8 kcal/mol nm$^2$ and is comparable to the strength of the hydrogen bonds. The associated free energy difference $\Delta G$ of transferring hydrophobic substance such as oil into an aqueous environment like water can be defined from the solubility constant $K$, $\Delta G = G(\text{oil in water}) - G(\text{oil in oil}) = -k_B T \ln K$, with $K$ as the equilibrium constant from the reaction of oil dissolving in water as in $K = \frac{\text{mol fraction of oil in water}}{\text{mol fraction of oil in oil}}$. With the mol fraction of oil in oil equal to 1 the hydrophobic substances will have a lower solubility in aqueous environment; this defines $\Delta G$ as a large positive number. Conversely, $\Delta H$ is small at room temperature and either negative or positive. We may measure the enthalpy change $\Delta H$ by using calorimetry, $\Delta G = \Delta H - T \Delta S$ therefore for these mentioned substances $\Delta S$ is a large negative number at room temperature. As such we can see a large entropy decrease for the system as a hydrophobic substance is dissolved in water.

The working theory is that the previously described entropy decrease is attributed from a partial ordering of water molecules which are surrounding the hydrophobic molecules. The given $\Delta S \sim 4 \text{kcal/mol} \sim 7 k_B T$ gives the suggestion that there must be many water molecules involved. The postulated mechanism to describe this ordering is that even though water molecules in water have a choice to form hydrogen bonds with their neighboring molecules, the water molecules that are closer to a hydrophobic surface are more restricted in their choice of hydrogen bonding partners and as such are forced to be partially oriented. With sufficiently high temperatures this ordering mechanism should melt away. With varying temperatures the hydrophobic forces also vary. Non-polar proteins enthalpies of transfer from inside the protein to the aqueous outside are increasingly unfavorable as the temperature increase.

The primary force governing the interactions of amino acids is hydrogen bonding. Each potential has enthalpy and entropy and each depends on temperature of the environment. The result of hydrophobic forces are protein folding whereas hydrogen bonding defines the specific structure of that folded protein by dictating secondary structures.
Conclusion

Reviewing the general properties of proteins gives us an insight into some key aspects and terms as they apply to protein biophysics. This fundamental knowledge is necessary in order to be able to build upon further details of biophysics and navigate your way through the maze of information. Additionally, some general cell biology may be required prior to the reader being able to fully digest the described material. Some basic molecular cell biology should be understood such as DNA transcription and translation, mRNA, ribosomes processes as mediated by tRNA, splicing, transcription factors, etc. are basic required knowledge. Having this base of knowledge allowed me personally to be able to see through the numbers and apply the logic of the physics to what I understood to be happening in the cell. Further information on the structures of proteins and how they fold begin to get more in depth and really ratchet up the difficulty of math and physics involved. This again is where I was able to draw on prior knowledge to understand post-translational modifications to proteins and how they affect protein structures as they develop into their native conformation.

The vast expanse and complexity of the protein universe is a very daunting challenge to undertake as a research study topic. My major for undergraduate is Molecular Bioscience and Biotechnology. I felt that being able to cross link the two worlds through physics would be a fascinating, endearing, and personal topic. Understanding the biological processes involved in protein synthesis certainly helped to draw parallels and build a solid base to apply the physics to. The intricacies involved in each little aspect of the protein world can be overwhelming. For every topic I touched on there was a plethora of data more to be had. I only scratched the surface. This is truly elementary protein physics. Even though I wanted to keep going more and more in depth the scope of this paper would not allow for it. It seemed even my current knowledge of biology and math was beginning to falter as the complexities of these protein interactions grew.

In the end this research left me very excited for the future of biophysics and in my own future. I am excited to be in a related field where I deal with proteins on a daily basis. The research being conducted on biophysics and protein physics are growing daily. With continued research the mathematical models of protein interactions will become very powerful. This will allow many medical applications. I imagine we will be able to in silico design and analyze proteins and their interactions. This understanding can be applied to vaccines, drug delivery, treatments of diseases, or diagnostics. If we can design accurate mathematical models that can predict the flow and delivery of a drug or vaccine then we may reduce development time of many helpful pharmaceuticals. This is just the tip of the iceberg; even though biophysics is not an entirely new field the recent advances in technology have really given the sub-field of physics a boost in development and application.
Figures

*Figure 1* – kinesin walking on a microtubule. Available from http://news.rice.edu/2015/02/23/motor-proteins-prefer-slow-steady-movement/ Kinesin walking on a tubule

*Figure 2* – Polypeptide structure. (a) Amino acids are linked together by peptide bonds formed by dehydration (b) polypeptides with free amino end (N-terminus) and a free carboxyl end (C-terminus). (c) A ball-and-stick model shows peptide bonds. Lodish H. et al.

*Figure 3* – Amino Acid Chart. From Dan Cojocari, Princess Margaret Cancer Centre, University of Toronto. Available from http://asterix.cs.gsu.edu/~weber/ProteinStructureAA.pdf

*Figure 4* – Standard structures of proteins. Berg JM et al.
**Figure 5** - The α-helix, common secondary structure in proteins. *Lodish H. et al.*

**Figure 6** – β-sheet, another common secondary structure. *Lodish H. et al.*

**Figure 7** – DNA converted into amino acid sequences of proteins by 1) transcription factors 2) initiation complex with polymerase transcription 3) Processing 4) translation and post-translational processes. *Lodish H. et al.*

**Figure 8** – Proline with its characteristic bend. Cysteine side chain with sulfur creating disulfide bond. *Lodish H. et al.*

**Figure 9** – Rotation between planar peptide groups in proteins. *Lodish H. et al.*
Figure 10 - Schematic energy landscape for $\psi$ rotations. \(^8\)Sneppen K. et al.

Figure 11 – Peptide bond association with hydrogen donor and acceptor with orientation to the backbone. \(^8\)Sneppen K. et al.

Figure 12 – Secondary structural elements span the protein molecule connected the loops. \(^8\)Sneppen K. et al.

Figure 13 – Quaternary structure showing Cro protein in a dimer. \(^6\)Berg JM. et al

Figure 14 – Quaternary structure of human Hemoglobin with two $\alpha$ and two $\beta$ subunits. \(^6\)Berg JM. et al
The diagram illustrates the changes in Gibbs free energy ($\Delta G$), enthalpy ($\Delta H$), and entropy ($T\Delta S$) as a function of temperature ($T$) in degrees Celsius ($^\circ C$). The temperature range is from 0 to 100 $^\circ C$.

The figure shows the Gibbs free energy change ($\Delta G = G_{\text{in}} - G_{\text{out}}$) as higher temperatures $T$ are applied. This change is associated with the dissolution of a non-polar solute, where the solute is represented by a circular shape and the surrounding solvent molecules are shown in a smaller circular shape.
References


A Look Into Photovoltaic Systems

Uche Nwambuonwo

Paradise Valley Community College

PHY112

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Dr. Durandet
Abstract

As mankind continues to advance in many facets of life including technology, longevity, and overall quality, there is no question that we need more energy to fuel this life. There are several issues surrounding the common ways in which we currently obtain energy. Some of the issues include the rate at which they are being depleted and the effects they have on our ecosystem. These issues are leading to new ways in which we can create affordable, clean sustainable energy that will meet the demands in our world today. One of those ways is through photovoltaic cells (PVC), commonly known as solar energy. In this paper we will explore the ins and outs PVC and see what the future holds for this form of energy.

Energy and our need to consume it

Mechanical, chemical and electrical energy are some of the basic forms of energy that we know exist today. In almost every activity that we accomplish in our daily lives as human beings, we need to consume some form of energy to accomplish them. The average human being is said to need about 2.9kWh of energy to accomplish his daily needs15, which is about 1060 kWh per year. In the 1890, the average energy consumption a year was about 5800 kWh; by 1970 it was 20200 kWh5,15. This was an enormous jump in just under a hundred years. Today the average human being consumes about 19000 kWh per year 15, this is still significantly higher than the 1060 kWh that each person needs for basic healthy life. This increase in consumption can be attributed to many factors mostly to our advancement in technology. With this increase in energy demand, the world has found ways to produce more energy, in comes fossil fuels and nuclear energy.

Fossil fuels

There are three forms of fossil fuels, coal, oil and gas. “85.6% of energy used in the U.S comes from fossil fuels.”2 Fossil fuels are known as nonrenewable forms of energy since we are constantly utilizing them and with no way of replenishing them5,11,15. Coal, retrieved through surface or underground mining is made up of carbon, hydrogen, oxygen, nitrogen and varying amounts of sulfur. Different types of coals have different amount of carbon and the greater the amount of carbon, the greater the energy output. 15

50% of the countries electricity is generated from coal but the pollution generated from coal is great and regulations are in place for coal industries to monitor the amount of pollutants.2,15

Gas was first used in the United States in 1816 and today accounts for 23.7% of its energy today2. A mixture of different gasses with the main ingredient being methane gas, it is the most environmentally friendly fossil fuel. Although industries are the biggest consumers of natural gas15, 60% of homes use gas for heating2,11. Natural gas will be depleted in 30-50 years2 if the rate at which it is being used continues.

Oil is the third form of fossil fuel. The demand for oil is so great in our world today and yet the production of it is sparse driving and keeping its prices high. Oil is refined to make different forms of fuel with gasoline being the important and most utilized. 37.2% of energy consumption in the U.S comes from petroleum2,11 and the U.S continues to rely more and more on other countries for oil. The production of and consumption of
petroleum contribute to air and water pollution\textsuperscript{5}, with majority of the pollution coming from the gasoline burned when cars omit them\textsuperscript{5,15}.

These forms of nonrenewable energy have been the source in which we obtain energy to fuel our lives. Along with nuclear energy, these fuels supply 78\% of our energy demands\textsuperscript{2,14}. The main issues with the use of these forms of fossil fuels are twofold, its damage to the earth via pollution and the rate at which it is depleting. With six billion people on the earth, consuming approximately $1.3 \times 10^{10}$ kW of energy each year\textsuperscript{14}, we are bound to run out. At this rate of consumption fossil fuel reserves will be depleted in 320 years and nuclear energy within 260 years\textsuperscript{2}. Though the number of years might seem outrageous and meaningless because none of us will live long enough to see it, it still has to be a concern because at the rate at which the world is growing (estimated to reach 10 billion people by 2050)\textsuperscript{2} the depletion might be even faster.

Along with its depletion rate, environmental issues are also a major concern with the use of fossil fuels. The increase in CO$_2$ into the atmosphere by these fossil fuels are said to increase temperatures on the earth surface which will be a catalyst to many ecological issues such as greenhouse effects and acid rain. These issues have led to the research of other renewable sources of energy that are more eco friendly as well.

Renewable sources

Renewable source of energy is any continuing flow of energy from natural resources. These are alternative forms of eco friendly energy that are either being used or explored today and account for 22\% of energy sources\textsuperscript{15}. Some of the well-known forms of renewable forms of energy are wind, solar, biomass, and hydro.

The Sun

The sun is necessary for all life. It gives light in the day and also brings warmth upon the earth. Without it our plants cannot grow, the process of photosynthesis (oxygen creation from plants) will not happen, humans will be unable to create vitamin D, and most importantly light skinned people will not be tan. The sun is also the greatest inexhaustible source of energy we have on the earth. Its energy is always present and available to us, free of charge, in fact “more energy from the sun falls on the earth in one hour than is used by everyone in the world in one year”\textsuperscript{10}. This available energy is also the cleanest form of energy we have compared to fossil fuels that currently cause pollution. “Heat from the sun causes temperature differences between areas, producing wind that can power turbines. Water evaporates because of the sun, falls on high elevations, and rushes down to the sea, spinning hydroelectric turbines as it passes.”\textsuperscript{10} 173,000 terawatts of solar energy strikes the earth continuously\textsuperscript{14}. That's more than 10,000 times the world's total energy use\textsuperscript{10,14}. All the stored energy of the earth’s reserve (coal, oil, and natural gas) is matched by the sun's energy in just 20 days of sunlight\textsuperscript{10}. The question is, with this abundance of solar energy, how can we convert it into affordable useable energy? Answer…photovoltaic.
What is Photovoltaic?
Photovoltaic is the conversion of light (solar energy) into electricity. The word photovoltaic is a combination of the Greek words phos, which means light and the word volt, the unit for electrical voltage.

Utility companies through a grid network of conductors distribute electricity to consumers. Traditionally, this is the electric power generated from a central source like nuclear, fossil fuel or hydro power plants and transmitted at high voltages through the grid to substations and then to industrial, commercial and residential consumers. As global energy need continue to increase, renewable energy sources such as solar, wind, micro hydro and biomass become significant. A photovoltaic (PV) system consists of solar (PV) module array or PV module and other equipments used to convert solar energy into electrical energy that is used to power electrical loads. Applications of PV systems started where connection to the grid is unavailable or cost prohibitive. These include space applications for satellite and spacecraft, portable applications for small loads like recreational vehicles, boats, preservation and transportation of medicine and vaccines in remote locations. Other valuable remote applications include road signs or signals, remote monitoring, lighting, water pumping, communication and off grid residences.

History
It had been common knowledge that the sun had the ability to create some electricity since the development of photovoltaic effect by French physicist Edmund Becquerel in 1839. However, the invention of utilizing Silicon as a means of generating electricity was an accidental one. In the early 1870’s, William Adams and Richard Day used selenium to create electricity and by the early 1880’s, Charles Fritts had invented the first PV cell by putting a layer of selenium on a metal plate and coating it with gold leaf. Fritts’ invention generated more electricity but not enough to be useful or compete with the gas- powered lighting that had become more affordable during this time. In 1931, a German scientist named Bruno Lange built solar panels out of selenium, but had the same issues that Fritts had encountered years earlier. Finally in the early 1950’s, two Bell laboratories scientist, Calvin Fuller and Gerald Pearson were attempting to improve silicon transistors for electrical equipment when by accident they created a PV cell that generated electricity when it was placed in sunlight. By 1953, the collaborative work of Calvin Fuller, Gerald Pearson and Daryl Chapin (another Bell Lab scientist), have created what we now utilize for solar energy.

Photovoltaic Effect
Although there are various semi conductors (can conduct some electricity in some conditions) used in solar cell manufacturing today such as crystalline silicon (c-Si), Gallium arsenide (GaAs). Some thin film technologies include; amorphous silicon (a-Si), copper indium gallium selenide (CIGS), cadmium telluride (CdTe). Crystalline silicon is the most popular and the element that this report is based on.

The building block for Photovoltaic solar technologies systems are solar cells which are a semiconductor material converting the sun’s radiation into direct current electric power. Silicon atoms have four valence electrons in its outer shell of its atomic structure. All the silicon atoms form bonds by sharing electrons with other silicon atoms giving them eight electrons making it a stable element but a poor semiconductor. In order to
make the atom unstable and make it a stronger semiconductor, a process called doping is performed. Doping is the process of altering the purity of silicon in order to make it a greater semiconductor, thus producing electrical voltage when exposed to sunlight\textsuperscript{5}. The silicon atoms are commonly doped with two different elements, phosphorus on the top layer of the silicon element and boron on the bottom\textsuperscript{5}. Phosphorus has five electrons in its outer shell, so when combined with silicon, an extra electron is present making it a negatively charged particle. Boron, which has 3 valance electrons, when combined with silicon creates a “hole” making it a positively charged particle\textsuperscript{1,5,8,11}. A P-N junction is created when the positive charged boron layer is combined with the negative charged phosphorus layer\textsuperscript{1,5,8}. This opposite attraction causes an electrical field when the extra electron from the top phosphorus layer finds a hole in the bottom positive boron layer. The magic happens when the photons from the sun hit the electrons in both layers knocking them loose, these electrons move around freely but don’t create any electricity until they reach the electrical field at the junction of the layers. The electric field pushes the electrons out of the junction and into the top layer of the silicon. The force of the moving electron pushes the electron out of the cell and into the metal conductor strips on the panel, generating electricity (See figure 3)\textsuperscript{1,5,8,11}. This electricity (direct current) flows from the metal strips, through a wire and into an inverter, which then converts the direct current into alternating current to be used in the home.

Solar Cell Manufacturing

Silica (SiO2) makes up more than 90% of the earth’s crust making it the second most abundant element in the world\textsuperscript{4}. Silica is reduced and melted between at temperatures between 1500\degree C-2000\degree C in a furnace. This process still keeps the purity of the silicon at 97%. This abundance of silica is another reason why the possibility of solar systems makes great sense and is worth the research that’s necessary to make it more efficient.

The process of manufacturing solar modules for solar energy generation starts with the quality fabrication of solar cell wafers. Purified silicon crystals are molten in a crucible and doped with boron to form the p-type silicon. Monocrystalline ingots are grown from single seed silicon crystal in a cylindrical motion and slowly drawn out and allowed to cool down. The ingot is cut into wafers using wire saws. The crystalline silicon wafers have a thickness of 180\,\mu m to 350\,\mu m.\textsuperscript{4,5} (See figure 4)

Crystalline silicon wafers are further processed to obtain wafers to be used in manufacturing solar modules. This is a very delicate process and requires significant precision in order to maintain its functionality. These processes involve chemical etching of the surfaces to remove the roughness created by sawing and also to create a texture for more surface area for the incident sunlight\textsuperscript{4,5}. Phosphorus diffusion is the process of infusing the phosphorus atoms in order to create the top n-type layer\textsuperscript{5,11,15}. Edge deletion is another process of removing the n-type material all around the wafer’s sides and bottom, in order to have a distinction between the n-type and p-type layers on the cross-sectional area of the wafers\textsuperscript{4,5}. Another process is the application of anti-reflection coatings. After drying, the screen-printing of grid patterns on the wafers and applications of interconnect ribbons on the wafer is performed. These are electrical contacts connecting one wafer to the next creating a top negative connection in a module. Finally
the entire back surface of the wafers are coated with a thin metallic layer of aluminum\(^4,5\) thus making the bottom positive connection. Wafers are arranged in series to form a module and with the interconnect ribbons going from the front contacts to the back contacts. The entire cell arrangement in a module is encapsulated with ethyl vinyl acetate (EVA) material\(^4,5\), this is then passed into a laminating machine where heat is applied, for water proofing the cell array, and aluminum frames are built around the laminate to give structural form to the module. A junction box containing by pass diodes is glued to the back of the module and the electrical wires are soldered in with MC4 connectors at the open ends for making series connection to other modules\(^5\). (See figure 4)

Electrical Characteristics Of Solar Cells

The current-voltage (I-V) curve characterizes all possible voltage and current operating points for a PV cell\(^5,15\). When voltage is zero on the curve, the current is at a maximum also known as short circuit current. When the current is zero on the curve, the voltage is at a maximum also known as the open circuit voltage. The power output of the solar cell can either be zero at both ends of the curve or positive integer any other point along the curve. The knee of the curve is the maximum power point, with corresponding maximum power current and maximum power voltage traced to the current and voltage axes respectively\(^5\). (See figure 2)

\[
\text{Power (P)} = \text{Voltage (V)} \times \text{Current (I)}\]

The irradiance and temperature under which a solar cell, module or array operates has an impact on the I-V curve. Five points on the I-V curve will be explained briefly.

Open Circuit Voltage (V\(_{\text{OC}}\))\(^5\)

This is the maximum voltage on the curve. The solar cell will be at open circuit condition or infinite load and there will be no current output

Power = V\(_{\text{OC}}\) x 0 Amps = 0 Watts

Short Circuit Current (I\(_{\text{SC}}\))\(^5\)

The maximum current on the I-V curve is the Short circuit current. Voltage is zero and therefore power output is zero. PV modules are current limited and therefore can be safely short-circuited for measurement purposes. However under sunlight, an array short circuit current can be significant to cause electrical wiring damage due to DC arcing in the contact. This can be avoided by using a conductor with switch or covering the array to prevent light on the modules. Only PV devices can be safely short-circuited. Other current sources like batteries and utility receptacles my not be short-circuited to avoid electrical and chemical explosion or fatality.

Power = I\(_{\text{SC}}\) x 0 Volts = 0 Watts

The current out of a PV module is directly proportional to the area and the irradiance. By doubling either or both of these parameters the power output is doubled or increased four times.
Maximum Power Point (PMP)

Every point on the I-V curve has a voltage and current values. There is a point on the curve where the product of the numbers is at a maximum. This is the maximum power point with corresponding maximum power point current (I_{MP}) and voltage (V_{MP}) on the axes of the curve. Anywhere along the curve has positive power output and is calculated from the following formula:

\[ P_{MP} = V_{MP} \times I_{MP} \]

For example what is \( P_{MP} \) of a solar module with \( V_{MP} \) of 19.5 V and \( I_{MP} \) of 2.56 A?

\[ P_{MP} = V_{MP} \times I_{MP} \]
\[ P_{MP} = 19.5 \times 2.56 \]
\[ P_{MP} = 50 \text{ W} \]

Other electrical characteristics that affect the performance of PV systems are includes fill factor, efficiency, temperature, system resistance and solar irradiance\(^4,5,11\). Temperature and solar irradiance are a function of the weather. Fill factor, which is a ratio maximum power to the product of open circuit voltage and the short circuit current\(^5\) as shown in the formula

\[ \text{Fill Factor (FF)} = \frac{P_{MP}}{(V_{OC} \times I_{SC})} \]

A decrease in Fill Factor over time is a good indication of degradation in the solar cells\(^5\).

Efficiency is the ratio of power output to the power input. Under standard testing conditions, the module has a nameplate rating for \( P_{MP} \), \( V_{OC} \), \( I_{SC} \), \( V_{MP} \) and \( I_{MP} \). The power output is the \( P_{MP} \). The power input is the product of the solar irradiance (\( W/m^2 \)) and the area (\( m^2 \)). For example, a module has a maximum power \( (P_{MP}) \) output of 50 W, when exposed to irradiance \( (E) \) of 1000 \( W/m^2 \), and has an area \( (A) \) of 0.395 \( m^2 \), what is the efficiency?

\[ \text{Efficiency} \ \eta = \frac{P_{MP}}{(E \times A)} \]
\[ \eta = \frac{50}{(1000 \times 0.395)} \]
\[ \eta = 0.127 \text{ or } 12.7\% \]

System resistances are two types; series resistance which are inherent in a PV system due to wiring lengths and connectors at various points in the system. Shunt resistance are leakages in the module cells or grounding faults in the array. Both affect the power output of the PV system. The best scenario is to have minimal series resistance and infinitely high shunt resistance.

Practical Home PV System Application

An example of a practical PV system in a home setting might look like this. Homeowners can offset their utility bills by installing rooftop grid (figure 6) tied PV system. Typically,
the homeowner will consult a solar installation company who comes to the home and performs an assessment of energy usage for the home and the size of the roof area. Based on the assessment, a PV system design is proposed to offset a percentage of the energy usage. Price of the PV system and the size of the roof area are limiting factors for the system proposed by the installer. When an agreement is reached and the homeowner signs the contract, design and installation begins. On a rooftop system, no more than 600V DC can be installed, these modules are connected in series (See figure 5) not exceeding the maximum voltage allowed. From the array, a PV output circuit goes through a DC disconnect which is similar to a switch to cut off power from the array itself. From the DC disconnect power goes to the inverter which converts DC power to AC power. From inverter, power is sent through a DC disconnect to the panel on the home, which then distributes power throughout the home. (See figure 1)

Based on the energy being used in the home at any particular time, the power from a PV array can provide the demand otherwise any extra power comes from the grid. When the array provides more than the home requires, energy is being added to the grid giving the homeowner.

Advantages, Disadvantages And The Future

The future of solar system is bright as its efficiency and technology continue to improve and prices are becoming more affordable as studies showing that prices of modules and inverters have dropped between 50% - 70% in the past 6 years. As mentioned earlier, there are many advantages to solar systems including minimal maintenance required, eco friendly, wherever there’s sun, custom design for the home. However there are also drawbacks with the greatest being the price and efficiency. Also the many parts needed to provide AC to the home can also be a drawback as well. Along with the price drop and government aid programs, there is no doubt that the use ok PV systems will continue to grow. In fact 18 gigawatts of was added to grid systems from 2008-2014.

Conclusion

After my research into this topic, I have become much more interested in this topic. The fact that solar energy has the potential to change the way we produce energy through something that is always available to us for free, the sun, is fascinating to me. Although there are obvious drawbacks that currently surround PV solar energy, the future is bright. As a young boy growing up in Nigeria and experiencing life without electricity, sometimes for months at a time, I would love to see solar energy implemented in the country of my birth to serve some of the 2 billion people in the world without access to electricity.
Reference


Figures

Figure 1

IV curve of the solar cell

The short circuit current, $I_{SC}$, is the maximum current from a solar cell and occurs when the voltage across the device is zero.

Figure 2

Electron and Current Flow in Solar Cells

Figure 3
Ripples in Spacetime Itself: Gravitational Waves

Erik Oswald
Astronomy 112
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Dr. Bill Sherry
Abstract

In this paper on gravitational waves, I will be covering a lot of what is known about gravitational waves. I will be going over the theory of general relativity from Albert Einstein, where he first predicted the existence of gravitational waves, among other things. I will also be covering the topic of how we actually managed to find and record these. This includes the LIGO system, and talking about the future of discovery in the project called LISA. Implications of these, as in how they change our perspective on what this Universe is, all the way to how it changes our views of binary star systems and the infinite unknown. The finding of gravitational waves was huge, and it is imperative to our successes in the future, as we move on in understanding where we stand in this Universe, and how the physics of it work.

Ripples in Spacetime Itself: Gravitational Waves

There are a lot of things in this universe we as a species do not understand. All of our theories and laws are our takes on trying to understand this vast unknown, and even then, these theories that can explain something 99% of the time, will still fail under certain circumstances. This has been shown to us time and time again, and as we learn more, we realize there is even more to learn. Gravitational waves are one such thing. These were something predicted long ago by a theory, and only about 100 years after this theory was first made, did the evidence show for this. Gravitational waves are ripples in the fabric of spacetime itself. Caused from potentially catastrophic events in space, these events manage to bend and cause ripples in gravity itself. They then behave and act with the same properties as other waves, traveling through space until it hits an object. Gravitational waves were, and still are, one of the most significant finds of modern astronomy, and all of astronomy for that fact. These can help us understand a lot more about the universe we live in, from painting a better picture of black holes, to helping us understand our theory of gravity better. The gravitational wave find was huge, and will help us as a species advance our space knowledge.

The discovery of the gravitational waves is a huge technological feat in of itself. The system that was made to detect these waves is called the LIGO system, standing for Laser Interferometer Gravitational-Wave Observatory, an interferometer that is used to detect gravitational waves, as the name states. According to LIGO’s official website hosted by Caltech, an interferometer is a device that uses lasers and mirrors to detect the smallest and most minute changes, like gravitational waves, which were said to have caused a wobbling effect thousands of times smaller than that of a size of a nucleus.

The LIGO system is a very large branch off of a device called a Michelson Interferometer, with a few modifications to enable the detection of gravitational waves. An interferometer works by using lasers that superimpose after being refracted and then reflected back. This process allows for the measurement of microscopic lengths, but even microscopic is not enough for gravitational wave detection. Due to physical and geographic limitations of space and size, the LIGO system is about 4 kilometers, or two and a half miles long between the refraction of the lasers, and the mirrors that reflect these lasers back at the end of the arm. This allows for even better detection of small changes, like the wobbling that a gravitational wave causes which is thousands of times smaller than the size of a nucleus or proton even. As it turns out though, even a two and half mile long laser is not enough to detect changes that are so minute, so this laser is bounced back and forth from mirrors before it reaches the refraction point again about 280 times, which makes the effective distance traveled to be about 1120 kilometers, or nearly 700 miles, all of this information being obtained from LIGO’s website. This is why physical and geographic limitations come into place, as 700 miles is not a manageable length to
get an uninterfered laser, and even if one could, Earth’s spherical shape also inhibits any possible solution.

Another testament to the amazing technology of LIGO is the what one might be thinking about at this point. If a machine made to sense the sensitivity something that is thousands of time smaller than a proton, how does it avoid motion from the rest of the Earth, like simple interference locally, or tectonic plate movements on the other side of the Earth? According to the LIGO website, LIGO uses forms of motion adjustment that are called “Damping.” First off, it is simply located in an isolated part of the Earth, where it can escape movement of everyday life like cars. Secondly, the use of “Active Damping,” which is a system within the LIGO system, that detects seismic events as they happen, and is able to counteract these events, allowing for adjustment to keep the lasers unwavering from unwanted movements. The next thing is “Passive Damping,” which is a system that holds the mirrors still on a contraption of four unique pendulums that are attached to the mirrors and hold it in place and keep the mirrors unwavering from anything that is not wanted, like all terrestrial movements. Lastly, LIGO is located in one of the purest vacuums on Earth, that is designed to keep all air molecules, and dust particles out, as to prevent any light refractions or disturbance that could send the LIGO system off. All of these, along with the other components of the system, help to make sure that the only thing that this massive and technology strenuous system detects are gravitational waves, not simple earthquakes or cars moving.

A couple of years ago, there were talks of using a LIGO type system in space, as to avoid all the nuances of a terrestrial based system. This would have been a space interferometer known as LISA, which stands for Laser Interferometer Space Antenna. As stated before, this would have been a space interferometer, that would have had the advantage of being able to obtain a larger distance in between lasers, and avoid any possibilities something like an earthquake messing up the lasers and setting off false alarms. Although there would not have been terrestrial problems, there are still things that can happen in space, like meteorites, or solar winds affecting the LISA system, which could have made for high reparation costs. Ultimately it was the costs that prevented this mission from going forward at that time. According to the NYTimes article by Dennis Overbye, this could soon be another mission though, as earlier this year, a test probe reaching its destination, called eLISA, that is supposed to be similar to the mother variant of this triangular system. The test was made to see if the point where the gravity of the Earth and Sun intersect, allow for motionless freefall. This test could prove to be extremely valuable, as if it is successful, we could have a full functioning interferometer in space, that is many times more sensitive than the ground interferometer we have currently. With the findings just made, it brings into consideration having a system as sensitive as this one would be. According the LISA section on NASA’s website, the plans state that the lasers would be around five million kilometers apart, which would allow for detections of the tiniest of gravitational waves, and as per usual, advance our knowledge of the Universe even more.

Gravitational waves were predicted nearly 100 years ago in the theory of general relativity from Albert Einstein. The theory of general relativity states that what we know as space is basically a piece of cloth, and this cloth is made from space and time, or the combined term, “spacetime.” It states that object with mass cause an indentation of this cloth, and that when another, less massive object gets caught in this indentation of gravity, that it will begin to orbit, or “free-fall,” into the more massive object. As such, the more massive the object, the more of a gravitational pull it will have, and the larger the area for one to get caught in its orbital path. So the part of the theory of general relativity that mentions gravitational waves is when
dense, massive objects get very close, or start to merge, they cause ripples in the very fabric of spacetime, and these ripples spread throughout the universe. This is much the same way as when a disturbance in a clear, still pond causes a rippling of waves across the surface, and these waves spread across the entirety of the pond. As such, these waves were first predicted in 1916, and it took us until 1974 to finally find an event that had the possibility of sending out these waves. Around this time, two neutron stars were merging, and at the point of mergence, it was predicted that this event would be sending out gravitational waves, but we did not have the tools to detect these at the time. Come 2015, with the LIGO system active, we finally could confirm the existence of gravitational waves, with real, certifiable proof that they existed. This particular wave was formed from two black holes, one of 29 solar mass, and the other of 36 solar mass, merged and created an event that was “catastrophic” enough, to cause a noticeable ripple in the spacetime fabric and allowing for us to feel this wave 1.3 billion years after the fact as it traveled through the vastness of space.

This discovery was one of the final pieces of the puzzle that is relativity, and truly helped to cement what exactly gravity is, and gives us a way to potentially measure gravity. Gravity has always been a theory to us, because although we can feel the effects of it, and it is always acting on us, we have no way to measure it, no way to see it. Being able to see these gravitational waves confirms that there is in fact a spacetime fabric, and that gravity can and does act on this, and that gravity can be “harnessed” and used to send out waves. The implications are huge, and they allow us to continue research and progress on this area of science with a clearer picture, a better understanding now of what we are looking at and observing, and what to look for. The observation of gravitational waves can help us to understand a whole host of interstellar objects and formations better. According to NASA, these finding could lead to the advancement of black hole studies, including formation and growth of massive black holes, watching how these massive black holes interact with their galactic companions on levels from local, to supermassive black holes in the centers of galaxies. They can help us to understand galactic structures, like how our very own galaxy operates and was formed. They also improve our understanding of general relativity, and how physics works under applied circumstances, and could potentially lead to new physics that explain things we did not understand or even know about before.

Gravitational waves give us a look into the way the Universe operates. With this discovery, and the confirmation of another portion of the theory of relativity, we now are able to apply other portions of this theory into explaining other parts of the Universe. One of the big theories now is what is the Universe. On a website called Einstein-Online.info, I obtained a lot of information related to Einstein’s theory, and how he predicts the Big Bang as the start of the Universe, and from there, we are trying to figure out the shape. With the gravitational waves confirmed, we are beginning to examine how these wave emit from their source, and then travel the Universe, and if they ever reach an end. Due to time and technological limitations, we cannot see further than the beginning of the Universe in terms of time. We can only look in an area around 14 billion years old, but we know that there is more out in the Universe than just that 14 billion year radius we can see. As such, it is hard to say what happens when we reach the end of the Universe, or if there is even an end, but we do have the theory of relativity which helps to explain what might be. As stated before, the Big Bang is the main theory as to how our Universe started, and from that we can observe and collect information and data on what is happening with the Universe currently, which is the expansion of it. We cannot say what it is expanding to or in, but we do know that the Universe is and always has been expanding, as per Einstein’s equations, and with the confirmation of the gravitational waves, we are able to more accurately say that it is
right, and we can start to use these waves to determine what is at the end of the Universe, or what happens. Common theories is Einstein’s work is that the Universe is a torus, or donut of some sort, so nothing actually ends, but everything just comes full circle and repeats indefinitely. Another part of this is that maybe instead of an indefinite loop, it is a more spherical shape, that takes on this donut identity, but when something originates and travels across the entirety of the Universe, it reaches a point where the wave more or less coalesces with itself, and it ends their. These both to lead to the mathematical model of our Universe, and that leads to a more or less cylindrical shape that has properties of both of these models. With the discovery of gravitational waves being new, all we have yet to do is speculate on what this means, but these could lead to us better understand what is actually happening in our Universe, because it is a still a huge, complex mystery.

Scientists have a very good understanding on what causes gravitational waves. The cause of gravitational waves is nothing too complicated, as it has to do with anything of mass making any sudden movement. It is a basic concept of energy, that energy can never be lost, only transferred. Waves are a form of this expulsion of energy. The easiest thing to think of for this concept is throwing a stone into a pond, and watching the waves of water ripple from the point of contact. These waves are formed as a form of energy being released, because as you threw the stone, you released the stored energy from the windup, and that made for kinetic energy. As the stone travels through the air, some of this energy is lost to heat transference, some is lost to the transfer of energy to air molecules due to friction, which also leads to heat, and the last of the energy is lost on contact with the water, where all remaining energy is transferred to the water in the pond, which it releases by sending out these rippling waves that travel the surface of the pond. As such, gravitational waves are most common among heavily dense or massive objects rotating and combining, to release a lot of this stored energy into space and effecting gravity itself, much the same way the stone released into the pond generated waves in the water. We know this because in 1974, a few scientists were watching the movement of two pulsar, neutron stars orbiting each other. They noticed that as these two objects orbited each other, and began to move closer together, their rotation started to slow down, quite significantly, which means this energy has to be moved somewhere, because energy does not disappear. The only thing that made sense was that these were losing their energy and creating gravitational waves. Alongside other things, they inferred that gravitational waves had to be dispelled, as that is what Einstein’s theory of relativity stated would happen when a massive, binary system orbited about each other in such a way. As stated before, this wave that was found by the LIGO system was created by a binary black hole system, each being extremely massive at 29 solar masses, and 36 solar masses. As the theory of relativity states, these are ideal conditions for emission of gravitational waves. A big problem that we have come to realize is that gravitational waves are very small, in that they do not seem to do a whole lot, as is showed by the LIGO being able to detect movements thousands of times smaller than that of the size of a proton. So for this particular event, it took the collision and a merging of these two extremely massive black holes for the LIGO system to pick up on the gravitational waves. Based on the observation of this event, the mergence of two black holes is a particularly catastrophic event, and the fact that we barely felt the gravitational waves, means dispersion and energy loss played a big factor in what little we felt, or will likely ever feel. A 1.3 billion light year journey is a long time, even for something like a wave, and as such, a lot of the energy was lost on the way to Earth.

LIGO works in a series of runs, and now that we have detected, studied, and observed our first gravitational wave, the future looks bright. As stated on Astronomy.com in an article by
Eric Bentz, The next LIGO run is said to starting sometime in the summer of 2016, and with what we know now, scientists are predicting that LIGO might be able to detect between 10 and 100 black hole mergers, which have shown us to produce gravitational waves. This leads to the hope that the next time LIGO runs, it could show us a slew of results, and we can further work towards the goals of understanding how and why our Universe is the way it is. Much the same way gravity is, black holes were only a theory, and we could only detect them through matter they would destroy, but could not directly observe them. Scientists hope that with the discovery of gravitational waves, coming from a binary black hole system, it leads to an advancement in being able to truly observe and record data from black holes, and help us to understand what one of the greatest mysteries of the Universe are. Another first with the observation that leads to an exciting future is the fact that a binary black hole system was never before seen or recorded before this incident. Scientists were not sure as to whether a binary black hole system was even possible, and even theories on this subject matter were very unsure. The discovery of these waves coming from a merging binary black hole system help us to show that not only are binary black hole systems a real thing now, but these systems produce gravitational waves that we can feel. On the same Astronomy.com article by Eric Bentz, as mentioned earlier, as we furthered our research on this topic, scientists have begun to notice and calculate that binary black holes are not actually all that rare, and that they happen more than we think. This changes our perspectives on our understanding of stellar evolution, and helps us to understand a whole field that is related to this subject matter. Binary systems are not uncommon, but finding a binary system where two black holes were 36 and 29 times more massive than our Sun was like striking gold. We can now understand better the evolution of binary system compared to a simple single star system. We can understand how two massive stars acted with each other during their death, and how they react now as black holes.

In conclusion, gravitational waves are one of the greatest discoveries of modern astronomy. This find shows us what takes place in areas of space that we only ever were able to observe. Now with LIGO and LISA coming soon, we are able to do more than simply observe space with telescopes, we can listen to space. This opens a new realm of possibilities for space, ranging from new theories and physics, to new and better ways to observe different objects and areas of space. This is the find that so many physicists never thought possible due to how little the gravitational waves actually affect anything at such a great distance away from the original event. LIGO was a huge step forward for the technological thinking of mankind, but with LISA being a new potential suitor for sensing gravitational waves, the possibilities to hear those in space that cannot be seen is ours for the taking. This find also changes our perspectives on most of the rest of the Universe we know. This find helps to paint a clearer picture of the beginnings of the Universe, as the explanation to how not just one, but two black holes managed to be so massive and so close to another is in the works. We also have learned that black holes can, in fact, be binary. For the longest time, we could only observe binary stars, which is a complexity in and of itself, but now with this finding, we can better explain binary systems in general, and this helps us understand the life cycle of a binary system, from birth to death. This event proved a portion of Einstein’s theory of relativity in a spectacular fashion. A 100 year old theory was proved to be correct, and under the perfect circumstances that were predicted. With this new found information, a whole new area of astronomy has opened up, and trying to figure out what this means now, and as we gather more information in the future, what this means then. All in all, the gravitational wave find was spectacular. First observed to be happening in 1974 according to Einstein’s theory, to now, in 2015, where for the first time, we could actually hear
the gravitational wave as it passed us by, nearly 1.3 billion years after the initial event occurred, and 100 years after this event was predicted to be heard. With the future of astronomy in the palm of our hands, scientists are getting to work on what this means for us, and are preparing for the next batch of data to come from LIGO in the near future.
Exploring the Red Planet

Alex Papamatheakis
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AST 112 12666
Dr. Sherry
Abstract:
The exploration of the unknown universe has always been a challenge since the beginning of the human kind. Mars, the fourth planet from the Sun in our solar system, has been the main target of human exploration during the last decades. The existence of Mars has been known since the times of Babylonians and Ancient Greece but during the last decades people have come closer to know the Red Planet more than ever before. It is very essential to review the known facts about the Red Planet and to understand its wonders and prospective. The human kind has achieved great goals in exploring Mars, but there are so many aspects that remain unknown and with the greatest question if there is water on Mars to be still debated. NASA and other international organizations around the world, are still working on exploring Mars, the Red Planet, with the hope that humans will be able to actually set a step on its rocky ground one day.

Body of Paper:

People have known the existence of Mars before even recorded history. It was first observed by the Babylonians 3600 years ago. It appeared on the sky as a fire ball and many other cultures had observed its existence. (Figure 1) Mars was known to ancient Egyptians, Greeks and Romans. It was appearing to them as a red star on the sky and they used to call it Har decher which means the Red One. Later in history Mars obtained its name from the Roman God of War. The exploration of the Red Planet has gone beyond naked eye and telescopes. During the 20th century the human kind achieved great goals in sending spacecraft missions on Mars. Many of the ancient mysteries about the Red Planet have been revealed. Today after spacecraft explorations, scientists discovered that the red color on Mars is caused by the rust on its rocks. Its surface is covered with iron oxide and tremendous dust storms move it around all the time. It is believed that billions of years ago Mars had water, oxygen and warm weather. These are the 3 conditions required to turn iron into rust. The name Red Planet is still though associated with Mars. Mars is the fourth planet from the Sun in our solar system and the one closest to Earth. The data According to NASA Web site Mars is about 142 million miles distance from the Sun and is about half the size of Earth with about 13259 miles circumference. Its surface area is about 28% of Earth’s with about 55,742,106 square miles. Since Mars is further from the Sun, its temperature is much colder than Earth. The average temperature is about -284 to 86 degrees Fahrenheit. The average temperature of the atmosphere is -81 F. The chemical materials that make up the planet are similar to Earth’s mostly iron rich basaltic rocks. The mantle is consisted of silicate rock. The core is most likely iron, nickel and sulfur but there is not information if it is liquid or cooled metal. The atmosphere is mostly carbon dioxide 95%, 3% nitrogen, 1.6% argon and some H2O vapor. Mars has an orbit period that equals 687 days in Earth and a rational period that equals 24 hrs, 37 minutes and 22.6 second so there are 668 days in a year of Mars. When Mars is at the closest point to Earth about 59 million km away, it can be observed with a small telescope. (Reference 6) According to Astronomy Today web Site Mars has two moons with the names of Phobos and Deimos. The moons were both discovered in 1877 by the astronomers Asaph Hall. Both moons are erratically shaped and it is believed they are asteroids that were caught by Mar’s gravitational draw some time long ago. None of them is large enough to have spherical shape and both have synchronous rotations. Their rotation forces them to always keep the same face in the direction of their parent plane. (Reference 5) In order to experience the conditions in Mars, people can visit the Haughton Mars project. The conditions created are high arctic with a rocky polar desert setting similar to the one in Mars. The reality is though that nothing on earth can be close to the conditions on the Red Planet. (Figure 2).
It is essential to understand that the exploration of the Red Planet became a great universal challenge for the last fifty years. It is very important to historically review the missions to Mars and the goals achieved. There have been a total of 43 international missions to Mars so far and there are more scheduled. Things were not always easy for humans in their missions to learn more and more about the neighboring planet. From the beginning of the human kind all people knew about the universe was coming from naked eye observations. Around 400 years ago Galileo invented the first telescope. Since after that the ambition to discover what is going on further than earth started to become a great challenge. The history of exploring missions in Mars began in the early 60’s. NASA has completed many missions some of them were a great success while others were a total failure. As it is stated in NASA web Site USSR started the first mission in 1960 but it was a failure since their spacecraft Karabl 4 didn’t even make it to Mars. In the time between 1962 and 1973 NASA built ten spacecrafts called Mariner to investigate Mars and the interior solar system. Mariner 3 did not get to mars but Mariner 4, just three weeks later on November 28 1964 was launched effectively after 8 months voyage. In 1964 NASA’s Mariner 4, was the first international success in exploring Mars. (Figure 3) The mission was a success by returning back to Earth with the first 21 images of the Red Planet. The success of the Mariner 4 mission was going to open the way for more future explorations. The ambition to learn more about Mars rose more than ever. NASA took the initiative on Mars explorations especially after some unsuccessful missions from the USSR. The Mariner 4 spacecraft did not actually land on the Red Planet but the images it returned were astonishing during that time. (Reference 7)

In 1976 the Viking Project became the first mission in the history of the United States to land a spacecraft safely on Mars. During these missions Viking I and Viking II gathered more detailed pictures of Mars. They also performed three biology experiments to look for signs of life. The experiments conducted showed no evidence of living microorganisms in the Martian soil. During the 80’s more missions were sent to Mars but without a success. According to Mars space Science article, there has been little new from Mars since the Viking mission. Without a spacecraft at Mars, there had to be another way of exploring Mars. The necessity of keeping the exploration alive, led NASA to the creation of a unique instrument. The Hubble Space telescope had been one of the few highlights of Mars exploration. Hubble space telescope cannot see surface details smaller than about 50 kilometers but can see Mars well enough to map temperatures, weather and color changes. The Hubble telescope also discovered that some meteorites on earth come from Mars. Hubble telescope is the first foremost ocular telescope to be found in space up to today. It is located on the ultimate mountaintop of our solar system. As it is places really high and above the deformation of the atmosphere, the rain clouds and the light pollution, Hubble has a clear outlook of the interior space. Scientists have used Hubble to study the most outlying stars and galaxies and the whole solar system. (Figure 7) Today the Hubble telescope remains active. It still provides NASA with valuable information not only about Mars but for the inner space also. (Reference 3)

During the following two decades no more missions were planned on the Red Planet. According to NASA In 1996 the Mars Global Surveyor (MGS) became the first successful mission after twenty years. This mission studied Mar’s atmosphere, surface and interior. During this mission it was observed that Mars has a very repeatable weather pattern. The weather conditions include many dust storms that reiterate in the same location within nine weeks at the same time it occurred in the previous year. MGS provided more detailed information about the atmospheric circulation and the dust storm occurrence as well as other surface properties. During this mission
it was observed that Mars’ surface includes many volcanic mountains. Olympus Mons the largest volcanic mountain in our solar system was discovered on Mars during this mission. Olympus Mons expends about 24km high above the lave plains. It has a bottom of about 600 km. (Figure 8) It was also discovered that volcanoes in the northern region are extremely massive and they distort the planet’s roundness. It was also observed a gigantic rift valley which was named Valles Marineris In this canyon system it was found that Arizona’s Grand Canyon could easily fit into one of the sides. A smooth surface that was also observed together with a better definition of shorelines leads to the conclusion that an ocean once existed in the northern plains. A spectrometer on MGS showed a large amount of hematite in the cratered highlands. Hematite is a mineral produced by chemical reactions and only in hot watery areas. Also there was an observation of magnetization of some ancient rocks although Mars does not have an active magnetic field today. The MGS cameras also gathered traces of sedimentary resources in the centers of old craters. There were also originated gullies created by new discharge of groundwater along the slopes of rifts and craters. This evidence indicates that heat may be interacting with subsurface water even today. The MGS continued to provide valid information on mars until it failed to respond on November 2, 2006. In January 2007 the MGS mission was terminated. Mars Global Surveyor was a successful NASA mission. During the same decade on December 4, 1996 another NASA spacecraft named Mars Pathfinder was launched for Mars. On July 4, 1997 it landed on Mars. This mission included a smaller size rover called Sojourner. This mission was able to analyze a great variety of rocks near its landing area. It was also able to collect soil data and more data on weather conditions. Based on the data collected, scientists discovered that a variety of Mars’ surface rocks displayed different compositions. Scientists came up to the conclusion that some of these rocks are a proof of more complicated geologic processes. This discovery was definitely a very significant one. Also some images from the cameras on Pathfinder also recommended that more water probably have flowed throughout that area than what it was believed before. The images increased the probability of water existence on Mars in the past. Overall the collection of images from the Pathfinder was great. During its mission it returned more than 16,500 images from the lander and a total of 550 images from the rover. (Reference7)

According to NASA In 2001 Mars Odyssey was sent for another mission to explore Mars further more. By this year, NASA scientists knew enough about the Red Planet’s atmosphere and weather conditions. The question of any life existence in Mars has begun to rise again. Odyssey is still in orbit. During its mission, it has collected more than 130,000 images and continues to send us information about Mars. The information gathered is focusing on geology, climate and minerals. Based on information from the Odyssey mission, NASA was able to produce detailed maps of minerals and chemical elements and spot covered water ice. Odyssey has provided extremely vital support to ongoing explorations of Mars. It relays data from Mars rovers to earth through the spacecraft’s UHF antenna. Since September 2011 the Odyssey mission has been operational uses spectrometers and electronic images to search for volcanic activity and the existence of any form of water today or long time ago. It is also searching for evidence that microscopic life had once existed or maybe is living now on Mars. (Figure 4) So far the spacecraft has identified large amounts of hydrogen in the Martian soil which supports the belief that large deposits of water ice might also be present somewhere in the interior of the planet. (Reference 8)
During the following years, NASA was not the only missioner to Mars anymore since other European countries and China organized different missions. In 2003 a European Space Agency (ESA) organized the Mars Express Mission. The spaceship was launched on June 2, 2003 and arrived at Mars on December 25, 2003. This mission included the robotic rover Beagle2. According to the “Mars Exploration Rovers” The mission of Beagle 2 was terminated on February 6, 2003 after an unexpected crash of the rover. During this mission ESA scientists announced that while the mission was in orbit around the planet’s atmosphere, water ice was first observed in the south polar cap. The Mars Express Orbiter still continues to provide scientists with high resolution images of the surface, radar sounding images of the inner surface and valuable data on the composition of the atmosphere. (Reference 2)

Based on National Geographic web site in 2005 the Mars Reconnaissance Orbiter project (MRO) was place in Mars’ orbit. This orbiter carries the most powerful camera ever sent in space. It can identify smaller objects such as a dinner plate. It also carries a sounder to locate subsurface water which is a very important consideration for future explorations on the Red Planet. It is also considered to be the first part of an “interplanetary internet” a very important service for future spacecraft missions. The MRO collected the strongest confirmation in human history that liquid water exists sporadically on Mars today. Scientists used images from a spectrometer on MRO to identify traces of hydrated minerals on slopes in puzzling lines that are seen on the Red Planet. They appear in several locations on Mars when temperatures are above minus 10 degrees and disappear at colder times. As of September 2011 the MRO spacecraft still remains standing by. (Figure 5) On May 25, 2008 a robotic probe named Phoenix landed near the north pole of the planet. It was the first spacecraft since the twin Viking to land on Mars’ surface. The Phoenix probe included many mechanisms recycled from the unsuccessful Polar Lander mission that was crashed near Mars’ South Pole. The Phoenix was solar powered and was not expected to survive the darkness and cold of the Martian winter. Its lifetime period was about three months. The Phoenix probe gathered sample soil that was more likely to include life traces which is completely different with the Viking Landers 30 years ago. Its digging arm exposed water ice several weeks after the spacecraft landing. In some trenches dug by the claws of Phoenix, a white matter was exposed that scientists believed it might be either water ice or mineral salts. This material was proved by scientists to be water ice. The probe’s cameras showed that small crumbs were spread by the digging graze and disappeared over the following days. This could be explained by evaporation of water ice. The situation of the atmosphere at the digging location would not have allowed the evaporation of dry ice. The remark of water ice in Martian soil by the Phoenix operation was the first one not located in either of the polar caps. (Reference 4)

For the years to come the exploration of the Red Planet continues even more and with much more success. As it is mentioned on “ Mars Exploration Rovers” on June 10, 2003 NASA also launched twin Martian rovers, Spirit and Opportunity. The Mars Exploration Rover (MER) mission landed on Mars in March 4 and 25 2004. Originally they were set on a three-month mission. The last communication with the Spirit rover was on March 22, 1010. In May 2011 NASA ended the Spirit’s mission and transitioned to a single-rover with just supporting the Opportunity rover. During its mission the Spirit rover covered an area of 4.8 miles of the Martian surface. It returned more than 124,000 images and conducted tests with its spectrometers and microscopic imaging tools Based on the data collected from both rovers, at some point in the past, Mars had fluid water at both landing sites. (Figure 6) Since after 2011, the Opportunity
rover continues to perform its mission on Mars. Actually on August 2011 it began its exploration of the Endeavour crater. (Reference 2)

Based on “Mars Exploration Rovers” the Mars Science laboratory (MSL) was launched on November 2011 carrying the rover Curiosity. Its mission will be similar to Sprit’s and Opportunity’s. The rover Curiosity is many times larger in size and much more technologically advanced. Its total cost was $2.5 and is the most sophisticated effort until now to search for life signs past or present. Its mission will be to a) determine if Mars was ever in the past able to support microbial life, b) characterize the Martian climate and analyze the evolution of its atmosphere, c) determine the geology of the planet including mineral composition of the minerals in the surface and subsurface, and d) learn as much as possible about this area in preparation for a future manned landing. Curiosity is about five times the size of Spirit and Opportunity. (Reference 2)

Since after 2010, more missions with less money required were sent to Mars. Based on NASA website in 2013 MAVEN (Mars Atmospheric and Volatile Evolution) became the second mission based on low budget. MAVEN will gather vital dimensions of Mars’ atmosphere. This data collected will help scientists to understand the remarkable climate changes that occur on Mars. MAVEN mission discovered that a long time ago, Mars had an atmosphere much denser that was able to support liquid water on its exterior environment. It is now believed that Mars might once have circumstances to support microbial life. A remarkable climate change occurred sometime and Martian atmosphere was lost. As it is observed a significant number of dry channels and minerals that are typically formed in water remain in Mars. This also leads closer to the belief that mars indeed had a watery past. Today though, Martian atmosphere no longer allows water to be steady at its surface. MAVEN is the first spacecraft collect data of the Martian atmosphere. MAVEN carries 8 science instruments that will gather data of the higher atmosphere. It is also focusing on the area about 80 miles above the planet in order to model the total higher atmosphere of the planet. (Reference 8)

In 2009 the United States and the European Space Agency announced their decision to cooperate and coordinate on Missions to explore mars. Their effort will be focused on 2016orbital and rover missions. The Mars joint Exploration Initiative (MEJI) program has the challenging goal of creating and launching robotic explorers able of recurring samples from Mars in the 2020’s. The ExoMars spacecraft which is officially called the Exobiology on Mars mission is a European project that according to “Mars Exploration Rovers” with the help of the United States is scheduled to launch sometime in 2016. The mission will include an orbiter and static lander in 2016 and a rover in 2018. The Trace Gas Orbiter (TGO) will deliver the ExoMars Entry Descent and landing Demonstrator Module (EDM) lander to the planet. The lander will analyze sources of methane and other gases on the planet while the TGO will be in orbit above the Martian surface. It is scheduled that two years later the rover will explore the Martian surface and near subsurface, searching for signs of life. It will also be examining water and chemicals and learn more about the evolution of Mars and evaluating the possibility of sending a sample return mission sometime in the future. The rover will use a drill that could go down to a depth of 6.6 ft. The TGO will be used as a relay station between the EDM lander and the rover on one end and earth on the other. (Reference 2)

The view about Mars has changed dramatically over the years of its exploration. Based on the information known so far, Mars was not always a cold, dry and geologically dead planet. The
prospective of a warm, wet, world where life may have arisen becomes more of a far reality of
the lost past. Over the years the Mars missions have achieved great goals. So much more is
already known about the Red Planet and more are to come. Nowadays the exploration of Mars
continues. Every day new discoveries rise and the debate of water existence in Mars has become
a universal issue. In national geographic web site it is mentioned that scientists speculate that
Mars experience the largest known floods in our solar system around 3.5 million years ago. No
one knows though where the floodwater did come from. As it is known today, Mars’ atmosphere
is too cold and thin to permit the existence of fluid water on the surface for long time. Images
from NASA’s missions suggest that underground water reserves and may break through the
surface as springs. The exploration of water in Mars is a constant process. (Reference 4) As it is
stated in lives science web site substantial tiny amounts of water would quickly evaporate in
Mars’ low pressure atmosphere. It is believed though that water from sources such as aquifers
could last long enough to pool left over liquid for at least a year. Recent research suggests that if
a great amount of water flowed from a source such as an aquifer, it could stay liquid on the
surface for a while, forming the puzzling features known as recurring slope line that appear on
Mars’ slopes during warm months. Despite its frigid temperatures, scientists believe that Mars
might be able to host lakes of water on its surface today. The belief of existence of some type of
water form in Mars changes completely the prospective of an actual human visit in Mars.
(Reference 1)

Weather it is possible for humans to visit Mars yet or not, several new missions are planned for
the coming years. Besides NASA, the European Union, Russia, China and Japan are planning
missions in exploring the Red Planet. These new missions will include orbiters, landers, rovers
and sample-return probes. It is actually the Russian Soyuz spacecraft that helps bring American
astronauts to the International Space Station (ISS) which is their only current destination. NASA
though remains very optimistic on further missions on Mars. According to NASA “Our quest on
Mars has been to follow the water, our search for life in the universe, and now we have
convincing science that validates what we’ve long suspected,” said John Grunsfeld, astronaut
and associate administrator of NASA’s science mission Directorate in Washington. If everything
stays on plan, NASA is planning another significant mission on mars this coming August after a
few years of delay. This time a NASA rover about the size of a small SUV will be landed on
Mars and will begin searching for signs that Mars once supported life or even if it still does.
(Reference 6) According to Space Program, this new mission scheduled will be the most
ambitious one on Mars yet, and the cost will reach $2.5 billion. This mission will mark a bright
spot for the United States space program. Although many goals have been achieved, the last few
years NASA has been facing great financial difficulties due to budget cuts. The Obama
administration has decided to withdraw United States participation from 2 unmanned
explorations to Mars. Both explorations were in association with the European Space Agency.
NASA officials though claim that in the future years, private contractors in association with
NASA are willing on spending billions of dollars to bring Americans back into deep space.
Currently they are working on preparing a new launch that will eventually take humans to Mars
and even beyond. According to NASA Administrator Charles Bolden “NASA is not only alive
but is poised to take the next big leaps into both human space exploration and scientific
discovery”. It is not just NASA that had difficulties in completing new missions. Although most
of the U.S missions have been successful for NASA it cannot be said the same for Russia. Since
1960 they only achieved 2 of the total 19 missions on Mars. (Reference 9)Whether NASA will

be able to support financially new missions it is not certain at the time. But the ambition on setting foot on the Red Planet will continue on.

Conclusion

The Exploration of Mars has been and still is the greatest challenge for the human kind. Over the years the information gathered from missions on the Red Planet have been extraordinary. Historically, NASA has been the greatest leader in exploring the Red Planet with other international agents to follow along. Since Mariner 4 mission in 1964 brought back the first images of the Red Planet, the humankind achieved greater goals in exploring our neighboring planet. Today we have valid information on the planet’s atmosphere, geology and climate. The data collected so far indicates that Mars supported life at some time in the past, but scientists have still to answer the question if any form of life exists on Mars interior today. The question of any water existence it is also still debatable.

During the last decades the exploration of Mars has been the target of space missions. The data we have collected so far is valuable but we are not just ready to step a foot on the Red Planet. Without a doubt, the ambition of actually sending a human on Mars will never stop until we actually achieve it. The Red Planet is the closest one to Earth and it is the one we have more information about. Future Mars explorations might be postponed for a while due to budget cuts but for sure the project will remain alive. The exploration missions on Mars will continue to gather more information and hopefully one day the human kind will achieve the ultimate goal of visiting the Red Planet.
FIGURES

Figure 1 (Mars as it looks like through a telescope)
www.nasa.gov

Figure 2 (The Solar System; Mars is the fourth planet from the Sun)
www.nasa.gov
Figure 3 (The Mariner 4 in 1964, was the first successful mission of the humankind on Mars)

www.nasa.gov

Figure 4 (The landscape of Mars as it was recorded from the Odyssey mission in 2001)

www.nasa.gov
Figure 5 (The Mars Reconnaissance Orbiter was placed in orbit in 2005)

www.nasa.gov

Figure 6 (The Opportunity rover still collects valid data on Mars’ weather conditions and geology)

www.nasa.gov
Figure 7 (The Hubble Telescope is one of the most significant tools ever designed to analyze data of the inner space)

www.nasa.gov

Figure 8 (Olympus Mons is the highest mountain on Mars and the solar system)

www.nasa.gov
References


   http://go.galegroup.com/ps/i.do?id=GALE%7CCX3727801503&v=2.1&u=mcc_pv&it=r&p=GVRL&sw=w&asid=2a41702c895b3f1d7593354ce1768321

   http://go.galegroup.com/ps/i.do?id=GALE%7CCX4019600123&v=2.1&u=mcc_pv&it=r&p=GVRL&sw=w&asid=05c48b06f26f23f4a39ddab717dbac4


Pulsars

Elizabeth Pinto

Paradise Valley Community College

AST112

Dr. Sherri

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Abstract

The following will notice Jocelyn Bell and her discoveries of pulsars. It will cover the history, formation, use, and characteristics of pulsars, including the supernova explosion and chemical reactions. Over time, many technological advancements have been made and will be noted. These advancements in technology have also given astronomers the opportunity to further uncover information regarding black holes, and the balance between time and gravity. The balance between time and gravity can further be noticed due to the abilities of pulsars. It will notice how these advancements will go on to change what we know about space.
History

In 1932, James Chadwick first discovered Neutrons. Before his discovery, the uncertainty principle stated that calculations had discovered electrons did not hold enough energy to stay within the nucleus (Nave, n.d., para. 1). Little did he know, his discoveries would not only explode scientific discoveries but also shed light on the outer space. A year later in 1933, astronomers Baade and Zwicky would then consider the presence of neutrons in stars. Astronomers considered this, and soon after in 1939, Oppenheimer and Volkoff would then put in print the theory of neutron stars (“Important dates” n.d., para. 1).

A group of representatives (including Bell) of the Cavendish radio astronomy group had a constructed a telescope. The telescope was available in 1967 when Bell accompanied Professor Hewish and started annualization. The telescope recorded and printed details of the information that had presented itself, and printed about 30 meters of chart paper daily. After two months of studying the data that arose, she found what she called “scruff”, or disruption within the documented signals that produced a brightly shining light. She speculated that these signals were not of any man-made developments. Looking back on previous transcripts, she found that this “scruff” had taken place, not beyond a distinct plot in the upper atmosphere. After three months, she found that these palpitates were no more or less than 1.3 seconds separated from each other. After making this discovery, Bell approached her supervisor, Tony Hewish, who told her that 1.3 seconds was much too rapid to be expected from a star, given its massive size. He concluded that the pulse from this “scruff” ought to have been man made. However, given the location of the pulse, he determined it had to have derived from outer space. Hewish and Bell explored many causes to account from such an uncommon pulse. They ruled out a reproduction of a radar from the moon, a satellite in a bizarre orbit, and quasars. They tried using another telescope in order to detect an imperfection in the original telescope but found nothing. Computations detected the pulses were aspiring well passed the perimeters of the solar system. They found the pulse period to be 16 milliseconds. The only explanation they could think of was “LGM” or little green men (aliens). What was thought to have been picked up were radio waves; however, after doing further research and data processing, they realised that the changes in frequency were a result of the mobility of earth rotating around the sun.

The group was not confident in making their discovery public, due to the unreasonable idea that little green men were trying to communicate with them. Bell, looking back, then found in multiple charts that this “scruff” was also deriving from an entirely separate part of the sky, and concluded that it must be from a natural source as it was exceedingly absurd that two alien cultures were attempting to access them concurrently. By the time January arrived, Bell and Hewish found two additional pulsations from separate origins. They named them “pulsars” or pulsing sources, a well-earned name. Together they composed an article portraying the first
source and proposed it to a journal called “Nature”, who made their discoveries public on February 24th, 1968.

Not too many days before the paper was made public by “Nature”, Hewish disclosed the information learned to his colleagues during a meeting of Cambridge astronomers. When listening to their conclusion, Professor Fred Hoyle proposed the remains of a supernova could be capable of making such a pulse. When the paper was published, it included the supposed cause of the pulse. Thomas Gold from Cornell University presented a more reasonable cause of the signals about three months later. He stated that the signals were due to neutron stars rotating on an axis. Gold declared that the stars were not displaying pulses, but rather giving off a consistent light beams from both ends of it’s axis. As a result of the signal being rotated, it was being picked up by pulses, which Bell had first described. He also made clear that only a neutron star could spin this fast, due to the way they were formed (Cambridge Physics, n.d., pp 3-6).

Formation

The process of a star becoming a supernova takes about 11,010,015 years. Only stars with a mass larger than 1.5 x 10^30 kilograms can become a supernova over time, this is approximately eight times more massive than our sun. Stars compose in very cold, dense sections of space, called molecular clouds (Cambridge Physics, n.d., p 8). Forces of gravity stretching the cloud greater than the durability of internal tension pushing outward causes the cloud to collapse, forming a new protostar. A protostar becomes a star when it receives its energy from nuclear fusion rather than gravity. (A star is born, n.p., p 1).

Due to gravitational attraction, a ball of hydrogen is formed and drawn together. When enduring a fusion reaction, it gives off heat, forming helium. The helium is then strained toward to middle of the star (core). This newly named star (now that it is only reliant on nuclear fusion) grows its helium core for about ten million years. Over time, the core runs dry of hydrogen, causing the helium to shrink due to gravity until its temperature is high enough to start it's on fusion reaction. This fusion reaction caused from helium forms carbon and oxygen. This operation takes about one million years. As the core progresses to reduce, heavier outputs (neon, sodium, magnesium) start fusion themselves- this takes about 10,000 years. In the next 15 years, more weight substantial metals such as silicon, sulphur, and magnesium form. In the last week of the star’s existence, silicon and sulphur combine to create iron. All of the preceding fusion resulted in energy beamed to the outer layer of the star. This prevented the gravitational collapse of the star itself. In the chain of events, the iron would perform fusion as well, but because there is no radiation to back it up, the star then collapses on itself. The force of gravity encroach on the iron atoms detrimentally. The force leaves electrons and protons no other options than to proceed to produce neutrons and neutrinos. A shock wave from the core pushes back exterior material
and assembles isotopes of each achievable element. This entire collapsing take only a few seconds. This massive explosion gleams more luminous than all of the galaxy put together. Heavier elements move at 15,000 kilometres a second. This results in a cloud of matter surrounding the central neutron star. This neutron star is around 30 kilometres across; however, it is extraordinarily dense at about $10^{18}$ kg/m$^3$. A sugar cubes worth of this material would weigh 100,000,000 tons. As the neutron star is formed, it spins on an axis. Out of each end of an axis are beams of consistently released radiowaves and radioactive particles, which look like light. This light can only be seen if it is pointing directly in one’s direction. As a result of the visual effect of its rotation, it has been called the “lighthouse” star. These particular characteristics not only make it a neutron star, but more specifically, a pulsar. Depending on the neutron star, it will rotate once a second or faster for the next 20 million years. In this example, the initial star would have to have been between 8 and 25 times the mass of the earth’s sun. The remaining neutron star is about three times the mass of the earth’s sun (Cambridge physics, n.d., pp 6,8-9). As a result of the pulsar’s mass, one would have to be moving 100,000 kilometers per second in order to not get sucked in, this is close to a third of the speed of light, which is 299,792 kilometers per second. The fate of this pulsar is to eventually turn into a white dwarf. The fastest pulsar recorded is 716 pulses in one second (Cain, video 1).

Had the star had 25 times or greater the mass of the earth’s sun, it’s core would have collapsed and had such great density that no type of radiation would have been able to be released. This, in it of itself, is a called a “black hole” (Cambridge physics, n.d., pp 6,8-9). If one were to see a black hole, they would actually be looking at what is called the “event horizon” (figure 2). Anything that dare pass the black hole must be accelerating faster than the speed of light in order to escape the black hole’s clutches. This happens to be quite impossible. Rule of thumb, stay away. Over a google of years, a black hole will eventually dissipate in a ginormous nuclear explosion (Cain, video 1)

Scientific Experiments

Many scientists have been learning more about gravity by conducting experiments using pulsars. Much attention was drawn to pulsars after research on them, resulting in two Nobel Prizes. Three scientists did experiments using pulsars, and made their results public at the American Association for the Advancement of Science gathering in Vancouver, British Columbia. Albert Einstein’s thoughts on gravity were made known when he released his theory of General Relativity in 1916 (Saxton, para 1-4). General relativity is defined from dictionary.com as “…all motion must be defined relative to a frame of reference and that space and time are relative, rather than absolute concepts…” (para 2). In a couple theories, gravity's actions change in accordance with the structure inside the neutron star. Stairs noticed that by tediously checking the rate at which pulsars pulse, that they can then calculate the characteristics.
She (Stairs) then noticed as studying many different neutron stars that its motions are not due to it’s structure, and stated that General Relativity was so far constant. The best research shown so far, supporting General Relativity, was regarding pulsars in a binary-star system along with additional neutron stars. The reason it was referred to as some of the best research, was because the laws of General Relativity were still in play with substantial amounts of gravity.

The mass of a pulsar can be calculated when applying the glitches of the pulse. When these neutron stars are young, the pulses are not always routine, and can occasionally glitch. These are due to the normal and superfluid in the crust collaborating. These pulsars are made up of loaded amounts of neutron material beyond nuclear frequencies. Most durations of pulses are between 1.4 milliseconds and over one second. The timing of these pulses are used to uncover gravitational waves from massive black holes from far away. Because these pulsars are emitting such high amounts of electromagnetic radiation, the loosing of the masses energy is due to the energy exerted from the rotations. Eventually, this explains why pulsars end up moving much more slowly. Amid these “glitches” the rotation boots for just less than 30 seconds, and then goes back to its preexisting revolution for a more substantial amount of time, up to hundreds of days. Pulsars PSR B0833−45 and PSR J0537−6910 are perfect examples of such glitches. These glitches are thought to happen do to the neutron superfluid reaching beneath the crust of a pulsar. The three parts of a pulsar are the outer crust, the inner crust, and the core. After the supernova occurs and the pulsar is formed, the outer crust is made up of crystalline, whose mass density is about 4 x 10^11 g cm^-3. The inner crust is neutron-loaded nuclei within a free neutrons. This is considered of superfluid character due to the intense temperature, or Tc. The core is constructed of liquid neutrons. Although the pulsar itself might slow down, this superfluid is not compliant. This then makes the superfluid operate as angular momentum. Over time, the difference between the rotations and the superfluid are incongruent. Eventually, the superfluid momentum affects the entirety of the star, making this glitch (G, Espinoza, Antonopoulou, Anderson, 2015).

Pulsars in Orbit with Other Stars

It wasn’t long after pulsars were discovered that astronomers found one orbiting another star. It was uncovered by Joseph Taylor and Russell Hulse in 1974. This earned them, too, a Nobel Prize in Physics for testing thoughts about gravity, including Einstein’s Theory of General Relativity. The two stars were found orbiting each other every eight hours. They are about 432,687 miles apart and are moving about 300 kilometers per second. The strength of the gravity between the stars is so intense that Newton’s theory of gravitation cannot not accurately portray their movements. Their oval orbits are not constant, but move slightly every year, about four degrees. Because of the amount of gravity between the star’s, radiation is formed and some orbital energy is lost; meanwhile, one centimeter is shed from each star daily. This too was foreseen in General Relativity (“The First Binary Pulsar,” n.d., para 1-3)
If two stars orbiting each other wasn’t enough, a pulsar was discovered orbiting two other stars. Pulsar PSR J0337+1715 was uncovered by Scott Ransom with his colleges from the National Radio Astronomy Observatory in 2007. They studied this construct continually for a year and a half. This pulsar in particular is 1.4 times the earth’s sun’s mass, and has a velocity of 366 revolutions per second. It’s first orbit is with a white dwarf, whose mass is only 20% of the sun. The second star, is also a white dwarf 41% the mass of the sun, and orbits between these two stars every 327 days. The group stated that there very well might be over 100 constructs as these in the Milky Way alone, and notes it’s rarity. So here is how it works: the mass of an object decides how much resistance or push its giving, or its gravitational mass. The equivalence principle states that the gravitational and inertial masses are one in the same. This is the same as two objects, regardless of their weight, free falling to earth at the same time, as Galileo once proved. Einstein's equation: E=mc^2 states that the energy of an object is equal to it’s mass. This means that the gravity of the object is completely equivalent to the mass of the object itself. This construct is another example of Einstein's General Relativity. This is very hard to prove in detail when it comes to pulsars, due to our current lack in technology and the weakness of the Milky Way’s gravity. It is only able to be calculated to parts per thousand. This construct gives the opportunity to test two preceding methods. The gravitation of PSR J0337+1715 only counts for 10% of its mass. The write dwarf on the other hand, only accounts for not even 0.001%. In other words, the pulsar and the white dwarf orbiting, is being orbited by the second white dwarf. The scientist conducting the study will most likely know which star is attributing to the most gravity within a year, according to Ransom (Cho, para 1-12).

Other Theories and Discoveries

One of the pieces of research that won a Nobel Prize was in 1933. It suggested that there could be a disruption between space and time had gestures of planets, stars, etc, given off gravitational waves due to their movements. Currently, astronomers are applying pulsars to their research of gravitational waves throughout the universe. By tediously recording pulses throughout the Milky Way, astronomers are trying to uncover calculations that prove gravitational waves were cause minor abnormalities. Their most significant reason for research is to uncover that gravitational waves were due to aftermath of black holes, cosmic strings, and occurrences post-Big Bang, which they are still working towards. Another scientist doing such research said that with improving telescopes, that they hope to make much more precise measurements and uncovering these gravitational waves in the decade to come. Pulsars are so researched and experimented with because of their density, which scientists are still unsure of their physics. The second Nobel Prize was given out when Pulsars were discovered later in 1967 (Saxton, para 9-14).
The brightest pulsar ever uncovered was discovered at Texas Tech University. This sighting was actually unplanned. The telescope was aimed at a dead star just after it’s supernova explosion. Scientists previously believed that these extremely luminous X-rays were black holes consuming close stars. When this happens, gas secretes out and light is dispersed, millions of times more luminous than our sun. They’ve noticed 20 of these lights so far that are coming from pulsars, rather than the violent death of stars by black holes. The beams on the pulsar found had the strength and brightness of 10 million suns and is similar in width of Pasadena. It is by far the brightest pulsar ever noticed and documented. By comparison, it is 10 times that of most other X-ray pulsars. Astronomers theorize that the reason this pulsar shines so brightly is due to its abilities to take matter from nearby stars (Davis, Brightest Pulsar Recorded).

Conclusion

Overall, I thought pulsars were definitely one of the most beautiful, and crazy astronomical discoveries ever recorded. I enjoyed reading about them, and learning their formation, from birth to death, the explosion of the supernova, and what makes the pulsars have their magnetic fields. Learning how they form chemically and what is within their outer crust, inner crust, and core, was extraordinary. Three Nobel Prizes given for the founding of the pulsar, the suggestions of correlation between space and time (and how the pulsars could help detect this), and uncovering one pulsar orbiting another star. All of the advancements in technology and the studies done to uncover more about pulsars was very well explained and frankly, quite amazing (figure 1).

I am very excited to see what is coming in the next couple years regarding exact measurements of density, and the effects between gravity and time that can be explored using the pulses of a pulser. It is truly hard to imagine that these were found by mistake. Astronomy generally isn’t my thing, but these discoveries were absolutely astonishing. I truly cannot wait to hear more from Cho, Davis, Cain, Bell, and the rest of my references. It is absolutely crazy how many advancements in technology have been made since the first discovery of pulsars. Over time I expect to see more accurate measurements and additional theories regarding how our universe works.
Figure one

Figure 2
References


The Physics Behind Bowling

Iliana Ponce

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Dr. Casey Durandet
Abstract
As simple as bowling may seem, it is actually a complicated sport. There are many aspects of physics behind the sport of bowling. Some of the major parts of physics that bowling has include Newton’s laws of motion, energy, trajectory of the ball, friction, and elastic collisions. Different types of bowling balls will affect the trajectory of the ball. All of these physics concepts will be explained in a way that apply to bowling.

History of Bowling
Bowling has a long and rich history which makes this one of the most prevalent sports in the world today. There was evidence found thousands of years ago that shows how long bowling has been around. A few artifacts were discovered that could be dated back to 3200 BC. In the 1930’s, a British anthropologist named Sir Flinders Petrie discovered a collection of objects in a child's grave in Egypt that appeared to be used for a primitive form of bowling. There was a crude version of the bowling ball and pins that were all sized for a child and were very primitive. This shows us that bowling has been around for more than 5000 years, making bowling one of the oldest sports on record.

There is a history of a variety of games that are much like modern-day bowling. In Europe there are nine pin variations, Petanque is popular in France, and there is also bocce ball which emerged in Italy and is identical to British lawn bowling. However, in 1841 the ninepin game was outlawed as a gambling game in Connecticut but this did not stop people from bowling. To get around the law, the rules of the game were changed in the same century just by adding one extra pin. When the ten pin game emerged it allowed people to keep bowling. In 1895, a variety of bowler’s clubs got together with a restaurant owner to establish the American Bowling Congress at Beethoven hall in New York City. Despite the fact that women had been bowling since the nineteenth century, this congress only allowed man. However, in 1917 the woman of bowling established a congress of their own named the Women’s National Bowling Association.

Throughout the years bowling became more and more popular in America as new technologies improved the sport. By the 1950’s, media embraced bowling and NBC had “Championship Bowling” making the sport grow even more. In 1951, the “pinboy”, responsible for setting up the pins, was eliminated to add an automatic “pinspotter” in 1952. The length of the ten-pin bowling lane is 18 meters. Bowling balls, electronic scoring, and monitors that allow us to view the speed and path of the balls were all upgraded. Due to its popularity, bowling alleys kept growing in numbers across the country and are still growing today.

Different types of Bowling Balls
The most important piece of equipment for this sport is the bowling ball. There have been several modifications to the bowling ball throughout the years. Originally they were designed in such a simple way that they didn’t have gripping holes for the thumb and fingers and the bowler had to “palm” the ball. The earliest bowling balls utilized were made out of wood called “lignum vitae” which were light weight and durable. However, these wooden balls did not have as much bounce as the materials used today. The first rubber ball was made in 1905. In 1914, a ball made with “mineralite”, a cement used to create durable surfaces, was developed by the Brunswick corporation. Since the 1980’s until now, both the internal and the external parts of the bowling ball have had significant changes.
Today, ten-pin bowling balls are created from urethane, plastic, reactive resin, or a combination of all three materials. They are manufactured with a hard outer covering and a weight block (symmetric or asymmetric) inside the core. Bowling pins can determine the alignment of the weight block to know where the finger and thumb holes need to be drilled. A bowling ball containing a symmetrical core divided into two equal sides. A symmetric weight block requires one bowling pin to interpret the position of the weight block located in the core due to its symmetry, where an asymmetric weight block requires two bowling pins due to its non-symmetry. An asymmetric core is designed so that when the ball is thrown down the alley it will sidespin and hook to the lane. This indicates that the weight and structure of the core will have an impact on the way the ball will spin and curve down the lane. Bowling balls are available in a couple of different sizes, which are five-pin and ten-pin. “Ten-pin balls are typically eight and a half inches in diameter and contain three finger holes. Five-pin balls, on the other hand, have no finger holes and are five inches in diameter”. Additionally, different bowling balls with different weights are designed for both adults and children. Adult sized bowling balls come in 10, 11, 12, 13, 14, 15 and 16 pounds. Bowling balls that are specifically designed for children are available in 6, 8, 10, 12, 14, 15, and 16 pounds.

Reaction of each type of Bowling Ball
There are different styles of bowling for which distinct types of bowling balls are available. Some bowling balls are designed for hook shots and others are for bowling straight. Different core shapes and different chemical covers are used for bowling balls designed for hook shots. Although the four types of bowling balls have the same shape, they each have unique traits and will act differently as they travel down the lane. The surface of the lane may affect the trajectory of the ball. The amount of oil the lane contains may impact the direction of the ball and the level of friction it will have. The more oil the lane has, the less friction there will be and the ball will roll faster. Urethane balls are soft and will drag down the wooden lane causing the hook potential to increase. Typically, urethane balls react the best to oil. Some chemical covers permit a bowling ball to hook more to the lane. Resin is one of the few types. Reactive resin bowling balls are almost identical to urethane balls. They are manufactured by including small pieces of resin to the identical mixture that is utilized for urethane balls. Reactive resin balls are sticky, therefore the ball’s grip increases and more hook potential is generated. Resin balls also react well to oily lanes. The resin cover is designed to move and absorb the oil on the lane, which creates a path where there is less oil and the amount of hook on the bowling ball is increased. Bowling balls made out of particle are almost like reactive resin balls, except there are small pieces of glass added which helps it grip the lane and it is much easier to control the spin and hook. The most common types of bowling balls and cheapest to manufacture are plastic, which are made out of polyester. Plastic balls are also known to last longer than any other type of bowling ball. The downside to plastic balls is that they have a tendency to only skid down the oiled bowling lane as opposed to rolling, making it more difficult to control the ball.

Newton’s Laws of Motion
Physics plays a major role in the sport of bowling. Newton’s three laws of motion (inertia, force, and motion), are all essential to understand why a bowling ball acts the way it does on the lane and how it collides with the bowling pins. Inertia exists when an object is in a state of rest or uniform motion in a straight line, unless there is an outside force acting on it that causes it to
change its state. Although all objects resist changes when they are in motion, some objects tend to resist changes better than others. An object with a greater mass has a greater tendency to resist changes. A bowling ball will remain in motion as it travels down the lane until there is an outside force that acts on it, such as when the ball strikes the pins. A force will always be applied to the ball. Newton’s second law says, “The acceleration of an object is directly proportional to the net force acting on it and is inversely proportional to its mass. The direction of the acceleration is in the direction of the net force acting on the object”. The greater the force applied to the ball, the faster the ball will go and the more accelerations it will have. When we deal with a ball of a greater mass, the acceleration will decrease. Newton’s third law states, “Whenever one object exerts a force on a second object, the second object exerts an equal and opposite force on the first”. This means that in every interaction, a pair of forces acts on the two interacting forces. The direction of the force on the first object is opposite to the direction of the force on the second object. The bowling ball will knock down the pins with some amount of force, and the pins will respond with an equal force.

Energy in a Bowling Ball
Conservation of energy is a very important physics principle that applies to a bowling ball. Energy can neither be created nor can it be destroyed. However, energy can be transformed from one form to another and can also be transferred from one object to another when some type of work is done. When the bowling ball is thrown down the lane, it starts with some amount of energy that is created when we swing the ball and it changes as it makes its way toward the pins. In bowling, energy changes into both potential and kinetic energy. By lifting the object to a certain height, there will be a certain amount of energy stored. Stored energy is called gravitational potential energy. Depending on the height from which the bowler is winging and then releasing the ball, it will have some amount of gravitational potential energy. The greater the mass of the object, the more potential energy the object holds. Likewise, the higher the object is, the more gravitational potential energy it will have. The greater amount of potential energy the ball has as it is released down the lane, the more it will bounce. As a result, the possibility for a hook-ball to catch friction against the lane is very minor. This is proven by the formula $U = mgh$, where $U$ is the gravitational potential energy, $m$ is the mass of the bowling ball, $g$ is the acceleration of gravity, and $h$ is the height of the bowling ball. From this formula we see that the potential energy is directly proportional to the mass and the height of the object.

Energy will then be transferred from gravitational potential energy to kinetic energy. At this point the ball is rolling down the lane toward the pins. Depending on the mass of the object and how fast the object travels, it will have some type of kinetic energy. The remaining kinetic energy once the ball is reaching the end of the lane is used to collide with the pins. The heavier the ball, the more kinetic energy it will have. The faster the bowling ball rolls down the lane, the more kinetic energy it will also have. This can be seen from the formula $K = \frac{1}{2}mv^2$, where $K$ is the kinetic energy, $m$ is the mass of the bowling ball, and $v$ is the velocity of the bowling ball. From the formula we see that the kinetic energy is directly proportional to the mass and to the square of the speed. Therefore, more energy is required to move an object that has a greater mass and more energy is also required for the object to travel faster.
Trajectory of the Bowling Ball
The trajectory of a bowling ball is a curved path. Analyzing the physics of bowling is very useful because it allows one to understand the factors that influence how the bowling ball curves, and how one can make the best possible shot when striking the pins. The goal is to throw the ball and make it travel at a certain angle. Throwing the ball at an angle improves the chances of knocking down all the pins and making a strike. A ball that is directed in a curved path, will hit the pins at a greater angle than a ball that is directed in a straight line. Controlling the degree of the angle is important to make the best possible shot. The ball should be thrown at an angle that is roughly 30 degrees to evenly spread the force of impact to the bowling pins. The ball should also be released with a force parallel to the horizon for maximum impact. The greater the angle, the better. Due to friction, the angular velocity changes direction as the ball travels down the lane.

Three Phases of Bowling
As the ball makes its way down the lane it passes through three phases, the skid phase, the hook phase, and the roll phase. The ball also goes through two transitions. Friction also plays a major role in the sport of bowling. Friction is responsible for slowing down objects that are in motion. “In bowling, it occurs between the surface of the bowling ball and the surface of the alley. If there is less oil on the lane, there will be more friction and the ball will travel slower. If there is more oil on the lane, there is less friction and the ball will travel faster”. The total amount of friction will also depend on the weight of the bowling ball. When the ball is first released and begins to travel down the lane, it is called the skid phase. As the ball is traveling down the lane, friction will reduce the ball’s speed and increase the ball’s revolution rate. To protect the lane, it has to be oiled especially during the first stage of bowling right before the ball starts rolling. The speed with which the ball is thrown will also have an impact on the ball’s overall motion. The skid phase is the shortest part of a bowling ball’s motion, where the ball isn’t actually rolling, but instead it is skidding down the lane. Once the ball leaves the skid phase, the first transition will occur. The ball will leave the skid phase and transition to the hook phase. During this phase, the ball will somewhat slow down while it hooks to the lane. Once the ball is hooked the initial speed can begin to pick up. The ball will then go from the hook phase to the roll phase, which is the second transition. The roll is the final phase, which happens when the ball has finished hooking and actually starts rolling. The second transition is also indicated when the ball enters it’s entry angle and stays constant. This is when the ball has the most power to hit the pins and take them out.

Elastic Collisions
During an elastic collision there is no loss of kinetic energy. When the bowling ball comes in contact with the pins, the pins move in separate directions and some amount of the kinetic energy is transferred from the ball to the pins causing the pins to bounce against one another until they all fall down and run out of kinetic energy. Any interaction between the bowling ball and two or more bowling pins is an elastic collision. Pins bounce against each other because momentum is conserved during the collision. Momentum is formed by the velocity and mass of the object.

Conclusion
In conclusion, we now know how physics can relate to bowling. I have learned that bowling is a sport that evolved as a child’s game and is now played by everyone, mostly in the US. The sport
of bowling grows day by day. The rules of the game are always changing because we are constantly thinking of new ways to improve the game. Today, bowling is seen as a fun activity to enjoy with friends and family rather than a competition. Bowling establishments keep thinking of new ways to attract people to go bowling. Food and drinks are offered at every bowling establishment to enjoy while enjoying a fun bowling game. I think that the reason why people enjoy bowling so much is because weather does not play a role in this sport and families can spend quality time together since this is a sport that both children and adults can enjoy. Although physics plays a major role in bowling, we don’t think of it that way and we just enjoy the simplicity of throwing the ball down the lane and knocking down the pins. Bowling is a sport that has demonstrated international appeal and will keep growing as long as we keep playing it.
Figures

Source: (Stephenson, 2015)
This image shows the curved path the ball should have to come in contact with all the pins and score a strike.

Source: (Stephenson, 2015)
In this image we see how the kinetic energy is evenly distributed throughout the pins.
References


Time: An Absolute Definition

Thomas K. Poulson

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Abstract:

Einstein revolutionized physics in the early 1900’s. His work on special relativity, leads to the coining of the term “Space-Time”. Einstein incorporates the concept of Space-Time to write his greatest work on general relativity. It is the grandest theory of the Universe to date. It accurately predicts the behavior of Mercury and the rest of the Universe. But, it has a few problems: it does not mesh with Quantum Mechanics; it has not lead to the Unified Theory of Everything; and dark matter and dark energy have been “invented” to explain observed behavior in the Universe. Could the issue be that Einstein never defines time? If a Unified Theory of Everything is to be discovered it is imperative that everyone agree on what is time. By using a “mind experiment”, as Einstein, Hawking, Faraday and so many others have done, we can define time. It can be shown that “Absolute Time” exists. And time can be expressed as a mathematical derivation of $E = PE + KE$.

Introduction:

Time is one of the great mysteries that mankind has not resolved. Everyone knows time keeps moving forward and everyone has experienced the results of time, growing older, but few people agree on what time is. Looking at a sampling of research papers on time it seems that science have never come up with scientific definition which is satisfactory to everyone. Einstein made a great effort in his work, “On the Electrodynamics of Moving Bodies”, to define clocks and the definition of simultaneous, but he does not define time. He never does. A clock is a device which measures time, not time itself.

Einstein did not define time, but many scientists, church leaders, philosophers and others have tried. The dictionary defines time as: the system of those sequential relations that any event has to any other, as past, present, or future; indefinite and continuous duration regarded as that in which events succeed one another. Of the many different definitions researched, this one is the closest to what time is.

Time is the motion which creates the events which succeed one another. By using a thought experiment, a conclusion can be reached that time is motion and absolute time is the relative motion of each object to every other object in the Universe.

Time Begins:

The current theory of the creation of the Universe is the Big Bang theory. In the beginning all mass resided at a point; this was the Universe. The Universe condensed to its most compact. It would have no space between stars, planets, atoms, quarks, all mass resides in one place. The Universe was one point. There were two things that existed; this point and space. Time had not begun. (See diagram 1)

The Universe was complete. For some reason, the Universe became unbalanced and exploded with a big bang; some say God did it.
With that “big bang”, the Universe exploded outwards and time began. Time began at the instant there was more than one particle in existence and at least one of them was moving. The Big Bang handled those rules quite easily, creating untold trillions of particles and expanding at a tremendous rate.

Each of these particles was affected by and affected all the other particles around it through gravity, collisions, and energy transformation. At any point since time began, each and every particle of the Universe would be in a unique pattern to each other. The next point, no matter how long you wait or how quickly you select, every particle of the Universe is in a location, many of them would be in a new location. Once again, every particle is affecting every other particle and the Universe would again be in a unique pattern. The movement leading to the “...events (that) succeed one another”, which can be measured by any number of clocks, is time.

**Thought Experiment 1:** Immediately after the big bang, the Universe was expanding. Each particle of the Universe would be in an ever-changing unique relationship to all the other particles. And that relationship causing and resulting from the movement is time. But, it will be much easier to understand time by using an extreme simplification of the Big Bang. (See diagram 2) The Universe contains an unknowable large number of particles at any given point in time, but only 5 objects will be used to represent them all. The series of diagrams 2a, 2b, and 2c have 5 objects in them and are labeled V, W, X, Y, and Z. In each succeeding picture the objects have changed to a unique position in relationship to each other. Each object has traveled to their position effected by all the other objects by forces which alter the courses of the objects. In the diagrams, the objects that reduce in size represent objects going into the paper, the objects growing larger are coming out of the paper. The red lines between the objects represents the distance between them.

The diagrams show succession of unique relationships. Each diagram is a selected point in time. Time itself is the smooth motion producing the change of distance between the objects.

If the objects represented the solar system, one can measure time in the year and day, as the planets move smoothly around the Sun and the Earth rotates.

If every particle of the Universe is included in the diagram you would have a diagram of absolute time. At any point in time, as measured by any clock, the Universe will always be in a unique new pattern.

A very useful tool to understand that “time is motion” is thought proof 1.

**Thought proof 1:**

Imagine everything in the Universe is absolutely frozen, nothing moves… not a planet circles a star, not a person walks, not an atom vibrates, not even an energy wave wriggles…nothing moves. *Time would cease to exist.* Now, imagine things un-freeze, everything just starts right back up from where it was, things are back to normal. Nothing will have changed, no planet has moved, no one would be aware of any hic-up in time and no instrument would have recorded anything so there could be no record of the “missing time”. In fact, there is no missing time, time ceases to exist without motion.
**Thought proof 2:**

Imagine everything in the Universe is absolutely frozen, nothing moves...not a person walks, not a planet circles the sun, not an atom vibrates: except one single ball moving in a straight line. Now time can be deduced, by the change in position of the ball moving relative to the rest of the frozen Universe.

If another ball was unfrozen, so that two balls are moving and the rest of the Universe still frozen. Time could be measured by the change of position of the 1\textsuperscript{st} ball moving relative to the rest of the frozen Universe still. Or time could be measured by the change of position of the 2\textsuperscript{nd} ball moving relative to the rest of the frozen Universe. Or time could be measured by the change of position of both balls relative to the rest of the frozen Universe and each other, this would be absolute time. One could continue unfreezing the rest of the Universe one item after another until the entire Universe was unfrozen. Time could still be measured by the 1\textsuperscript{st} ball moving relative to the rest of the Universe. But “Absolute Time” is the change of position of every object in the Universe relative to each other.

**Independent Research:**

On the other side of the world, Dr. Julian Barbour, PhD. (University of Cologne, 1968) a theoretical physicist came up with the same idea of “Time is Motion”, independently and through a different rational.\(^7\) Einstein pointed out, in his famous equation \( E = MC^2 \), that mass and energy are interchangeable. Dr. Barbour uses this fundamental concept in his approach by using a mathematical expression for the law of Conservation of Mass and Energy.

Dr. Barbour used different symbols than this paper. This paper which uses symbols currently used by the book, *College Physics, 10\textsuperscript{th} edition* by Raymond Serway\(^8\). Dr. Barbour’s argument is as follows:

Bodies with mass have potential energy (PE) in their mutual attraction;

\[
PE = g \frac{m_im_j}{r}
\]

Taking this and applying it to all the bodies of the system, the formula looks like this:

\[
PE_{ij} = g \sum_{i-j} \frac{m_i m_j}{r_{ij}}
\]

Where: \( g \) is Newton’s gravitational constant, \( m_i \) is the mass of the body “\( i \)” in the system and likewise \( m_j \) is the mass of body “\( j \)”, and \( r_{ij} \) is the distance between the objects \( i \) and \( j \). The
The summation symbol means take all the pairs \(ij\) in the system once, calculate the gravitational attraction for each and then add them up.

Objects in motion have kinetic energy, \(KE\).

\[
KE = \frac{1}{2} m v^2
\]

A very good approximation of velocity is \(v = \Delta x / \Delta t\). The delta symbol indicating a very small quantity. Taking that approximation and applying it to all moving objects results in the following equation.

\[
KE_{ij} = \sum_{i-j} \frac{m_i(\Delta x_{ij})^2}{2}
\]

In an isolated system the sum of the kinetic energy and gravitational potential energy remains constant at all times; Conservation of Energy.

\[
E = PE + KE
\]

Substituting into \(PE\) and \(KE\) and taking the isolated system as the entire energy of the Universe \((E_u)\) the equation looks like this:

\[
E_u = g \sum_{i-j} \frac{m_i m_j}{r_{ij}} + \sum_{i-j} \frac{m_i(\Delta x_{ij})^2}{2}
\]

In this form of the equation, time shows up. Electrical fields, magnetic fields and such are manifested through the motion of charged objects and are intrinsically included in the equation.

Rearranging the equation to isolate time on one side results in this equation:

\[
\Delta t = \sqrt{\sum_{i-j} \frac{m_i(\Delta x_{ij})^2}{2(E-PE)}}
\]

Where \(PE_{ij} = g \sum_{i-j} \frac{m_i m_j}{r_{ij}}\)

\(E, \ g\) and the mass are constants, the only variables are time and displacement. This equation says that time is dependent on the movement of every object in the system. A dimensional analysis of the equation resolves to \(\Delta t = \Delta x \text{ second/meters}\). Of course, seconds and meters are defined by man.
The following equation shows instantaneous velocity expressed without using time. Dr. Julian Barbour writes, in *The Nature of Time*, about this equation, “… (it) expresses the truth that only relative quantities have objective meaning. Speed of body “i” is not the ratio of its displacement to an abstract time increment but to …the displacements of all the bodies in the system.”

\[
\mathbf{v}_i = \frac{\Delta \mathbf{x}}{\Delta t} = \frac{2(E-V)}{\sqrt{\sum_i m_i(\Delta X)^2}} \Delta x
\]

**Ramifications:**

What are the ramifications of this concept of time? Author Walter Isaacson, in his article *How Einstein reinvented reality*, writes, “With his special theory of relativity, Einstein had shown that space and time didn’t have independent existences but instead formed a fabric of space-time.”

Einstein’s beautiful equations on relativity are still valid, they have to be. They have been confirmed many times over. But at the edges, relativity fails in several profound ways. Dark energy and dark matter have been “invented” to make those equations work with the observed Universe. Neither of these have been found. Another major unresolved issue is that relativity does not mesh with quantum mechanics. Could these problems be resolved by understanding the true meaning of time?

The concept of Space-time has led to many bizarre areas of scientific exploration through mathematics. They are mathematically wonders of imagination which produce; worm holes, time travel, multi-universes, and other non-provable ideas. It is unknown if space can be bent or not because there is no information from “Space”. The only information that comes from “Space” is actually the particles and energy of the Universe traversing through any given sector of space at any given moment. And time has been misunderstood since man became rational. Space-time has to be thought of differently. Space-time can be thought of as the relative movement of objects through space creates time.

The vast distances of space make the measurement of time relative. A super nova in the night sky is not exploding simultaneously with the observers on earth looking at it. The super nova could have happened 100,000 light-years away. But when the super nova exploded every particle in the Universe was in a unique position. Over the years the photons from that explosion traveled across the vast expanse of space. At any point selected, those photons and every other particle is in a unique position. After 100,000 years the photons from that explosion reach Earth. A person looking up would see it explode. It was not simultaneous it was relative.

A train traveling near the speed of light doesn’t stretch, photons hit the eye in a sequence that makes it look like it stretches, so the brain reads stretch. In absolute time the train is a train. Every particle of that train is in a distinct location to every other particle in the Universe.
the photons that bounce off that train and strike the eye are in distinct positions in the Universe in relation to every other particle including the train. Absolute time can be used to explain this.

Absolute time shows that now is unique and can’t be repeated.

Absolute time can only go forward. For time to go backwards it would require all the particles of the Universe to once again be in a pattern that it previously was in. Due to the unimaginable trillions of particles in the Universe, getting them all back into a previous unique position would be impossible.

Fortunately, mankind does not have to use absolute time, we can’t. The Universe has forces which imparts a regularity to this motion on all dimensional scales. These “laws of nature” have slowly been un-raveled by a series of brilliant scientist, philosophers and mathematicians. On the macro scale, gravity is the dominate force and it accounts for the motion of the earth around the sun (a clock that measures a year), the rotation of the earth around its axis (a clock that measures a day), the tidal pull of the moon and all the other movement and interaction between heavenly bodies. These were the clocks of early man. Mankind’s clocks have improved since then with town clocks in the bell tower, to stand up grandfather clocks, to watches, to digital clocks, to atomic clocks. Modern science and technology need an absolutely precise measurement of time. In 1967, the Bureau International des Pods et Mesures (BIPM: an intergovernmental organization through which Member States act together on matters related to measurement science and measurement standards) defines the second as:

\[ \text{The second is the duration of 9,192,631,770 periods of the radiation corresponding to the transition between the two hyperfine levels of the ground state of the cesium 133 atom.} \]

Mankind has produced super clocks but still does not have a unified understanding of Time.

Conclusion:

Mathematics is the “language” which describes the functioning of the Universe. Whether it is, “there are 5 cows in the field” (5 cows) or “Force equals mass times acceleration” (F=MA) math is trying to describe the real world.

Once a few “laws” were developed, such as Newton’s law of gravitational attraction \( (F_g = g (M_1M_2/ r^2)) \), mathematics became an investigative tool. Neptune was mathematically predicted before it was observed. But mathematics is only a description of nature, not nature. When mathematics and nature disagree it is not nature which is wrong.

Sir Isaac Newton was right with his description of gravity. His equations and calculations worked perfectly fine on earth. But, in the late 1800’s, Mercury was discovered not to be following the prescribed path according to Newton’s concept of gravity. A few scientists of the day thought a planet on the back side of the Sun, dubbed Vulcan, was causing this discrepancy of movement. Of course Vulcan was never found and it was Einstein who revolutionized physics with his equations which account for the small discrepancies in the orbit of Mercury.
Einstein’s equations developed through his work on general relativity are the best that mankind has produced to accurately describe the Universe. They are correct. Except “Dark Matter” and “Dark Energy” have been hypothesized to bring the Universe back in line with the equations. Are dark energy and dark matter the Vulcan of today?

Relativity is alive and well. Only time that has “meaning” to the observer is used. The farmer uses seasons, the worker uses days and hours, and scientists use the duration of 9,192,631,770 periods of the radiation corresponding to the transition between the two hyperfine levels of the ground state of the cesium 133 atom. They just need to measure time. But theoretical physicists must understand time, not just measure it.

Scientists have the best tools they have ever had to study nature, but if science is ever going to develop a *Unified Theory of Everything* then a basic agreement of what time is, must be at its foundation, and time as defined by this paper is a beginning.
Figures page 2:

Figure 2b

Figure 2c
End notes:

1. Ashtekar A. Time in Fundamental Physics. Essay from: Penn State Workshop on Time and Cosmology. 21 December 2013; University Park, PA


The Relationship between Lower Extremity Power and Faster Acceleration Times in Athletes

Shaun Reghabi

Physics 112
Professor Dr. Durandet
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Abstract

The goal of this research was to find the relationship between lower extremity power and faster acceleration times in athletes. Within this paper we dive into some of the basic physics concepts, like that of Newton’s laws of motion, needed to help get a fundamental basis on the application of sprint mechanics and power production within sprinting acceleration. Additionally, the use diagrams were used to help better illustrate these key points throughout the paper. Numerous amounts of studies have been conducted at elite levels of competition to find whether or not specific lower body test and exercises contribute to quicker acceleration times. These studies ranged from use of sport specific exercises/tests and there relations to power output/speed to different linear performance assessments.

Content

The ability to create maximum levels of power output is an essential part in the world of athletics. These high outputs of power correlate to almost all aspects of athletics, whether it be running, jumping, swimming, or etc. Because of this, numerous studies and mass amounts of research have been compiled over the past few decades to help understand how to increase power output in its relation to sprinting in athletes (3, 5, 6, 8, 9). When looking at power output, one can argue that the topic of acceleration will always come up, fundamentally in the world of sports you cannot have one without the other. What comes to mind when we hear the words acceleration? For most, the first thought that comes to mind is a race event whether that be cars or athletes pitted against each other to obtain the fastest time in any given distance. We can define acceleration as the rate of change of velocity per unit of time (1, 12). But, to better understand its application towards athletes and its relation specifically towards power we must define a few key terms.

When looking at acceleration and power in athletes we must have a simple understanding, or a basis of kinetics. Kinetics is the study of forces acting on or produced by an object (1). By having a basis, we can understand the magnitude of force a person produces into the ground, while accelerating, has an equal and opposite reaction on that person’s body. Therefore, kinetics is imperative to the person producing the movements for acceleration (1). The second set of key terms we must understand are scalar and vectors quantities. Conceptually, understanding both of these terms are critical to the fundamental biomechanics of acceleration in sprinting (2). A scalar is a quantity that only has a magnitude, but no direction, i.e. speed, mass, and power (12). Whereas, a vector quantity has both a magnitude and a direction, i.e. acceleration, velocity, and force (12). When taking a look at sprinting and acceleration, two questions always come up. 1. How much force can be applied to the ground? 2. In what direction can that force be applied to the ground to obtain the desired direction we want the body to move? (10, 11). Typically when we are talking about acceleration, we are talking about vector quantities.

Finally, our last key pieces to laying the foundation to understanding the application and relation we must look at Newton’s 1st law of motion, inertia, Newton’s 2nd law of motion, force, and lastly Newton’s 3rd law of motion, reaction. Newton’s 1st law can be defined as an object at rest stays at rest and an object in motion stays in motion with the same speed and direction unless acted upon by another outside force (12). Therefore, if someone is not moving at all, i.e. zero meters per second, and suddenly needs to accelerate, like a 40yd sprint, they need to overcome inertia. Knowing that when we need to break inertia to get moving there are some unique
characteristics of force and power output that allow us to do so (2). Newton’s 2nd law can be defined as the acceleration of an object as produced by a force is directly proportional to the magnitude of that force in the same direction, and thus is inversely proportional to the mass of the object (12). Essential, looking at figure A, Newton’s 2nd law states that force is equal to the mass of an object times its acceleration (A.2). Lastly Newton’s 3rd law states that for every action, there is an equal and opposite reaction (12). Therefore if an athlete is sprinting puts 100 N of force into the ground, there will the same 100 N of force pushing back in the opposite direction in which the initial applied force was being placed. Now, let us look at the relationship between the two in a mathematical perspective, in figure A, we see that power is equal to force times velocity (A.1). Furthermore, in figure A, we see that force is equal to the mass of an object times its acceleration (A.2). By taking those two equations we can find a quantitative relationship between power and acceleration (A.2, A.3).

Now having a basis of understanding, let us take a deeper look at some notable studies that have been conducted over the past decade pinning a link between lower extremity power and sprint acceleration. Research has found that high lower extremity power production test’s like broad jump, 1-repetition max squat, or 1-RM squat, and max vertical jump testing have a “significant relationship” towards 10 yard acceleration times (13). The strongest relation of all power production test’s and sprinting acceleration came between the broad jump and 40 yard sprinting times (13). With 40 yard sprinting, an athlete must generate a great enough force to break inertia and accelerate quickly to reach a maximum velocity early in the sprint in order to post a good time (14). Concurrently, in a broad jump an athlete must generate a great enough force to break inertia and explode of the line to reach the farthest distance possible (14). Within most data found there were some deviation from novice to elite athlete on whether lower extremity power had any relationship towards acceleration times (6, 7). Those novice with lesser skill showed greater relationship of training for higher lower extremity power and quicker acceleration times, than compared to those top level performing elite athletes with greater amounts of skill (7). The reasoning behind this, is because those with lesser skill see greater impacts of training in short amount of time, i.e. dramatic strength gains and power output, compared to those at the highest level who see minimal amounts of impact on training in shorter amounts of time because they’re already at such a high capacity (1).

Even though there is an overwhelming amount of compiled data at hand, this does not always transfer to every single application and in this position to constant prediction of faster acceleration times. Because of this another question arises, will only lower extremity power contribute to faster sprint acceleration times, or are there other contributing factors that aide in the pursuit of quicker times? In the world of strength and conditioning, some of the main goals of a coach are to improve the athletes overall power, strength, speed, mobility, and stability (2). In the regards to faster times, most of these coaches use strength and conditioning programs to help increase overall power output and maximize muscular strength in an effort to improve acceleration and absolute speed (2). If we take look back to an example in figure B we have two hypothetical athletes who are both identical in mass and velocity but, differ in power output. Which one will have the quicker acceleration and thus the faster time? Looking at what we are given, athlete #1 produces 20 Watts of power which is double the amount of power compared to athlete #2 at 10 Watts. Using the given totals we have an acceleration of .2 m/s² for athlete #1 and an acceleration of .1 m/s² for athlete #2. This makes sense seeing that the acceleration of athlete #1 is double that of athlete number #2, because the total power output of athlete #1 is
double that of athlete #2. We can safely assume that in this scenario that athlete #1 will have a faster overall time than athlete #2, because athlete #1 will take a shorter time to reach max velocity. (B.1, B.2).

In continuing using the past scenario of the two identical athletes (B.1, B.2), we know that the one who generates greater power will have greater amounts of acceleration and therefore the faster sprint times. But, how is it that in some real world scenarios we have those who are within the same mass range, yet produce less amounts of power sometimes beat those who generate more power in sprint acceleration times? Is it because of specificity training to sport, or perhaps the use of optimal technique while sprinting? What about neuro muscular adaptation? Looking at previous research, there is not just one specific answer to this question (4, 5). One major contributing factor to faster sprinting times in athletes, is optimal sprinting mechanics (10, 11). The reason why the athlete with the smaller amount of overall power output have the faster times is because the athlete was able to put “the power to the ground” thus using what power they had more efficiently, compared to the athlete with the greater power output and non-optimal mechanics (10, 11).

When looking at optimal sprinting mechanics we must refer back to some of the key terms defined earlier in this paper. Again since we are talking about acceleration, we must deal with vectors. One the most important things when looking at sprinting mechanics is the ability to generate as much horizontal force as possible in the least amount of time while maximizing efficiency (2, 10, 11). Optimal sprinting mechanics involve very critical use of forces and their direction of vectors (1). When we take a look at free body diagrams in figure C, the object moving in figure C.1 has a small vector of force but does manage to get as much of that force in the horizontal axis as possible, being more efficient and therefore covering more distance in a shorter time. Compared to the first free body diagram in C.1, figure C.2 has a greater force but, will not move as efficient because the vector in which the force is pointing is not in an optimal position, therefore will take longer to cover the same amount of distance. So, to have quicker acceleration times in athletes you must not only have lower extremity power, but also optimal sprinting mechanics.

Conclusion

After researching through many different articles and different books I have come to the conclusion that these findings showed a very strong relationship between lower extremity power and acceleration times in athletes (5, 6, 7, 8). Looking over the present data I had within this paper, I found that those with exercises and tests with the similar direction in which the force was being produced as sprinting i.e. broad jump, had a stronger relation of power to quicker times ratio than those exercises producing forces in different directions i.e. vertical jumps and 1-RM squats did for power to quicker times ratios. All exercises and testing’s had a strong relationship between the two, but it was most notable that exercise imitating the same direction of movement had a stronger relation. For those who are not in the strength and conditioning field, you can see that specificity of training will have some of the greatest impacts on results. That being said those coaches and athletes looking to gain faster times in sprinting, can benefit from doing lower extremity workouts focused more towards horizontal force production. Also, both coaches and athletes can benefit from application of optimal technique by helping put that new found power down to the ground more efficiently.
My final thoughts to this paper, is that I had great amount of findings from multiple different studies, but I thought to myself that looking throughout the data there was not enough information pertaining more recreational based level athletes. There is way more than enough information on elite and amateur players, but those players only fit a small niche compared to the vast amount of recreational athletes. I feel that hopefully in the future more studies will be conducted on this matter in order to benefit the greater number of people playing sports.
Figures

A.1) $P = F \cdot V$
A.2) $F = m \cdot a$
A.3) $P = m \cdot a \cdot V$
A.4) $a = \frac{F}{m \cdot V}$

B.1
$m = 10 \text{ kg}$
$v = 10 \text{ m/s}$
$P = 20 \text{ Watt}$
$a = \frac{v}{P}$
$a = \frac{20 \text{ N m/s}}{10 \text{ kg} \cdot 10 \text{ m/s}^2}$
$a = 2 \text{ m/s}^2$

B.2
$m = 10 \text{ kg}$
$v = 10 \text{ m/s}$
$P = 20 \text{ Watt}$
$a = \frac{v}{P}$
$a = \frac{20 \text{ N m/s}}{10 \text{ kg} \cdot 10 \text{ m/s}^2}$
$a = 1 \text{ m/s}^2$

C.1
F
N
mg

C.2
N
mg
F
References

Human Anatomy and the Application of Physics: Levers and Center of Gravity

Brian Rellihan
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PHY 112
Professor Swingler
Abstract:
Kinesiology is the study of human movement with consideration given to physiological and mechanical actions based upon evolution. The structure of our anatomy is a complex coordination of skeletal structures (lever systems) that produce work. That is, muscles provide force in conjunction with joints to move over a range (distance) that is related to the length and alignment of the skeletal system. The range and work can be further limited by moving center of gravity (COG). The purpose of this paper is to understand body mechanics based upon principles of physics, specifically levers and COG, when considering anatomical structures in a human body that make up the lever system.

To understand the function of COG and levers within human anatomy we need to not only cover the underlying physics but also the anatomical orientation and construction of the human body. Specifically focusing on musculature system, and joints, which form the mechanical systems that are the basis of this paper.

With over 600 muscles in the human body it is amazing to think that they are coordinated to work in conjunction with each other for us to walk, talk, lift, and breath, just to name a few. Each perform specific functions to serve a physiological need. Through their movements our bodies apply principles of physics in the areas of leverage (lever systems) when applying force over a range, and adjusting for moving center of gravity (COG).

We, and all vertebrates, have three muscle types: Skeletal (voluntary), Cardiac (in-voluntary), and Smooth (in-voluntary).

- Voluntary skeletal muscles are attached to skeleton, appear long and striated, being composed myoblasts (hundreds), fused together, end to end. These movements are controlled by somatic nervous system. We can see these visible in biceps, or quadriceps when an individual has a low body fat percentage (males approximately 8% or less). This paper will focus on skeletal muscles.
- Cardiac muscles are located in the heart and also appear striated, yet composed of smaller cells. They are an examples of involuntary muscles (as with smooth muscles), which do not require deliberate thought. These muscles are run by the autonomic nervous system. An example is the heart, no thought is required to keep it beating.
- Smooth muscles are inside organs, other than the heart, and needed to produce slower contractions. They can be found in out gut, pupils, and blood vessels.

Skeletal muscles are attached to the bones with tendons, connective tissue made up of very durable collagen fibers. They are very strong, typically maintain a tensile strength greater than most soft tissue. As with most structures in the human body muscles and tendons have also evolved to function optimally in our environment(s). In recent history the palmaris longus has been studied. This small muscle and adjoining tendon in the wrist has been noted absent in 16% of subjects in a 2001 study. The palmaris longus is good for gripping, and observed in many climbers. This is not a necessity as we no longer need to hunt for our food, flee from predators, therefore it is being selected against, evolving away from that trait.

Charles Darwin introduced the idea of natural selection to reflect diversity and the selection of preferred traits, resulting in our current human anatomy and physiology. Geoffrey Miller,
evolutionary psychologist is quoted, "Parents could basically choose which sperm and egg get to meet up to produce a baby based on genetic information about which genes contribute to which physical and mental traits," 5 These physical attributes make us larger, stronger, and faster which are directly related to the physiology of lever’s in the human body. 8

Joints, the pivot points in the human body, play a critical role in the fluidity of movement when discussing the lever system. 1 Joints are organized by their flexibility, and fall under three categories: diarthrosis, amphiarthrosis, and synarthrosis. 1 Diarthrosis, sometimes called synovial joints, move freely and provide the least amount of friction in human lever systems. 1 They tend to be much more complex and encapsulated with fluid (synovial fluid). 1 The ends of the bones are capped with a very smooth cartilage Amphiarthrosis move slightly and would add resistance, therefore require additional force to move resistance. 1 We can see these represented between the bones in our vertebral discs (spine). Finally the synarthrosis do not move, and aren’t considered when opposing resistance when exercising. An example would be the sutures, collagen structures, which hold the cranium bones together. 1

Not impermeable to defects and damage, joints (pivot points) suffer from arthritis and other inflammation. Diseases can attack not only the specific joints but the surrounding encapsulated area. It tends to be more prevalent in the hands, shoulders, knees, neck, and lower back. In some cases additional fluid is pumped into the area to supplement the synovial fluid, or more evasive surgical procedures that can replace entire joints.

Archimedes of Syracuse, 287 - 212 B.C., a scientist, inventor, engineer, and mathematician, is quoted, “Give me a place to stand, and I shall move the Earth”. 3 It is believed that he was referring to The Law of the Lever. Ahead of his time, he postulated that the magnitudes are in equilibrium at distances reciprocally proportional to their weights when you have balance. 3 [See Figure 1.] A lever system is a rigid rod (or bone in the case of a human Kinesiology) that rotates on a pivot point (or joint) to apply force at a different point. These systems are composed of two forces, an effort force and a resistance force, acting on a fixed pivot point. The benefit of a lever is to create a mechanical advantage, the amount by which the system multiplies the force put into it. This system is represented by the equation: \( W_1 \times D_1 = W_2 \times D_2 \). 2 Another way to consider this equation is when balance occurs the moments are equal and opposite. 2 A moment is the distance time the force expressed as \( M=(F)(d) \) and occurs when the force applied acts upon an object and it begins to twist. 2

The longer the distance the lower the required force and vice versa. This principle is based upon the force applied as compared to the force expressed. A lever system elevates input force to create a larger output force or distance. The benefit of a lever is to create a mechanical advantage, the amount by which the system multiplies the force put into it.

Based upon where the pivot point is the mechanical advantage can be greater or less than one. To calculate the mechanical advantage ratio you divide the output by the input. If the resistance arm is longer than the force arm then the mechanical advantage is less than one. Whereas if the resistance is less than the force arm, then the mechanical advantage is greater than one.
Class 1 Levers have force applied to one end, axis in the middle, and the load is at the other end.\textsuperscript{8} A good example would be a seesaw where people sit at both ends and the weight from the heavier individual (force downward), would cause the lighter person on the other side up in the air (load). The axis could be moved towards the heavier individual, thus shifting the amount of force downward needed to lift the load side. In humans, the Splenius capitis muscle (back of the neck muscle), acting across the Atlanto occipital joint (neck, pivot point) to balance the head, much like a teeter-totter with the force on the ends.\textsuperscript{1} [Figure 2]

Class 2 Lever places the resistance in-between the pivot point and the force applied, and the mechanical advantage is always greater than 1.\textsuperscript{8} A common example would be a wheel barrel. Lever in the human body can be seen in the plantar flexion in the foot at the ankle where Metatarso-phalangeal joint (pivot point), the resistance is the center of gravity and the force is the pull from the Gastrocnemius and soleus (calf muscles).\textsuperscript{1} [Figure 3] There needs to be a force then to overcome body weight.

Class 3 Lever places the applied force in the middle and the pivot point on opposite ends with the resistance. Mechanical advantage on these levers is always less than 1.\textsuperscript{8} Tweezers and scissors are good examples. This lever system is the most common lever in the body. The applied force is in between the pivot point (joint) and the resistance. An example is the bicep muscle pulling up on resistance and the scapula/humerus (shoulder/upper arm) are fixed in place as well as the antebrachial (forearm) and metacarpals (wrist) are fixed. The effort in this case is located between the spinovial hinge (elbow), pivot point, and the resistance, where the muscle attaches to the ulna.\textsuperscript{1} [Figure 4]

Finally the rods in the lever system are represented by a skeletal system. The human body consists of 206 bones.\textsuperscript{11} It provides structure, protection, produces blood, and stores calcium. Bone has a dense and strong outer layer encapsulating a spongy marrow where blood is produced. When inside the body, being fed nutrients bones have a bit of a flex, albeit limited. This needs to be considered when referring to bone as the ‘rod’ in a lever system. The lever is expected to maintain consistent, lacking any flexing when balancing forces. If the contraction of the bicep curl was re-visited and the ulna pulled up (Class 3 Lever) was considered the flex, even slight would throw off the calculation of forces applied and resistance. That being said the basic fundamentals of levers still apply and result in a mechanical advantage.

There are different between males and females skeletal systems. For example a females pelvic region is at an angle of 100 degrees of rotation from ground verse a males which is less than 90.\textsuperscript{11} This has an impact on posture and therefor COG (when moving) for the sexes. Later differences on COG between males and females is discussed in relation to body composition and structure.

When addressing levers and the human anatomy center of gravity (COG) needs to also be considered. Processed in our bodies we have the ability to shift or create small movements to counter COG. It is another example of an involuntary system in most cases. COG was also discovered by Archimedes when he was working on the concept of levers.\textsuperscript{8} COG is the point of an object where the weight is concentrated and will balance. For simple geometric shapes it occurs in the middle. If an object is at rest is it considered at equilibrium, and the sum of the
forces acting upon it are zero. When an object takes a shape that is non-uniform the COG will be
affected. Stability is also affected by the size of the base along with the COG.\textsuperscript{9} The smaller the
base the easier for a shift to occur in the COG. Relating back to levers, an object will balance
when the forces acting upon either side of the COG are equal. This is also dependent on the
location (distance) of the forces in proximity to the COG.

The average humans COG is 3cm below the Umbilicus (navel) when sanding straight up.\textsuperscript{11}
Females tend to be lower as they have a disproportionate amount of weight in their lower to mid-
section. This place on the body would be considered a balance point upon which the body acts
to equalize the forces on either side of the COG. Base support, where the feet are positioned, in
most cases helps maintain equilibrium. Positioning the fee one in front of the other, a walking
position, creates a base with greater resistance to forward and rear forces. Whereas, placing the
feet side by side shoulder width apart, is more protective against lateral forces. [Figure 5]

Height of the COG changes also with body positioning. As the COG moves closer to the support
base, the greater the angular displacement that is possible. An individual standing straight up
with feet close together can easily be pushed of his base; COG easily shifts from side to side.
This is why athletes use a stance that focusing on bending he knees and dropping their hips.
While the vertical line of where the COG lies on the body is the same, repositioning the lower
extremities makes the change on the horizontal plane. Such a positioning creates a lower COG
with greater stability. [Figure 5] This can be seen in football lineman, they begin in a crouched
over stance with the COG between their hands feed. As the ball is snapped they propel
themselves forward, like a runner, shifting their COG in front of them expecting a collision and
keeping a low center of gravity but focused on pressure from the front. If the opposing lineman
simply let the by, they would fall as the COG is to far beyond their orien-tation.

When you take the mass of a human and drop a line down the person if the weight is equally
distributed between the support, the feet. If weight is shifted to one side of a person then the
center also shifts accordingly. This changes when there is movement. For instance when a
person is running the COG is in front of the legs, propelling them forward. Should they stop in
place, with the COG in that position, in front of them, nothing would prevent them from falling.
It can be said that to move forward, specifically running, is a series of falls with the expectation
that the next step will prevent a collapse. Then as deceleration occurs the COG shifts back from
in front of the feet to in-between the feet in line with the abdomen.

Through the study of COG one can break down what resistance a movement is placing on the
body, and on specific muscles. Through the shifting the COG there is a change in the muscles
that are being emphasized. As mentioned prior, running which changes the angle of the body
emphasizes different muscle fibers than walking. From a therapeutic standpoint, it allows for
individuals to not only move in deliberate ways but limit use to areas on the body that are prone
to injury or already injured.

Musculoskeletal disorders (MSD’s) are the result, weight management, overuse of joints and
muscles, in conjunction with poor body mechanics. Most commonly, due to improper body
orientation when doing daily activities.\textsuperscript{9} Work related MSD’s are the leading factors in missed
work days and illness. In 2011 MSDs cost businesses 50 billion dollars, accounting for 30
percent of all employee compensation.\textsuperscript{9}

Sprains, tears, and strains can be prevented by proper ergonomics and managed movement in bending and lifting. Ergonomics is fitting a job to the individual by focusing on posture; specifically positioning the joints for repetitive movements.\textsuperscript{9} Added attention is given to the back, neck, and head areas. Those places on the body have joints that have the aforementioned, diarthrosis, amhiarthrosis, less flexible joints.\textsuperscript{1} Due to the lack of pliability and increased weight of the general population the is increasing pressure put on all joints. The twisting and bending easily damages the connective tissue, or cartridge. Best case scenario only resulting in inflammation and irritation but more likely tears.

To limit injury and prepare joints, ligaments, and muscles for activity stretching provides a good first step. It not only warms the areas of the body but increases range of motion (lever) but enhances blood flow and nutrients. Additionally it temporarily elongates muscle fibers and joint fibers. With elongation it makes it harder for and injury to occur, but should it happen, recovery is quicker as there is already a good environment of blood and nutrients.\textsuperscript{7} Post stretching after an activity also helps circulation and healing. Activities that promote muscle engagement actually create small micro tears. These tears heal and create a stronger, more durable muscle. A resulting condition is Delayed Onset Muscle Soreness (DOMS).\textsuperscript{7}

Obesity is a body mass index (BMI) of over 30, with greater than 1/3 of U.S. adults falling into this category and 17\% of adolescents.\textsuperscript{12} It is seen to impact with disregard to socioeconomic considerations or ethnicity. Causes range from decreased activity, increased caloric intake, that are based upon personal choice. Additionally, as the body ages, hormone levels naturally decrease, specifically testosterone, growth hormone, and, dehydroepiandrosterone.\textsuperscript{13} With modern medicine hormone levels can be checked and, if prescribed, elevated to “normal levels”. In many cases, a healthy lifestyle that includes healthy food choices and daily activity to elevate the heart can quickly improve conditions.

The weight of a person impacts their center of gravity and lever system. A study from Ohio State University and Colorado Behavior Risk Factors Surveillance System tells us that one in five (27\%) adult obese males and one in five (21\%) obese females had injuries.\textsuperscript{12} As compared to 17\% average males and 12\% average females.\textsuperscript{12} The additional weight not only puts extra stress on the joints but limits range of motion. Also, changing from one posture to another, moving, shifts COG. For a larger person this can lead to imbalance, and in some cases falling. Sometimes this can be used as an advantage, per the example give of the football lineman. For most people though, excessive weight and COG is a problem that can lead to greater issues.

The professions of Occupational Therapy and Physical Therapy deal with these challenges. In the case of Physical Therapists (PTs), they teach people to prevent or manage their physical challenges so that they will achieve long-term health benefits. Their techniques to promote the ability to move, restore function, reduce pain, and prevent disability. Additionally, PTs work to prevent the loss of mobility prior to occurrence by developing programs suited to an individual’s range of motion, and lever system.\textsuperscript{10} Occupational Therapists (OTs) focus specifically on daily tasks, range of motion and motor skills, due to impairment or defect.
As for weight management, Board Certified Nutritionist work with individuals to encourage health by promoting quality food choices. In some cases they will work in coordination with a PT or OT to provide a comprehensive plan for clients.

**Conclusion:** The human body has evolved over thousands of years to function as a coordinated system. The resulting movements are founded in principles in physics. Kinesecology, is the study of how the mechanical advantage is based upon a system of muscles, joints, and bones in the body. Voluntary muscles cover the human body and deliberately activate in response to needs or stimuli. They provide the necessary force to act upon the lever system and counter a resistance (force). Joints are pivot points, or fulcrum, on which bones connect and have varying amounts of pliability to flex due to the cartridge that binds them. The skeletal system, specifically bones, are the rod in the lever on which the forces act. This knowledge, specifically COG and levers, provide guidance as to the limitations of the current anatomical structure which is the root of MSD’s. Therefore ergonomics can work with same standards to improve current injuries and prevent potential hazards. Physical Therapist, Occupational Therapist, and Nutritionist assist in the rehabilitation and prevention of injuries by applying the same theories. Regardless, it is the responsibility lies with the individual to limit the opportunity for injuries through stretching prior to activity for increased circulation and range of motion, and weight control to limit the additional forces on the joints.
FIGURES

Figure 1.

\[ M_1 \times a = M_2 \times b \]

Figure 2.

First-class lever. The applied force and the load are on opposite sides of the fulcrum.

Figure 3.

Movement completed.
References:


10. Staff writer for APTA. Available at: http://www.apta.org/AboutPTs/


The Physics of Hemodynamics

Kiersten Rheinschmidt

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Dr. Casey Durandet
ABSTRACT

Hemodynamics is the study of how blood flows through the body. Starting with the electrical pump of the body that follows the combined gas law, the body’s capillaries which follow henry’s law, to the valves in our legs that help fight gravity, and last how the total fluid energy of the body follows the first law of thermodynamics.

INTRODUCTION

The human body is incredibly complex electrical system and still has many secrets. But science has helped solve many of its mysteries one of which is how blood flows through the body, this field of science is called hemodynamics. Part of hemodynamics has to deal with the forces the heart has to go through while pumping blood and are described by blood pressure and blood flow. Hemodynamics involves an incredible amount of physics, from total fluid energy to the ability for blood to pump back up a leg. The study of hemodynamics has aided in understanding, curing and treating diseases such as congestive heart failure and hypertension.

One of our major organs, is the heart, without it we wouldn’t function, it carries our antibodies, our blood, oxygen, and even body waste to and from every tissue in our body. The heart beats with a frequency of about 1Hz which is about a minute per circulation. The valves in the heart control the direction the blood flows and are controlled by the pressure in the two ventricles. The heart follows the combined gas law where when the heart contracts the volume will decrease which in turn causes the pressure to increase. Pressure can modulated by three different variables being intravascular system, inotropy (contractility), and vasoactivity. The amount of work done by the heart can be calculated by the formula 1.

Blood vessels are a closed system of tubes that transported blood throughout the body where arteries carry oxygenated blood and veins carry deoxygenated blood. The smallest forms of these vessels are capillaries which are just big enough to allow for gas exchange. Henry’s law plays a role in this gas exchange because he states that each gas has its own partial pressure, this is why we can breathe the air which is mostly nitrogen without dying from nitrogen poisoning, and nitrogen is only soluble into our blood stream in extreme pressure situations such as scuba diving. When using the term blood pressure with describing veins and arteries it means the pressure put on the walls of the vessel not the pressure of the blood flowing though them. Vessels have elastic type walls which allows for the ability to take on various blood pressures. The size of the vessels plays in the velocity the smaller the radius of the vessel the lower the velocity lowest in the capillaries. The velocity and area help determine the rate of flow as shown in formula 3. The size changes through the vessels also adds in some resistance which can be calculated by formula 4, this explains that the resistance can be calculated by the change in pressure divided by the rate of flow.
Once you get into the cardiovascular system of the lower body veins have to start to go against gravity, the heart alone can’t cause the leg to pump the blood back up to the heart again. Blood is able to go back up the leg by muscle contraction and relaxation coupled with valves that are throughout the veins. The valves prevent the blood from flowing in the wrong direction and the muscle contraction is what allows for the valves to be opened and the blood to flow up, when the muscle relaxes again the valves close back up. A common disease associated with this blood flow is Chronic Venous Insufficiency (CVI) which is where your legs no longer have the ability to pump enough blood back up the leg. This is caused by high blood pressure (hypertension) Like mentioned before blood pressure is the pressure put on the walls of the vessels with this high pressure inside of the vessels the valves that control blood flow can be damaged and swelling in the leg will begin to accumulate. Diseases like this are why it is so important to understand the physics behind the blood flow, by watching the blood pressure in the body damage and disease can be prevented.

When it come to the total fluid energy of the blood in the body it follows the first law of thermodynamics, energy is not destroyed or created. Going through the vessels the fluid energy flows from high to low. Remembering that energy is not created or destroyed the total fluid energy of the body is higher than the highest fluid energy level. This is because the total fluid energy is made up of a sum of all of the fluid energies from the varying vessel radii. The Bernoulli effect comes into play here as well, it goes along with the law of conservation of energy. This effect is where fluid pressure is decreased where velocity is increased this is because the pressure energy is what allows for the kinetic energy to be made.

CONCLUSIONS

What is cool about this subject is that it applicable to just about any health sciences field, whether it is for a human doctor a veterinarian, or even a physical therapist, understanding blood flow can be very useful information. Hemodynamics plays into many cardiovascular diseases such as hypertension (high blood pressure) and congestive heart failure, understanding hemodynamics and its regulators to help cure treat and understand some of those diseases.

The field I have always wanted to go into is veterinary science, specifically equine science. Just like human’s horses can have heart disease just the same, some of their hemodynamic mechanisms are even more complex than in humans. When it comes to the horses leg it doesn’t have the same muscle mechanism as a human to pump blood back up through the leg, instead it uses the hoof. What is called a hydraulic cushion is formed in the hoof, this is where various structures in the hoof are compressed down when the hoof is stepped on, this locks blood into the veins in the foot and forms the hydraulic cushion, when the hoof is no longer being stepped on the blood is released and shot back up the leg. Looking at formula 4 again if you think of it as a high pressure state and the blood is locked in place therefore the flow is null the rate at release is much larger than if there was an initial flow. Once the blood has been shot back up the leg though the horse leg has the same mechanism as a human’s leg and uses valves to keep the blood flowing in the correct direction.
FORMULAS

1. $\Delta W = \int Fdr$
2. $P_g = hpg$
3. $F = vA(1/t)$
4. $R = \Delta P/F \text{(mmHg/ls)}$
5. $-E = P + \rho gh + \frac{1}{2} \rho v^2 c$
6. $P_1 + h_1 = P_2 + h_2$

FIGURES

Figure 1: This chart shows that pressure decreases the farther away from the heart.
CITED REFERENCES


3. Nave R. **Date of Publication**. Bernoulli Equation.[cited 11/1/15]. Available from: http://hyperphysics.phy-astr.gsu.edu/hbase/pber.html


The Physics of a Classic Yo-Yo

Name: Bryan Romero  
Date: April 20, 2016  
Course: Physics 112  
Professor: Michael Swingler
Abstract

A childhood classic toy known as the yo-yo has been around for many years. The fundamental properties of the object makes it possible to go up and down an attached string to the center of mass, the axle, due to the potential and kinetic energy. In order to calculate the acceleration of a given yo-yo heading toward Earth, the momentum and inertia must be understood as they refer to the mass and velocity of the yo-yo at rest and in motion. From there it can be given that the velocity changes as the yo-yo approaches to and away from the axle, which is referred to as the acceleration. Newton’s Second Law of Motion then makes it possible to explain the force applied to the toy being equal to the mass multiplied by the acceleration. Additionally, due to the rotational motion of the toy’s circular shape, torque (T=I(α)) can be used alongside F=ma in order to isolate acceleration as a variable. By doing so, the result is that the acceleration is equal to two-thirds the acceleration because of gravity. Furthermore, by modifying the axle to the toy, kinetic and rotational motion will allow for the device to spin at the axle until centripetal force weakens and requires the tug of a string to return the device to the user.

Introduction

The yo-yo has been around for many years, with it going as far back as 500 B.C. in Greece (Lucky’s 2016). Fast-forwarding to present day, people can find variations of the classic toy on store shelves, including those with built-in lights and other modern mechanisms that add to the basic functions of the device. While observing an individual operate it is all it takes to understand how to play with the toy, knowledge of entry-level physics makes it possible to fully appreciate the simplicity of the toy. By analyzing the functions and forces acting upon the object, the acceleration of the yo-yo toward Earth can be calculated as it rolls down its attached string. More specifically, comprehension of the individual aspects of rotational motion, torque, momentum and inertia all come together to tell the mathematical story of the yo-yo.

The Basics

The standard yo-yo design consists of two circular pieces connected at the center with a component known as the axle. Attached to the axle is a piece of string wrapped around several times, with a loop at the end for the user’s middle finger. Once the user is ready to use the device, they’ll rotate their hand and release the toy with a slight downward throw coming from the wrist. As the yo-yo rolls down the string it will eventually reach down to the axle, where the operator will then lightly pull back up on the string, causing the toy to roll back up. Once the individual has collected the wound up toy, it is ready to be used once again.
Energy

With the basic knowledge of the toss of a yo-yo, the steps can be broken down to: the object being at rest, rolling down the string, reaching the axle, and going back up until the point of return. As the yo-yo is being utilized, energy goes back and forth between kinetic and potential (The Science 2016). At the beginning phase of using the toy, it’s in the users hand and above ground. Due to the positioning of the object, it has stored energy in the form of potential energy, because of the gravitational pull that occurs when it’s released. However, if the yo-yo started on the floor then there wouldn’t be potential energy because it would already be at zero. Being that it’s all in relation to the positioning of the item to the ground, the length of the string plays a significant role due to it setting a limit on the distance the yo-yo can fall until it reaches the axle.

During the process of the toy rolling down the string, the potential energy translates into kinetic energy until it reaches the axle and wraps around the other direction. A pull on the string causes this upward direction, with kinetic energy still being able to be observed until it once again returns to the potential energy point in the user’s hand. For a visual representation of kinetic and potential energy, refer to figure 1 in the figures section of the paper.

While this mechanism occurs, air resistance reduces the amount of energy being released as it moves down, up, and rotationally. Additionally, the friction of the string and axle take away some of the energy as it’s in the process of rolling up and down the string. It’s important to note these variables because the yo-yo isn’t being utilized in an isolated system where air resistance and friction are negated.

Momentum and Inertia

Momentum and inertia play a huge role in the process of understanding the physics of the toss of a yo-yo. The momentum can be altered by changing the mass of the toy, force in regard to velocity, and by changing both of those variables together (Momentum 2016). The mass of a given toy will affect the velocity and ultimately the momentum, as the heavier a yo-yo is the more momentum it has. A stronger throw of the yo-yo will also cause the momentum to be much more due to the velocity increasing. Likewise for the contrary, a weaker toss and lighter yo-yo will have less momentum in comparison.

Additionally, once in motion inertia helps explain the predictability of the object staying in motion once it begins to roll down the string. As the simple definition of inertia is resistance to change, friction and other opposing forces results in lowering the overall acceleration, and the yo-yo experiences inertia (Propulsion 2016). Once the friction of the axle and string occurs, the pull of the string results in a change of direction of the momentum and inertia of the yo-yo. Due to this, an observation of the yo-yo rolling back up the string to its original starting position in the user’s hand will occur.
Acceleration, Torque, and Newton’s Second Law of Motion

The velocity of the yo-yo as it rolls down the string doesn’t remain the same. Simply put, the change of velocity is known as acceleration and it increases as it gets closer to the center of mass where the axle is located. Likewise, the acceleration changes as it goes up the string. Another way of viewing acceleration is imagining a car that is driving, and they speed up to merge into another lane. Being that the driver changed the speed and direction of the vehicle, they accelerated.

Newton’s Second Law of Motion explains this occurrence, as the forces applied to the mass of the toy results in a change of acceleration over time. This is where the equation: force equals mass multiplied by acceleration, becomes useful (F=ma) (Newton’s 2016). When the yo-yo is tossed, the forces are out of balance because a force was applied to it. Additionally, the mass of the object and gravity act upon the yo-yo as it falls, which adds to the acceleration over time.

With this in mind torque comes into play as well because it deals with the twisting force of the object. Torque is calculated as force multiplied by the lever arm perpendicular to the center of mass, or moment of inertia by the angular acceleration (as explained in the next section) (What is torque 2016). Another way to envision torque is by referencing the common see-saw seen at a local park. The lever arms can be considered the children on each side of the see-saw, and the center makes it possible for the kids to go back and forth. However, if one of the kids sits closer to the midpoint, the other will raise them up more easily because of the torque being reduced. With this in mind, the string of the yo-yo is the lever arm in the function of the yo-yo, which comes into play when looking into the process of rotational motion.

Rotational Motion

Given that the shape of a yo-yo is circular, rotational motion can be used to trace the exact amount of distance the object moves over time. The distance is recorded based on the angle (theta), which the term angular is used to define a point of reference. In order to observe this, an imaginary perpendicular line from the center of mass (the axle) to the outer edge is envisioned, where point X will be made. Once in motion, point X will be moved to a different location and the angular distance (represented with theta) should be recorded. Additionally, angular acceleration (known as alpha) can also be calculated by considering the amount of space X has been displaced in a given amount of time. In short, angular velocity (omega) can be used as the distance angle theta has moved over time. Furthermore, angular acceleration (alpha) is used as the change of angular velocity over time (Ilectureonline 2016).
Acceleration of a Yo-Yo Towards Earth

In order to make an educated prediction of the acceleration of a yo-yo heading toward Earth, observation of the two major forces acting on the object must be made. The downward gravitational pull of the object rolling down the string, and the mass of the toy contribute to the net downward force. It’s important to note that the downward force is considered to be positive because they’re contributing to the acceleration. On the other hand, the tension of the string results in an upward negative force due to resisting some of the acceleration.

Newton’s Second Law of Motion tells us that the net force is equal to the mass multiplied by the acceleration (F=ma) of the yo-yo. However, being that the yo-yo is rotating, torque (T=I(alpha)) will be used alongside the given law to make the calculation. With torque being equal to the moment of inertia multiplied by the angular acceleration, we can set the equations equal to each other and find the acceleration which will be done in the following section.

The Logic

Setting up the F=ma equation to T=I(alpha) to be equal to one another is relatively straightforward. With some algebra, linear acceleration is equal to the radius of the yo-yo by angular acceleration, alpha (a=R(alpha)). Additionally, alpha can be written as acceleration divided by the radius of the yo-yo (alpha=a/R). Furthermore, being that the yo-yo is a circular disk, inertia will be equal to one half the times of mass multiplied by the radius squared (I=1/2mR^2). By filling in each variable for T=I(alpha) with its broken down components, T*R=(1/2mR^2)(a/R), the R’s cancel out which leaves T=1/2ma.

Filling in the variables for F=ma also makes it possible for us to find acceleration, as the net force can be filled in as mg minus T (F=mg-T). Which from there, we can fill in the rest which ultimately results in: mg-1/2ma=ma. Now by adding 1/2ma to the right side of the equation, we isolate the equation to equal: mg=ma+1/2ma. Due to this, we can multiply to both sides of the equation 2/3, and cross out the mass (m) to get: g=3/2a. From there, we can reduce all the way down to a= 2/3g, which is two-thirds the acceleration due to gravity itself (Physics 2016). Refer to figure 2 in the figures section of the report for visual representation of the calculations.

The Sleeping Yo-Yo and Other Modifications

Yo-yos today have a much greater variety than they did many years ago, as people can now choose a toy that has built in modifications. These new modifications can also affect the acceleration toward Earth, as they often add more mass to the object. As an example, some yo-yos have lights that turn on when in motion in order to create a visual effect as it rotates. Not only are lights added to each side of the yo-yo to enhance the user’s experience, but it also prevents the toy
from being out of balance. The end result is that while additions can be made to the classic toy, the physics of the device still need to be considered in order to make it functional.

Centrifugal and centripetal forces are the major components of physics that the newly modified auto-return and clutch yo-yos utilize in order to create its desired effect. Centrifugal force is caused by rotational motion at high speeds, resulting in a force going toward the outer surfaces of the yo-yo. On the other hand, centripetal force results in an inward force (Centripetal 2016). Modifications to the axle of the yo-yo can also create an interesting effect that is known as making the toy sleep. As the user makes a toss of the toy, the yo-yo rolls down the string as it normally does. However, once it reaches a newly added spindle located at the axle, the yo-yo continues to spin instead of requiring a tug of the string to return the toy. As the yo-yo loses energy over time, the centrifugal force from spinning weakens and centripetal force takes over, causing the clutch to lock the spindle in place in order to engage the necessary upward force. The additional weight adds to the overall acceleration and centripetal force, which can often result in the sleep time going well over thirty seconds.

Conclusion

The classic yo-yo has been around for many years, creating many great memories for kids and bringing back nostalgic feelings to adults every day. While modern modifications can be added in order to make the yo-yo more interesting, the underlying physics of each toss remains the same. Due to this, these additions build upon the foundation in order to remain functional. From the downward force that acts on the yo-yo, to friction against the axle and string running past each other, every variable effects the acceleration of the yo-yo as it heads toward Earth. Even further, each component makes the physics of each toss possible and tells the mathematical story of the yo-yo.

I believe that the yo-yo will continue to be around for many years to follow for various reasons. Due to the toy being relatively cheap and easy to produce, it makes for a great choice for fun. Even further, the standard yo-yo can teach us a lot about physics as it has several variables that contribute to its function. Due to this, I believe it’ll always serve as a learning tool. As each component is analyzed and fully understood, those lessons can be related to other situations that contain the same functions. Even further, by understanding the yo-yo on a smaller scale, future inventions could result in the yo-yo being as a foundation of their research.
Figures

Figure 1: Potential and Kinetic Energy

Figure 2: Calculations
References


The Impact of Physical Training and Biotechnological Advances on Prevention and Management of Diabetes and Diabetes-related Complications

Author: Parastou Sazegar
Date: 11/19/2015
Course: PHY112
Section Number: 12834/12901
Instructor: Dr. Casey Durandet
Abstract
As diabetes continues to be known as the top five leading cause of death in the United States, the importance in understanding the complications of this disease grows rapidly. National diabetes Statistics reported 29.1 million people which corresponds to 9.3% of the U.S. population, diagnosed with diabetes in 2014. (1)
“According to the Centers for Disease Control and Prevention, in 2007, almost 24 million Americans had diabetes, with one-quarter of those, or six million, undiagnosed” (2)
If not properly managed, diabetes can cause head-to-toe damage. Untreated diabetes leads to complications such as cardiovascular disease, nerve damage or neuropathy, kidney damage or nephropathy, eye damage, foot damage, hearing impairment, and Alzheimer’s disease. (3)
Although diabetes cannot fully be treated it can however be managed. This article focuses on the impact of physical training and biotechnological advances on prevention and management of diabetes and diabetes-related complications.

Diabetes classification
Diabetes mellitus is a metabolic disorder that causes elevated glucose level in the blood or hyperglycemia and abnormal carbohydrates, protein, and fat metabolism. Based on the present or absent of insulin and proper functioning of its receptors there are currently five categories of disordered glucose homeostasis:
- Type 1 diabetes “insulin-depended diabetes”
- Type 2 diabetes or “non-insulin depended diabetes”
- Impaired glucose tolerance/impaired fasting glucose
- Gestational diabetes
- Other rare forms such as Maturity-Onset Diabetes of the Young (MODY) and pancreatic diseases (4)
In type 1 diabetes there is a total lack of insulin. This autoimmune disease is caused by destruction of beta cells in pancreas and is usually triggered by environmental factors.
Type 2 diabetes on the other hand is associated with impair insulin action in adipocytes, hepatic, peripheral, and muscle tissue as well inadequate secretions from the beta cells. Type 2 diabetes is usually caused by sedentary life style and is mostly seen in obese patients.
Glucose intolerance or impairment may vary in degree. Although the exact mechanism is still unknown but some underlying causes can include decrease hepatic glucose uptake and glycogenesis, mostly seen in people with liver disorders such as cirrhosis. Other contributing factors can include insulin abnormalities and resistance in muscle tissues and some peripheral and hepatic tissues. Hormonal abnormalities and improper insulin action leads to to an abnormal glucose concentration in the bloodstream.
Gestational diabetes is seen in pregnant women during their second or third trimester. During pregnancy some hormones including human placental lactogen or chorionic somatomammotropin and cortisol are secreted that act as insulin antagonist. These hormones inhibit glycolysis or glucose catabolism and promote lipolysis or metabolism of fat. As a result, the insulin requirement to reuptake the extra glucose concentration from blood stream increases. This condition can lead to various degrees of glucose intolerance and diabetes. (5)
Maturity-Onset Diabetes of the Young (MODY) is caused by a series of gene mutations causing a defect in pathway of insulin production and and secretion. MODY usually develops during late adolescents and is usually misdiagnosed as type 1 or type 2 diabetes.
However, unlike type 1 diabetes, patients diagnosed with MODY have negative lab results for the presence of autoantibodies and unlike type 2 diabetic patients, they are neither obese nor insulin resistant. (6)

Complications
In most cases untreated diabetes leads to other complications such as cardiovascular disease, nerve damage or neuropathy, kidney damage or nephropathy, eye damage, foot damage, hearing impairment, and Alzheimer’s disease. (3)
Proper treatment and management of symptoms in diabetic patients is critical, because most complications involved with this disease develop gradually over time and become life-threatening. For example, diabetic patients are at a higher risk of developing silent heart attacks. Due to the effect of diabetes on nerve damage patients’ sensory neurons do not carry information to the brain properly. Therefore, they do not feel pain or other commonly known symptoms related to heart attack. On the other hand, high blood glucose level, hyperlipidemia, and high cholesterol level that is often seen in diabetic patients increases the risk of arterial blockage and development of stroke or heart attack. (7)

The Impact of Biotechnological Devices on Recovery of Lost Functions in Diabetic Patients
The number of people who are affected by diabetes is increasing dramatically due to modern lifestyle. Consequently most of diabetic patients suffer from diabetes-related complications such as foot ulceration, restricted movement and amputation. As a result of these complications the quality of life dramatically decreases in affected patients. Therefore, many scientists have been devoting their resources on research about biotechnological devices that assist diabetic patients recover some of their lost functions. One of the most advanced and effective devices designed for this purpose is wearable robot. This high tech device aims to increase the quality of life, and decrease the cost of health care for the patients in the long run. Nonetheless, if this technology accomplishes its goal it will be a great help to deal with psychological effect of diabetes-related disabilities. (15)
A normal gait is a movement that contains both stance and swing phase. The action of lifting foot from the ground and stepping back to the ground is the stance phase. The swing phase is the duration between when the toe is off the ground until heel contacts the ground. (15) In patients with abnormal gait, adequate dorsiflexion that is the upward movement of foot is often a problem in the swing phase of gait. Similarly during the stance phase, an adequate planter flexion that is the downward movement of foot towards sole is problematic for these patients. In patients who are suffering form lateral paralysis of lower limb, swing phase is longer in order to avoid loosing balance and trip over.
The Exoskeleton Robots have been able to give some motion and movement to the affected part of the body. However, their movement is very restricted and rigid. It also takes a large frame to surround the affected limb or foot in order to mimic movements. The main focus of current studies is on creating an actuator that is able to draw the same actions but smoother, with a more manageable size, and a less power required.
Foot drop gait is not a disease by itself. It is one of the complications of diabetes. In this condition lower limb muscles of patients become paralyzed. This consequently causes dropping of forefoot and affected patients are not able to walk. The Active Ankle Foot Orthoses (AAFO) is a device designed to empower patients who are suffering from foot dropping gait condition.
The Active Ankle Foot Orthoses is composed of a basic orthoses, an ankle joint and an actuator. Orthoses is a brace that supports the affected limb and aims to correct its alignment. The actuator consists of two motors, the angular motor and the linear motor. The angular motor or potentiometer is attached to the ankle joint and the linear motor or potentiometer is attached to a backside of tibia with the help of a spring. The current that is sent through the motors and the resistance inside them are known. Therefore, it is possible to measure the produced force by taking the potential difference between the two known values. In order to measure the force applied to the device by the ground, there is a capacitive force sensor that is placed in the bottom of the device. (16, pg22-24)

The AFOs joint can be adjusted via changing its rigidity and resistivity to maximize the quality of gait for different patients with different gait abilities. This is the idea behind the Lower Limb Orthoses Magnetorheologic (MR) brake. The Magnetorheologic brake is placed at the ankle joint and can be adjusted depending on the gait quality of patients. The brake is applied by increasing or decreasing the resistivity of the device via changing the electromagnetic field inside the AFOs. Information about time required to control the brake is precisely calculated by a sensor embedded in the sole of the shoe that detects compressive force and another sensor that is located behind the tibia of the AFOs and calculates the bending moment. The MR brake enables the AFOs to maintain a constant dorsal flexion during a swing phase. The MR brake can also detect shock when heel hits the ground and in response, it can create a smooth stance phase. (15)

The third types of AFOs are called Artificial Pneumatic Muscle Powered AFO. Pneumatics is a Greek word for breath. Pneumatics technology follows a similar principal as hydraulics technology but instead of liquid, in pneumatics, compressed air is used to produce power. The Pneumatic muscle powered AFOs are composed of four essential parts including, a carbon fiber shell, a hinge join, and two artificial pneumatic muscles providing the dorsiflexion and planter flexion. The pneumatic muscle that is placed in the back of tibia contributes to planter flexion and the pneumatic muscle that is located on the foot contributes to dorsiflexion. The air pressure in each of the pneumatic muscles is adjusted independently. Therefore artificial muscle force is proportional to electromyography (EMG). (17)

The next type of AFOs is the Powered Ankle Foot Ortho (PAFO). Although there is yet to have such a device in market, the effect of this device is incredibly important. The support of this robotic device rehabilitates gait movements of patients. Since the energy required for acquiring these movements remains the same, this robotic device is very effective for use in recovery treatments. Also because of device’s active support timing push off is decreased noticeably. As a result, less stress is applied to the ulceration area and the healing time shortens considerably. An important accomplishment for this device is its changeable controller. The controller’s structure can be used in different stride lengths and is designed to adjust to different walking speeds. Therefore, even if later on patient acquires an abnormal gait behavior, the device will teach the patient the correct pattern over and over again. This will decrease the risk of long-term complications. Powered Ankle Foot Ortho consists of two core components including, mechanical and electrical parts. In order to provide torque to the ankle joint, a motor that is linked to a lead screw in series with a spring are placed behind the patient’s leg to assist with desired movements. (16,pg24)
The fifth robot is the Hybrid Assistive Limb (HAL). This device is a body suit robot with rotational motors at the joints that allow the robot to move. Movements of body parts are the result of signals from the brain to specific muscles. The first component of the HAL consists of translating the signals that are detected by many sensors that are placed on the skin. These signals are processed and translated into specific movements by HAL’s translating sensors. The second component is the motor component, which results in the ability to move different body parts. The action of the robot is a little bit ahead of the body movement in order to provide support for affected body parts. (16, pg26-27)

The sixth device is called Exosuit. The function of this device is intensifying the sustain endurance of muscle. In contrast with the mentioned devices, Exosuit does not have any rigid parts. It is mainly made out of nylon and straps. Exosuit provides a very small power to support the person who is wearing the device. The sensors embedded in the sole of the shoe function in acquiring the information that is required for the motors to adjust the power. (16)

Finally the very last device discussed in this paper is the MR Fluid Damper AFO. This device is considered a breakthrough in robotic devices. This is because of the support that the device provides to the wearer in maintaining a normal gait as well as enabling the person to walk on steep surfaces and go up and down the stairs. MR Fluid Damper has four different modes. The first mode is called the Damp Mode. During this mode, sudden movement of foot is restricted.
while the foot is down by a moderate damping. The second mode is called Free Mode in which a free motion is maintained and there is almost no damping. Lock, the third mode functions in maintaining the foot upwards during the swing mode. This action has a high dampening factor. The last mode is called the Free Down mode and has an opposite result from the third mode. During the swing phase, this mode assists in downward movement of the foot with a very small dampening factor. (16)

![Exosuit developed at Harvard University [Xue, 2014]. (16.fig2.8)](image)

**Physical Activity’s Effect on Type 2 Diabetes**
According to doctor Ingrid Strauch: “Type 2 diabetes is often called a “Lifestyle Disease”.” (8) Lack of exercise and bad habits are some of the main risk factors for developing diabetes type 2. On the other hand through proper training and exercise patients can take steps towards managing
the complications caused by this disease and further reduce the risk of developing coronary artery disease and other cardiovascular related diseases.

In a study that was published by doctor Sanjai Sinha on diabetic patients some acute effect of physical activity including increased glucose production, fuel mobilization, elevated level of muscle glucogenolysis, increased activity of glucose transporters/GLUT-4, and decrease lipolysis; that overall leads to decrease glucose concentration in blood serum was observed. (8)

In a study done in Dunlap by a group of researches, it was observed that physical training enhances insulin-mediated glucose uptake from skeletal muscle via local mechanisms. In this study that was published in American Journal of Physiology, seven male subjects were recruited and trained for a period of 10 weeks. Subjects were monitored for their arteriovenous differences in both legs via euglycemic hyperinsuliemic clamp (clamp1) method. “Within 2 weeks before the start of a 10 week training program the following measurements were made for both legs: isokinetic, concentric strength of quadriceps femoris at angular velocities of 30, 60, 120, and 180/s” (9)

Procedure
Using computer tomographic scans (CT) the circumference and cross-sectional area of subjects’ thigh and calf were also measured and recorded. During the bicycle training, subject wore a special bicycle shoe on one foot while their other foot was resting. Before a three-step clamp, subjects’ urine volume was measured. For each step of clamp a 50-ml solution of insulin infusate containing insulin, saline, and 2.5-ml serum albumin was prepared (9). Insulin was administered at each clamp step as a 2-ml bolus at first and then at a constant infusion rate of 258 µl/min for 120 minutes. At 5-minute intervals subjects’ arterial plasma glucose concentration was measured. Subsequently computer-adjusted glucose solutions were infused to maintain a constant concentration. Lastly, subjects’ blood sample was collected for measurement of hormones and other metabolites. (9)

![Effect of Training on Insulin-Mediated Glucose Uptake in Human Muscle, Fig.1.](9)
Result of the Dunlap study
Based on the data collected in the study, it was shown that physical activity enhances insulin-mediated glucose uptake, storage, and glycolysis (glucose breakdown). An explanation that was given by the author for increased insulin sensitivity seen in subject’s trained leg was muscular hypertrophy. (9) Training leads to increase in cross-sectional area of muscles, hypertrophy, and increased capillarization. (10) According to the author, this could increase the number of insulin-receptors or enhance their affinity for insulin binding. (9) These findings correspond to the hypothesis that professional training and physical activity can help diabetic patients manage some of the complications caused by this disorder. Type 2 diabetes is associated with impair insulin actions and its inability to properly activate metabolic pathways that lead to glucose uptake from blood stream and its storage in hepatic tissues. (4) As it was observed in doctor Sanjai Sinha’s study, physical activity enhances responsiveness of these metabolic pathways that function in maintaining a proper glucose homeostasis. Another interesting finding of the Dunlap’s study was the duration of these effects. The study showed that while exercise enhances insulin-mediated glucose disposal in muscle tissues, its effect fades as training stops. In other words, exercise is only effective in managing diabetes’ complications if it is continuous. (9)

In a similar study the relative benefits of a 4-month endurance and strength training in improvement of type 2 diabetes-related cardiovascular risk factors was studied. (11) In this study a total of thirty-nine participants that were diagnosed with type 2 diabetes were recruited. Twenty-two people were randomly assigned to a strength-training program and the remaining seventeen people were assigned to an endurance-training program. (11)

Endurance training
Muscular endurance-training helps in sustaining muscle activity over a long period of time while making simple tasks such as going up and down the stairs easier. Muscle endurance-training enhances the ability to do something over and over without getting tired. (12) Cardiovascular endurance-training aims to enhance the ability of the heart to pump blood and oxygen while body is going through long periods of movement. It also aims to maintain a steady respiratory rate throughout a long workout session. (12)

Strength training
Strength training increases strength, muscle mobility and hypertrophy. It improves balance and helps in maintaining bone density. (13),(14) Engaging in regular ST programs can prevent many disorders associated with a sedentary lifestyle, such as cardiovascular disorders, diabetes and obesity. These activities increase mobility in people with arthritis and by maintaining bone density and strength; they can prevent bone related disorders such as osteoporosis. (13)

Procedure
In this study the endurance-training group were required to complete a 15 minutes training session on a cycle ergometer three times per week.
A similar workout session was designed for the strength-training group. Subjects were required to participate in a 10-minute low intensity pre-workout training. During the first two weeks weights were kept at minimum. After the third week, the workout sessions were advanced to three repetitive sets per muscle to induce hypertrophy. The strength-training program was designed to increase upper body strength. (11) At the end of the 4–month period participants were assessed and the collected data between two groups was compared.
Result
According to the study, after completion of the 4-month training program, there were significant improvements in both ST and ET groups. In the ST group, there was an increase in maximum muscle strength, a decrease in body fat mass and a significant systolic and diastolic blood pressure reduction from the baseline. Insulin sensitivity (IS) was measured using computer homeostasis model assessment (HOMA) in both groups. According to the data collected from the study, insulin sensitivity had significantly improved (P.04) in the ST group. However there was no significant IS improvement seen in the ET group. (11)

Based on the improvement in insulin resistance and reduced glycated hemoglobin (HbA1c) that was seen in the study, ST has a significant beneficial effect in glycemic control in diabetic patients. “The strong association observed between muscle size and glycemic control supports the importance of muscle tissue in IR in type 2 diabetes.” (11)

According to the study, endurance training’s effect on IR and HbA1c however were only moderate. Another interesting finding of the study was the effect of ST on metabolic factors that contribute to cardiovascular related diseases. Based on the data collected from the study, after 4 months of straining-training, there was a significant improvement in atherogenic lipid abnormalities and reduction in triglyceride, LDL and total cholesterol levels. (11)

Conclusion
The aim of this paper was to provide a close review of studies on the impact of physical training and biotechnological advances on prevention and management of diabetes and diabetes-related complications. As it was seen in dr. Sinha’s study, continuous physical activity enhances responsiveness of metabolic pathways that function in maintaining a proper glucose homeostasis. This function is highly beneficial in type 2 diabetic patients who suffer from insulin resistance and improper glucose homeostasis. (13)

According to the Simpson’s study, exercise and physical training can help patients avoid developing diabetes-related cardiovascular disorders by improving patients’ atherogenic lipid abnormalities, triglyceride, LDL and total cholesterol levels. (11) Exercise improves the overall health of patients and assists them in management of their diabetes and its complications.

On the other hand, new breakthroughs in biotechnology research has introduced more effective and promising robotic devices that can help patients recover some of their lost functions. Diabetes-related complications decrease the quality of life for the effected patients and over time can lead to foot ulceration, restricted movement and amputation. These advanced devices aim to assist these patients with their daily activities while providing them with a shorter recovery time. (16)
References


One Giant Leap for Mankind

Maggie Scott
4/20/16
Astronomy 112
Section 12666
Dr. Bill Sherry
In July of 1969, the United States' Apollo 11 was the first mission to send a man to the Moon (10). When NASA had met Kennedy's goal to put a man on the Moon before the decade was finished (8), a new perspective of the universe was born. The goal of this paper is to address the history of space exploration and the triggers that led to this mission. The extensive process and events of the trip itself will be broken down, as well as the findings during the mission. Lastly, this paper will also address the impacts of the mission on science and technology, as well as the impacts on society as a whole, and possible impacts of future opportunities of space travel.

As Neil Armstrong took the first step onto the Moon, his famous words were, "That's one small step for man, one giant leap for mankind" (7). As an estimated amount of 530 million people were watching this take place on television (7), the world's perspective of the universe had begun to shift. During a heated period of the Cold War, thousands of people within NASA worked on an extensive game-changing project that proved to be a success (7). With the phenomenal outcome of Apollo 11, there was a huge impact on the way people thought about space and its possibilities, as well as the motivation to continue space exploration. Even to this day, the effects of the Apollo 11 mission can still be seen, and one can only guess what the future might hold in terms of space travel.

The history of lunar exploration mainly emerged from the intense environment of the Cold War between the United States and the USSR, which began in the late 1940s. During this time, both countries had goals of showing superiority in specific areas such as military developments, technological advancements, and space program achievements. At one point during this constant competition, both countries set their eyes on the sky. Although space exploration was valued for the increased knowledge of the universe and space itself, it was also important to both countries because it could provide more militaristic opportunities (5).

The Space Race between Soviet Russia and the United States began in October of 1957 with Sputnik 1, the first satellite that launched into orbit around the Earth (10). The USSR continued with their space success a month later by launching Sputnik 2. Inside Sputnik 2 was the first living thing to orbit Earth, a dog named Laika(10). After the launch of Sputnik 2, a back-and-forth space rivalry began between the United States and the USSR. From 1957 to 1969, over 40 missions were attempted by the two countries (10). Some of these missions had goals of orbiting Earth, orbiting the Sun, reaching the Moon, launching satellites with specialty equipment, capturing images of the universe, orbiting the Moon, and lastly, sending men to the moon. One of the last ventures that took place during the Space Race was the United States' mission of sending Neil A. Armstrong, Michael Collins, and Edwin E. "Buzz" Aldrin Jr. to the Moon.

According to NASA, the goals of the Apollo missions as a whole were to establish the technology to meet other national interests in space, achieve preeminence in space for the United States, carry out a program of scientific exploration of the Moon, and lastly, develop man's capability to work in the lunar environment" (11).
The process of getting these men to the Moon and back home successfully proved to be a challenging and suspenseful task for those in space as well as for those on Earth. At 9:32 a.m. on July 16th, 1969, Saturn V, a three-stage 363-foot rocket with 7.5 million pounds of thrust propelled the men into the Earth's orbit (7). In Figure 1, the intensity of the blast-off can be seen by observing the trail of heat and smoke Saturn V left behind. After one and a half orbits around Earth, Apollo 11 departed from its orbital path by re-firing the third-stage engine to reach a velocity of 24,200 miles per hour and began the journey towards the Moon (3). As the Apollo moved towards the Moon, the gravitational pull of the Earth cut its speed in half, the slowest speed being 2,000 miles per hour. This process is called "Translunar Injection" (3).

Apollo 11 began its orbit around the Moon three days later at a speed of 4,000 miles per hour, relative to the moon (3). Its initial orbit was a path shaped like an egg, with its low point of 61.3 nm and its high point of 168.8 nm (3). The path then evened out a bit, transitioning into more of a circular shape with points of 65.7 nm and 53.8 nm (3).

The day after that at 1:46 p.m. is when the crew split up, with Neil Armstrong and Buzz Aldrin aboard the Eagle lunar module and Michael Collins remaining in the Columbia command module to oversee operations occurring on the surface of the moon (3). As Collins continued to orbit the Moon, Armstrong and Aldrin headed for the surface of it. Figure 2 is an image of the Eagle lunar module, which is equipped with special landing gear that would notify the crew to turn off the descent engine (7).

Roughly an hour and a half later is when the process of landing began. Having a low point of 8.5 nm, the lunar module was put in an elliptical orbit above the moon (3). When this low point was reached, the Eagle lunar module exited the Moon's orbit and began on an arching path to the surface of the moon. To confirm that the module was on track, there were two check points of different altitudes and locations - the first being 7,600 feet in altitude and 26,000 feet laterally from the next check point, which was an altitude of 500 feet and being adjacent to the landing zone (3). Both check points were met, and the process of landing occurred according to plan.

The landing zone was to be in what is considered Mare Tranquilitatis, which translates to the Sea of Tranquility (2). This was the chosen landing site because the location is relatively level and smooth. However, as the lunar module made its way to the surface of the Moon, the astronauts noticed that their pathway led straight into a crater that was covered with a number of massive boulders, which would obviously prevent a successful landing (7). At the last minute, Neil Armstrong had to take manual control of the lunar module four miles away from the initial landing site to steer it away from the danger. At the same time, Eagle's computer began to sound of alarms that were not easily recognizable. Fortunately, the alarm was merely because the computer was trying to do too many things at once (7). Soon, at the other end of operations at the home planet, the next statement heard was, "Houston. Tranquility Base here. The Eagle has landed" (3). There was only enough fuel left to sustain the Eagle for 30 seconds longer. It is estimated that the Eagle ended up landing 120 miles southeast of the Maskelyne crater (3).
The trip home was routine, and uninterrupted. Throughout the entire trip back, there was only one correction that was made to the course - a slight change in velocity of 4.7 feet per second (3). Because the weather in the intended landing destination was bad, the landing site was shifted 215 miles away. On July 24th, the crew landed in the Pacific Ocean (3).

While Neil Armstrong and Buzz Aldrin were on the Moon, they were productive in making observations, collecting samples, and conducting experiments. The first observation made was by Buzz Aldrin looking out the window of the lunar module. He stated, "It looks like a collection of just about every variety of shapes, angularities and granularities, every variety of rock you could find. It looks as though they're going to have some interesting colors to them" (3). Once Neil Armstrong stepped out of the craft, he immediately noticed that the surface material of the Moon seemed to be fine and powdery, as can be seen in Figure 3; however, his steps sunk a quarter of an inch or less (3). Armstrong quickly gathered a few pounds of the material that made up the surface of the moon. He noticed that once he dug deeper than five to six inches down past the Moon's surface, he would come upon a hard, consistent material (3). Between the two men, around 50 pounds of samples were bagged up and stored in the Eagle. These samples mainly included the powdery surface material, and rocks found on the surface (3).

The surface of the Moon was described as different shades of grey. The temperature on the moon ranged from 180 degrees Fahrenheit in the sun and -160 degrees Fahrenheit in the shade (3). The angle at which the two men stood affected and seemed to change the color of the Sun. At an angle of 0 degrees, the Sun looked white (3). Lastly, the way in which Armstrong and Aldrin could move was surprising. There had been concerns and reservations about movement on the Moon and how it could be challenging with a space suit that limited flexibility; but these fears were soon settled as the men reported to televisions on Earth that movement was simply different. It was important to know one's center of mass, and there were specific ways to move that made moving and turning easier.

Some of the experiments the men conducted had to do with measuring solar wind. A small, rectangular mechanism with foil-like material attached to a horizontal rod, as well as an American flag with a rod along the top edge to keep the flag extended out both helped gather information (3). Other demonstrations were conducted, specifically for the audience down on Earth, at home or in the control room. On the trip home, a broadcasted program showed Michael Collins attempting to pour water out from a spoon; the physics weightlessness were apparent, due to the fact that the water remained in the spoon when the spoon was upside down. The process of drinking water was demonstrated, and a gyroscope was illustrated with a spinning can of cheese (3).

The work that the men did during the mission of Apollo missions led to many discoveries and new understandings regarding the Moon. The new information led to discoveries about the Moon's environment, age, composition, history, form, and surface. Before Apollo, the moon was nearly a complete mystery. Little was known about its
characteristics; fortunately, the successful trips proved to definitely be a "giant leap for mankind."

The history of the Moon is one of the major aspects that were addressed during these missions. Due to its extensive amount of surface craters and rocks, illustrated in Figure 4, scientists were able to use rock samples to estimate the age of the Moon. According to the Smithsonian National Air and Space Museum, the youngest rocks found on the Moon are the same age as the oldest rocks found on Earth (4). The ages of these Moon rocks range from 3.2 billion years in the lower valleys of the Moon to 4.6 billion years in the higher terrain (4). Nearer the beginning of the Moon's history, the Moon was melted in great depths, forming what is called a "magma ocean". Scientists came to this conclusion because of traces of ancient rocks with a low density that once had floated to the top of the magma ocean. After the magma ocean had run its course, a series of massive asteroids hit the Moon, creating dents that lava flows eventually filled (4).

Continuing with discoveries of the Moon's surface is the fact that all Moon rocks were developed through processes that contained very little water and high temperatures (4). As a general rule, the three main types of rocks found on the Moon are basalts, anorthosites, and breccias (4). Its surface is comprised of what is called lunar regolith, which is referring to the many fragments of rock and the fine powder Neil Armstrong had observed during the first few moments of stepping out of the Eagle (4). What he had not noticed, however, is that the surface contained a nearly complete and very unique record of solar radiation (4). The mass of the Moon is not equally distributed, as its center of mass is more towards Earth; the gravitational pull of Earth could be the possible cause of this lopsided characteristic (4).

Lastly, the Moon is similar to Earth and other planets in many ways. Scientists believe that by calibrating its craters using rock samples, the key to estimating the age of other planets may be unlocked as well. Its internal structure is also similar to Earth's, as it has a crust that is roughly 60 km thick, which has been affected by meteorites and volcanoes (4); it also has a generally uniform lithosphere, which occurs at a depth of 60 km to 1,000 km deep, and an athenosphere that is semi-liquid and occurs at a depth of 1,000 km to 1,740 km (4). It has yet to be proven, but there is a possibility that there is an iron core similar to Earth's, due to slight traces of an ancient magnetic field (4).

Because of its relatively close proximity to Earth, as well as the fact that similar materials are found on both planetary bodies, scientists believe that the Moon and Earth are genetically related. Also, they believe that ancient occurrences and processes that took place on both the home planet and the Moon can be understood by looking at the history of the lunar surface. It is phenomenal just how many facts were gathered during the Apollo mission; this was one of many advancements in society.

An additional and completely independent arena that was heavily impacted by the Apollo 11 mission and the other Apollo missions was the technological aspect of things. To move a group of men and a ton of machinery required cutting-edge technology that did not exist at the time; on top of that, the Apollo 11 mission required unique ways of
communicating, whether through broadcasting to televisions, or a (very) long-distance call to the President (7), or even constant interaction with those managing the mission down on Earth. The growth and expansion of the world of technology was a necessity, even though it was years ahead of its time.

A particular example was the absolute need of NASA for a machine or program that provided a dependable navigational system for the astronauts on their way to the Moon. A number of people with PhDs in Physics were requested to create a way to softly land "on a moving target hurtling through space a quarter of a million miles from Earth" (12).

Eventually a program was created; however, it was not completely reliable and would have to be altered a bit while the crew was already in space. This produced a new issue: the need for portable computer machinery. The spacecraft had to be able to hold this computer, and at the time computers typically took up an immense amount of space. A new invention, the integrated circuit, that was still in its experimental phase was looked to for a solution. Fairchild Semiconductor worked with NASA, which bought one million of these circuits, in order to create a reliable system for the astronauts (12).

It is clear that this project was successful, because Apollo 11 ended up being a major victory. The point of this example is to illustrate how so many groups of people came together to push the envelope, and push past the current boundaries of present-day technology. According to The Guardian, in 1969, two employees from Fairchild Semiconductor began a new computer company and called it Intel (12). The advancements made in an attempt to send men into space also cause advancements in other areas of culture and technology.

These advancements soon trickled down into everyday life. The portable computer technology brought onto the spacecraft was later on seen in pocket-sized calculators carried by college students in the 70s, then home computers in the 80s (12); then in the 90s came the internet which linked people from across the world, and now it can be seen in laptops, smart phones, video streaming, credit card swiping devices (9), and social networks such as Facebook, Twitter, and Instagram. The effects of the Apollo missions can still be clearly seen today.

Other often over-looked advancements that directly originated from the Apollo missions can be seen in items like liquid-cooled garments worn by firefighters and race car drivers, which were initially worn by astronauts under their suits (9). The infamous freeze-dried food astronauts ate in space are now used as military field rations (9). This simply reveals the thorough process put into play by NASA - they utilized every possibly beneficial technology in order to have a successful trip, specifically in terms of keeping the astronauts safe.

The effect that the Apollo 11 mission had on society in 1969 was astounding; and the truth is that we are still experiencing some of its benefits. Culturally speaking, nearly the entire world joined in on the event, as much of the historical trip was recorded and broadcasted. 74 nations sent messages of encouragement to the Apollo crew, and over
3,000 news reporters from 55 different countries, not including the United States, covered the event (3). To say the least, this event did not simply serve as a "one-up" on the USSR for the United States during the Cold War; instead, it provided a unifying realization that mankind has the ability to explore further than ever before. For many people, dreams came true. For everyone, a new reality was born.

Apollo 11 created a new perspective of life for many people. The trip to the moon was the embodiment of victoriously accomplishing the impossible. It inspired many people to follow after new ideas that seemed intimidating. For example, Jeff Bezos was completely inspired by Apollo, and after creating the 90 billion-dollar company known as Amazon (1), he fully links his motivation and inspiration to the Apollo missions of sending men to the Moon (12).

An additional example is of Martin Sweeting, who founded the satellite company SSTL, which was a brilliant new start to communication and interaction with others. According to The Guardian, he was often encouraged that his idea was too risky and that he should go out and get a real job. He believes that because of Apollo, he would did stick with his idea (12).

Both of these allegories are simple and personal, but science really can be personal. To capture the heart of countries at a time, to inspire generations, to deeply affect so many different branches of the economy and of production, and on top of that, to intensely increase the world's knowledge of life outside of the atmosphere is quite the victory. Apollo 11 made such a large footprint in history, and now it is simply a question of when the next major, ground-breaking footprint will be made.

Regarding future possibilities in terms of space exploration, the current hot topic is that of space tourism, which is exactly what it sounds like. Companies are building and planning spacecraft that will hopefully one day transport cargo and people into space. There are already companies with promises of these kinds of tourist trips. Companies like Virgin Galactic with its craft named SpaceShipTwo are already taking down payments for future trips, for which the total price is hundreds of thousands of dollars (6). Another company, SpaceX, is currently working on spacecraft and rockets that will transport more cargo and people than any craft since Saturn V (6).

Years from now, traveling to the Moon could be like going on a vacation - special but not impossible. It is amazing to think that the contributing factor of that possibility was born in the 1960s when President John F. Kennedy announced that, before the end of the decade, the United States would be sending an American to the moon safely (8). In 1969, not only did NASA send one American to the Moon, it sent two. Apollo 11 was a bold statement in the midst of the Cold War, but that was not the only motivating factor. In additional hopes of gaining new knowledge and continuing to explore, NASA conducted a thorough process of sending astronauts to the Moon; this trip proved to be impactful for decades to come in the realm of science and technology, as well as culturally. It has even paved the way for future developments of space tourism.
FIGURES

Figure 1: An image of Saturn V, a three-stage 363-foot rocket with 7.5 million pounds of thrust that transported the Apollo 11 astronauts into the Earth's orbit. Reference: NASA. Saturn V Take-Off. NASA [Internet] [cited 2016 Apr 20]. Available from: https://www.nasa.gov/mission_pages/apollo/apollo11.html

Figure 2: An image of the Eagle lunar module, which was equipped with special landing gear that would notify the crew to turn off the descent engine. Inside was Neil Armstrong and Buzz Aldrin. Reference: NASA. Eagle Lunar Module. NASA [Internet] [cited 2016 Apr 20]. Available from: https://www.nasa.gov/mission_pages/apollo/apollo11.html
Figure 3: An image of the Moon's surface, which seemed to be fine and powdery; however, his steps sunk a quarter of an inch or less. Also illustrates the different shades of grey of the lunar surface.

Figure 4: Due to its extensive amount of surface craters and rocks, scientists were able to use rock samples to estimate the age of the Moon. The ages of these Moon rocks range up to 4.6 billion years.
References


The Untold Truth of Tanning Beds

Tedi Serreti
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Physics 112
Professor Michael Swingler
Abstract

On this research paper will be discussed the technology behind tanning beds. The positive and negative effects of tanning beds on the human body will also be mentioned. The tanning salon industry will also be used to discuss how people are manipulated into spending money on ruining their skin.

Body Paragraph

The origin of tanning beds goes back to 1906, when it was used on patients with calcium deficiency. People who needed to create stronger bones benefited by this technology for a short period of time. But how did tanning bed create stronger bones? Vitamin D is produced by sunlight and allows absorption of calcium. And the absorption of calcium in the body creates stronger bones. The use of tanning beds for medical purposed ended when a man called Friederich Wolff thought about tanning beds in a different way. He used the tanning beds on athletes to make their bones stronger but he also noticed a darker skin complexion on these athletes. He opened his first indoor tanning bed salon in Europe and later in America. Over the years, tanning beds have advanced drastically thanks to technology. Nowadays tanning beds obtain air conditioning, radios and numerous safety features. Today tanning salons are a multi-billion industry. However, many people do not fully understand the risks taken while using these tanning beds.

The majority of people who use tanning beds do not know the science behind this technology. Although they might know most of the risks associated with tanning beds, they do not really try to learn about the science, but they genuinely believe to the employees of the tanning salons. So what really is tanning? Tanning is the skin’s response to ultraviolet (UV) radiation. Under the exposure of ultraviolet radiation, skin cells produce a brown pigment, called melanin, as a defense mechanism, to protect themselves from further ultraviolet exposure. The darkening of the skin or tanning, is basically body’s natural defense mechanism to prevent further damage from ultraviolet radiation. Tanning beds are a source of ultraviolet exposure and sufficient exposure can cause adverse health effects on human’s health. The different types of UV rays determine what kind of effects will happen to a human body when there is too much exposure.

There are three kinds of UV radiation and they are classified by their wavelength. UVA rays (315-400nm) have the longest wavelength, these can enter deeply into the skin and cause premature aging, immune system suppression, eye damage and skin cancer. On the other hand, UVB rays (280-325nm) are intermediate in wavelength, but they are nearly 1000 more efficient than UVA in causing a suntan and associated skin damage. When UVB comes in contact with skin, it can cause tanning, burns and skin cancer. Luckily, only 10% of UVB reach the surface of earth. Both UVA and UVB rays can cause cancer and are responsible for premature aging and skin damage. UVC rays (180-280nm) have more energy than UVA and UVB rays, but fortunately they have the shortest wavelength and react with ozone layer. Even though they do not penetrate deep into the skin, UVC rays are able to destroy nucleic acid in the cells. UVC rays are often used in germicidal lamps to kill bacteria and other organisms. All three of the UV rays have the ability to produce damage on the human body so it is best to be completely cautious while
handling any of them

Tanning beds are the main source of artificial UVB rays. Epidemiologic evidences suggest that artificial UV radiation causes melanoma, a type of skin cancer. The radiation that tanning beds transmit to the body are the same type and amount of UV radiation as the summer sun at noon, and sometimes more. Artificial radiation is no different from natural radiation, both kinds can cause skin damage, and possibly can cause melanoma and other skin cancers. UV radiation damage to DNA skin cells can result in mutations that promote cancer. Even if someone never develops cancer, the damaged DNA can increase his or her children’s risk of developing it.

There is no safe or permanent tan. Tans do not last for a long time, but the damage to the skin is almost permanent. Repeated UV exposures can cause photoaging, which consists of wrinkles, sagging, loss of elasticity and sunspots. Yet regular tanning salons clients think that tanning makes them look more youthful and it gives them a positive self-image. They like to look in the mirror and feel good about their healthy golden glow. When in actuality they are causing their skin to age prematurely. So what will stop people from going to tanning salons? There are healthier alternatives to get a tan. Recently tanning sprays and lotions have hit the market. These products have become more popular because they are less time consuming and very easy to apply. There is still the look of youth and tan skin, but the risks of any type of skin damage is significantly decreased.

But UV rays do not always cause harm. There are some skin problems that are treated with UV light. A drug called psoralen is given for a treatment known as PUVA. The patient is treated with UVA radiation after the drug collects in the skin. Another option is to only use UVB radiation and for this treatment option drugs are not needed. Some people in states that do not receive a lot of sunshine such as Washington and Oregon have a vitamin D deficiency. In order to gain more vitamin D, they usually take tablets containing the vitamin or use a UV light and shine it on their skin. These uses of UV rays are very beneficial to those who have these skin problems, but caution is always heeded when coming in contact with the UV rays.

Today, tanning is seen as a trend. According to Dermatology Nursing magazine, the intensity of the UV light emission has decreased over the past thirty years, but the desire to be tan has increased drastically. On an average day, one million Americans visit tanning salons. And 92% of these people say that they understand that tanning is dangerous, but they think they look better with a tan. The conclusion of these statistics is that Americans are associating tan with beauty. Now, there are tanning salons that offer different levels of tanning beds for different skin type. Every person now can select the amount of tan they want, so for instance if customer’s skin tone is medium white, there is a tanning bed tailored for this individual customer. Also many salons have upgraded their tanning beds to stand-up beds. This new upgrade provides a better and all-over tan. Although, now there are new features since the original lamps of 1906, safety measures should be taken. Eye protection is required for individual who want to use the tanning beds or booths. Cataracts are known as side effect of constant UV exposure. Also, it is highly recommended to apply sun lotion 30 minutes prior UV exposure. Sunscreen and sunblock are not completely effective at preventing cancer or other skin damage, but any type of prevention is better than no prevention at all. Regular customers must do monthly check-ups to detect early signs of skin cancer. If an individual has questions about artificial tanning beds, they should only be consulted with a physician. The employees at the tanning salons are not
professionals, their job is to welcome costumers and sell products. They know nothing about someone’s health history or small suspicious marks on skin. Most of them are not well informed about the risks associated with UV radiation.

Conclusion

In conclusion, these days, tanning is seen as a trend. According to Dermatology Nursing magazine, the intensity of the UV light emission has decreased over the past thirty years, but the desire to be tan has increased drastically. On an average day, one million Americans visit tanning salons. And 92% of these people say that they understand that tanning is dangerous, but they think they look better with a tan. The conclusion of these statistics is that Americans are associating tan with beauty. Now how can the use of tanning beds be decreased? Well first and foremost, tanning salons should stop glorifying tanning beds and start talking about the serious effects that come with being exposed to UV rays. Since there are also a lot of other products that can make skin tanner without exposure to UV rays, people should be promoting and using those products over those that harm the human body. All in all, there are a lot of dangers, as well as a few benefits, while being exposed to UV rays. But those who decide to use tanning beds need to ask themselves if using tanning beds is really worth the risk of aging prematurely, having bad skin, and even the possibility of being diagnosed with skin cancer.
This is a picture of light spectrum. We can see that UV radiation has a short wavelength, a high frequency and relatively high energy.
References


Electroconvulsive Therapy (ECT): Optimization of the Efficacy and Safety of Treatment with Modern Techniques

Kiarash Shams
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PHY112
Dr. Durandet
Abstract—

The use of ECT for nearly three-quarters of a century has been used to ameliorate the symptoms of severe psychiatric illnesses by inducing localized, controlled seizures in the brain during treatment. Modern techniques improved upon the traditional approach of conducting ECT by modifying the number of stimulating electrodes that are used and the arrangement of the electrodes on the scalp. Furthermore, advancements in electronics have made it possible to control the duration and current of each pulses that is delivered and the frequency of pulses that are scheduled for delivery over the course of an ECT session to minimize the side-effect profile experienced by patients. Charge dosing optimizations have been made through the years and effective foci of seizure induction have made it possible to stimulate regions that directly provide therapeutic results for the patient. If the treatment approach is not viewed as a “one-size-fits-all” solution and follows the “benchmark” approach towards individualizing treatment protocols, then the efficacy of ECT can be clinically significant with a favorable side-effect profile.

Modern Use of ECT —

At present, the United States health care system approaches treatment of mental health disorders predominantly with the use of medication. Pharmacological intervention is the first line treatment of choice amongst practicing psychiatrist and the staggering number of prescriptions filled for antidepressant medication ranks 3rd overall in the United States health care system.\cite{4} It is noteworthy to consider that most of these patients are perhaps not “depressed” which is suggestive by the name of the drug class, but are rather seeking relief from anxiety, bipolar, attention-deficit, eating, and a wide range of other disorders that seem to be alleviated with the use of these drugs. In many cases, patients are prescribed typical or atypical antidepressants, which help make the struggles of daily living for these patients substantially more manageable. However, there is a subset of patients who are plagued with mental disorders that are treatment-resistant against common antidepressant medications and treatments. These patients usually represent the more severe end of the spectrum of mental disorders and their individualized treatment courses are more difficult to define by the supervising psychiatrist. Nonetheless, the two distinct, yet interdisciplinary practices of physics and medicine have coalesced to find better solutions to treating such treatment-resistant patients. One of these famous treatments is called electroconvulsive therapy (ECT) and the principles behind the manner of its operation and efficacy follow some of the very early and fundamental ideas of electricity studied in physics.

ECT Technique—

Since the introduction of ECT in 1938, the basic apparatus set up has remained fundamentally similar from the time of its invention; however, necessary modifications have been made through the years to improve the safety and efficacy of the treatment. The traditional set-up includes a pair of stimulating electrodes that are placed laterally on the left and right brain hemispheres; this is referred to as bilateral (or bitemporal) ECT, which signifies the placement of the two electrodes above the patient’s ears on a region of the scalp that marks the location of the temporal lobe of the brain (\textit{Figure 1}).\cite{2,7} Newer methods include the option to change the positioning of these electrodes into other arrangements: right unilateral, left anterior right temporal (LART), or bifrontal placement (near the front crown of the scalp).\cite{7} Bilateral ECT has
been clinically proven to be more efficacious than unilateral placement and patients respond faster and to a greater degree with bilateral treatment. The supervising physician, however, may typically opt to initiate unilateral ECT before bilateral due to the more favorable side effect profile of the former.\(^\text{[9]}\) Despite the arrangements used, the electrodes are used for the sole purpose of delivering an electric shock to the patient’s scalp (Figure 2).\(^\text{[2]}\) Typically, the voltage set on the apparatus is 225 volts delivered via an AC source, but certain cases have deemed a maximum of 450 volts to be generally safe. The voltage intensity ranges used in ECT, regardless of strength, is very painful to the patient and a local or general anesthetic is advised during treatment.\(^\text{[5]}\) The electric current delivered is usually chosen within the range of 800-900 milli-amperes and the electrodes are charged to anywhere between 45 and 800 milli-coulombs.\(^\text{[8]}\) In addition, a circuit ground electrode is used to measure the difference in potential between the voltage produced at the stimulating electrodes and the designated ground electrode on the patient’s chest. An electroencephalogram (EEG) is hooked up to the patient’s scalp to simultaneously monitor the changes in neural electrical activity that occurs due to the stimulating electrodes that cause deviations to the natural electrical activity of the brain. The use of EEG becomes critical to determining whether the voltage delivered was adequate enough to initiate efficacious treatment. Due to the electrical nature of the treatment, and considering the fact that the skeletal and cardiac muscle systems of the human body are electrically excitable systems, the physician may choose to run an electromyogram (EMG) to monitor skeletal muscle activity and an electrocardiogram (ECG) to monitor cardiac activity. The electric shock delivered to patients in the past often had adverse effects such as uncontrollable muscle twitches and dangerous cardiac stress; therefore, a muscle relaxant, such as succinylcholine, is used to induce relaxation, but not complete paralysis of the muscular system during treatment.\(^\text{[9]}\) Following treatment with a muscle relaxant, the supervising physician often closely monitors respiration because the human diaphragm, which is controlled by skeletal muscle, is responsible for reducing the pleural pressure around the lungs enough to cause voluntary inspiration and becomes slightly inhibited with a muscle relaxant; for this reason, oxygenation through a respirator is advised during treatment.\(^\text{[9]}\)

**Fundamentals of seizures and their artificial induction with ECT**

The stimulating electrodes placed on the scalp are responsible for inducing localized, controlled, and observable seizures.\(^\text{[1]}\) Seizures are generally classified as either local or global hyper-excitation of neurons (brain cells) or groups of neurons that transmit signal information electrically via their axonal projections. These excitatory cells generally have a certain threshold that must be surpassed before they are able to relay their signal to the next neuron. This electrically excitable neuronal signaling occurs extremely frequently, between networks of billions of neurons, and bestows the magnificent processing power of the human brain. It is, however, when this pattern of signaling is disrupted to an overwhelming degree (i.e. hyper-excited) that seizures are observed. The goal of ECT is to induce such seizures albeit in localized parts of the brain and under controlled settings with the intent to achieve therapeutic results.\(^\text{[1]}\) Since variation between neuronal excitability exists between patients, it can be expected that the threshold necessary to thrust a region of the brain into a seizure episode can vary as well.\(^\text{[9]}\) It is the goal of the practitioner to gauge sub-threshold stimuli that are delivered and titrate the stimulus intensity upward to ascertain the threshold necessary to induce a localized seizure. Brain cells usually fire their electrical signals across their axonal projections in complex, yet
“normal” patterns relative to other healthy humans. The axonal membrane, which carries this electrical signal, is capable of accomplishing this because of the difference of charged, ionic concentration gradients established by sodium, potassium, chloride, and other ions across the membrane, which generate a membrane potential difference. When a threshold membrane potential is achieved, the neurons begin to fire and this is normal; when they reach their threshold more frequently than intended, and begin firing erratically and uncontrollably, this is abnormal and it is clinically classified as a seizure.

**ECT Parameter, Dosing—**

The intent of the electrode is to induce such abnormal brain patterns because it has been clinically shown to help improve a wide range of psychiatric illnesses.\(^9\) Once the seizure threshold of the patient has been determined, recent research has suggested a necessary stimulus that generates a response of 3-5 times the threshold necessary to initiate a seizure.\(^9\) A stimulus that generates such a response and is tailored to the individual is often more important that the absolute intensity of the stimulus in terms of clinical efficacy.\(^8\) A modern approach referred to as the “benchmark method” calls for the physician to assess the physiological events of the patient—peak heart rate, the intensity of peaks observed in the EEG scans, motor seizure duration, etc.\(^9\) For bilateral electrode placements, the appropriate initial ECT dose charge was found to be 2.5 mC per year of patient age using a 900mA current and 1msec pulse width.\(^9\) For unilateral placement, double the dose charge should be used (e.g. 5mC per year of age, 900mA, 1msec pulse width). It is important to note that a higher charge dose used correlates to a less favorable side-effect profile.\(^8\)

**Figure 3A** shows the EEG reading that typically is observed when a suprathreshold seizure stimulus is delivered to the patient.\(^5\) The erratic, high amplitude and frequent activity recorded by the EEG used to monitor electrical activity in the cortex denotes the patient is experiencing a seizure.\(^5\) **Figure 3B** activity is more diminished and portrays diminished neuronal activity that is observed following the electrical stimulus and seizure.\(^5\) The quality of the ECT can be ascertained through the strength of the seizure delivered.\(^6\) A lower stimulus typically produces the least efficacious results but also has a more favorable side-effect profile.\(^9\) Conversely, a stronger stimulus produces the most efficacious results, but also yields a more unfavorable side-effect profile. As such, a stronger electrical dosage produces a stronger neuronal depolarization and generates a larger seizure focus, which causes the seizure to affect more neural structures (i.e. increase in side-effect profile).\(^9\) The placement of the electrodes (i.e. whether it is bilateral vs. unilateral) also affects the strength and thus the quality of the ECT achieved. The closer the distance between the two stimulating electrodes, the smaller the volume of the seizure focal point, and the lesser the threshold charge necessary to induce a seizure.\(^9\) Dosing of the shock delivered is an important parameter that dictates the frequency and duration of treatment necessary to observe improvements in symptoms of a clinical significance.\(^9\) Stronger shock doses do not require as frequent of pulses of shock per session of therapy as weaker shocks require to achieve therapeutic levels of clinical significance.\(^6\)
ECT Parameter, Pulse Width and Frequency—

Modern ECT apparatuses feature current regulation and monitor the duration in which an electrical pulse is delivered. [8] The pulse width is the “window period” that delivers the dose charge. At present, research shows that ultra-brief pulse widths limit stimulation of brain regions involved in cognitive functioning, and thus improve the side effect profile for patients. [8] The briefer the pulse width that is delivered, the greater the increase in clinical efficiency is observed because the stimulus focuses on brain regions that regulate mood rather than regions greater associated with cognitive power and functioning. [7] Furthermore, as Figure 4 shows, the briefer the pulse width, the lesser degree to which the cell body of neurons (i.e. soma) are excited and induced to depolarize (and subsequently transmit their signal). [6] This is a significant finding because clinicians can minimize the extent to which the cell body of neurons are excited, which may potentially improve the side effect profile of ECT. [6] Pulse widths of ~0.5 msec have been found to maximally excite axonal projections of neurons, with minimal excitation of the neuronal bodies (i.e. somas) themselves. [6] This is significant because such a finding reveals that with ultra-brief pulses, a patient’s cognitive performance may return to baseline very quickly upon treatment. [8] Furthermore, shortening the pulse width and increasing the number of pulses delivered (i.e. the frequency) over the course of an ECT session has shown to dramatically reduce the adverse effects of ECT, while still maintaining efficacy of the treatment. [6]

ECT Parameter, Electrode placement—

The spatial arrangement of the stimulating electrodes is critical to achieving maximum therapeutic benefits and minimizing cognitive side effects. [8] The first line ECT treatment is usually right unilateral placement, which proceeds with the use of one electrode placed above the right ear, near the temple of the patient. [9] Since one electrode is used, a much larger stimulus dosage is required to achieve therapeutic levels; response rates of 45 percent are observed with an average dose of 175 mC and nearly 65-80 percent with dosages as large as 400mC. [7] However, the latter is often not recommended because patients often experience unwanted side effects such as impaired cognition and memory. In general, unilateral placement of the electrode embodies a lower brain volume in the seizure focus, lower efficacy, as well as a more favorable side effect profile. [7] Bilateral placement of the stimulating electrodes requires a smaller dose than unilateral placement—typically half the dose—and a larger volume of the brain is encompassed between the two electrodes. [8] As a result, there is a higher level of seizure generalization (i.e. more widespread seizure permeation), greater efficacy, and an unfavorable side effect profile. [7] This could be due to both hemispheres of the brain contributing to basic functions, and processing different seizure stimuli concurrently but differently; as a result, one side has the ability to mitigate to some degree the disorientation the other hemisphere has sustained. However, when both the left and right hemispheres are affected as in bilateral placement, the brain’s processing power is compromised to a greater degree and it has largely forfeited its ability to conserve the integrity of its processing power. [7] It is interesting to note that the efficacy of right unilateral ECT can match that of bilateral ECT by increasing the stimulus administration of the former electrode placement to nearly six times the threshold level of the patient; this result will reveal the same side effect profile as the more effective bilateral placement. [8]
**Duration of seizure induced by ECT—**

Prior to the use of muscle relaxers in ECT for the cessation of muscle twitches observed when an individual is experiencing a seizure (typically characteristic of the grand mal class of seizures), the event of a seizure was readily observable. In modern practice, muscle twitches are inhibited and EEG is used as the diagnostic tool to observe the brain under a seizure (Figure 3A). For a seizure to be of therapeutic value, a minimum duration of at least twenty-five seconds is advised. Occasionally, patients may not respond long enough for an efficacious treatment outcome, and administration of parenteral caffeine dosages from anywhere between 250mg to 1000mg may be required to excite neuronal tissue and lower the threshold necessary to achieve and maintain a seizure. On the other hand, some seizures may potentially last longer than expected; seizures lasting longer than five minutes are classified as Status Epilepticus episodes and may potentially damage brain tissue. These findings show the significance of using a correct duration for the localized, electrically-induced seizure to achieve therapeutic results and avoid extensive damage to brain tissue.

**Mechanism of Action—**

The ultimate goal of ECT is to produce localized seizures in brain tissue to observe the well-documented benefits of this procedure for patients suffering from extreme mood disorders (such as major depressive disorder), psychotic disorders (such as schizophrenia), and manic-depressive disorders (bipolar disorder), and anxiety disorders (such as obsessive-compulsive disorder). Although the electrical stimulus intensity is set by the apparatus and titrated up according to the patient’s individually varying needs, only about twenty percent of the applied charge actually enters the skull; this is because the oily secretion of the scalp bestows high impedance to the applied current. As shown in Figure 3B, there is a marked period of postictal suppression that is observed followed the induced seizure, and after about thirty minutes of “cooling-off”, the EEG scan shows a return to pre-seizure scans. Although ECT has proved effective in ameliorating symptoms for nearly three-quarters of a century, it’s exact mechanism of action still remains to be determined. However, there are a few working theories that provide the best explanations we have for now. One theory suggests that the seizure itself causes alternations in the firing of chemical messengers in the brain. After the electrical signal travels down the axonal projections of neurons, the signal reaches the terminals of the axon; here, the influx of Ca²⁺ ions into the axon terminal induces vesicles to release chemicals (i.e. neurotransmitters), which relay the signal to the next neuron to fire a signal. As such, signaling in the brain is characterized by the interplay between electrical firing and chemical signaling, and the stress induced by the localized seizures may cause some changes in the inherent pattern of this neurobiology, which for some individuals the dysfunction of may have due to a genetic predisposition to pathology in the form of mental illness. After ECT, certain neurotransmitter systems are moderated to increase sensitization to some chemicals, desensitize to others, and improve the neurotransmitter turnover (i.e. recycling for reuse) of key chemicals. These effects often mimic the effects of drugs that are observed with drugs, but it is fascinating to observe the brain autonomously making these changes after stimulation. Another working theory involves the rebound the brain experiences from the induced seizure. It is known that seizures increase blood flow, which subsequently revs up the metabolic rate of seizure-induced tissue,
consuming more glucose and oxygen than it regularly does. However, there is a marked decrease in metabolic rates and activity of the frontal lobes after thirty minutes of post-ECT seizure induction (Figure 3B) and this reduction in activity seems to explain the therapeutic outcome, as well as the temporary impairing cognitive side effects.

Side Effect Profile—

The most documented adverse reaction to ECT is memory loss and it is affected by many factors including electrode placement, stimulus waveform (i.e. modulated by pulse width), site of seizure initiation, and pattern of activation. Although most complaints regarding memory issues are transient and improve with time, the degree to which a patient experiences this side effect may be reduced by delivering a lower stimulus charge at the expense of reduced efficacy of the treatment. Furthermore, it has been postulated that such memory loss cannot be attributed to neuronal injury from the electrical stimulation imposed by ECT. It is interesting to note that anatomical evidence supports the opposite of such assumed neuronal death and/or atrophy; the disorientation experienced by patients of ECT may perhaps be due to the neurogenesis of cells in the hippocampus (gateway region in the temporal lobe of brain responsible for memory formation). Hippocampal atrophy may occur in patients with psychiatric illnesses, and ECT has been shown to restore to some degree hippocampal volume and improve memory processes in the long-term (as long as the procedure is not excessively used).

Conclusion—

Although modern research and practice has redeemed some of the bad reputation ECT has accrued through its history of use and popular culture, I am confident there are aspects of ECT that have not been fully exploited or understood. Although modern use of pharmacological agents have seemed promising, the induced rebound that is often observed in extremely severe cases with ECT may have potential for revealing deeper mysteries regarding the inner-workings of the human brain. Although the methodology of the process remains fundamentally the same since its inception, ECT has been thoughtfully and empirically optimized through the years by changing the number and placement of electrodes; the frequency, duration, and strength of stimulation; and employing modern electronics (ECG, EEG, constant current-controllers, etc.) to improve the therapeutic outcome. However, there appears to be certain domains that still need adequate research, such as improving the seizure focus (i.e. the spread of the localized seizure) to observe more controlled results with a more favorable side-effect profile.

Although the memory impairments that constitute the major side-effect from ECT seem unappealing to most patients, there is a balance between pros and cons that needs to be considered. First and foremost, psychiatric illnesses are notorious for impairing hippocampal volume and function and this in itself is enough to impair memory functions in the long run. Decreased interest in life (i.e. one of the hallmark signs of depression) has repeatedly shown to impair a person’s concentration, overall cognition, and ability to recall information. In this light, ECT may prove very beneficial to such extreme patients and may offset the “pros” adequately enough to improve the overall quality of life of patients suffering from such ailments. However, with the advent of newer technologies that may improve the delivery of the shocks and the treatment in general, we can expect improved efficacy and remission from ailing psychiatric
illnesses, as well as decreased side-effect profiles under more controlled conditions, in addition with the added benefit of revealing more about the brain and the way we regulate our mood and thinking patterns.
Figures—

**Figure 1:** Bilateral placement of the stimulating electrodes. The electrodes are placed above the ears on both sides of the scalp and right next to the temples. This stimulates the frontal and mediotemporal lobes of the brain. [7]
Figure 2: Schematic drawing of the traditional ECT set up. [2]
Figure 3A (top); 3B (bottom): The top EEG shows the seizure-induced brain via ECT. The bottom EEG shows the seizure induced brain immediately after cessation of ECT treatment. [5]

Figure 4: These two depictions show the variability in excitability of the axon vs. soma (cell body) of a brain cell (neuron) and the degree to which excitability changes with varying pulse width (msec).
References—


The Atmosphere of the Solar System Planets

Maziyar Soroor

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Professor Swingler
Abstract

From the beginning of human history humans curiosity about life outside of blue planet was always a dream and in order to reach that dream humans have to be aware of their solar system planet’s atmosphere. Without right atmosphere and pressure living on surface of a planet can be extremely hard. In this article there are remarkable information about the atmosphere and material of Milky Way planets that people can improve their knowledge after reading it.

1. Mercury

It is the smallest plant in our solar system, slightly larger than the moon. Its surface temperature can go from super-hot 800 degrees Fahrenheit to freezing 300 degrees below zero. Mercury is almost 48 million miles from the Earth and it makes that planet dangerously close to the Sun and making Mercury possibly the most dangerous planet to reach. Its atmosphere is so thin and barely detectable that rocks and floating material in space can easily hit the planet’s surface without burning, that’s why there are so many scars and holes on planet’s surface. Mercury is formed from 42% oxygen, 29% sodium, 22% hydrogen, 6% helium, 0.5% potassium and water vapor.

Like the Moon, Mercury has no plate tectonics and is geologically inactive. It is so small and its gravity is so weak that it is unable to hold on to an atmosphere. Mercury’s lack of atmosphere means it probably has no weather to create winds or clouds to blow across the horizon. Sunlight is so intense on Mercury because of its distance from Sun that its surface eroded during billions of years. Despite of this intense heat, ice may actually be found in the deep crater of planet’s north and south poles. Radar data has shown that permanently frozen water may exist in the areas that sunlight cannot reach them. Even with water on Mercury, without an atmosphere it is impossible to live on this planet and because planet is orbiting so close to sun makes it so hard to study.

2. Venus

Venus atmosphere is consist of carbon dioxide 95.5%, nitrogen 3.5%, sulfur dioxide 0.015%, argon 0.007%, carbon monoxide 0.002%, helium 0.001%, neon 0.001%. It is the planet that is completely covered by clouds, winds ripper around the upper atmosphere at over 300 miles per hour. About 35 miles above surface the winds are whipping around in a very aggressive pattern that basically carries the clock features around the entire planet every 4 days. Venus is closer to the Sun than Earth and its heat from the the Sun that drives the wind. The circulation of Venus’s clouds is quite simple, there is a warm side of Venus and cold side, as air warms up and rises on the dayside of Venus it spreads around to the night side. So the atmosphere circulation on Venus is quite simpler than Earth. Venus’s winds circulate faster than the faster hurricane on Earth but they look very similar. One of the great mysteries of Venus is that there are hurricanes at poles of Venus especially in the southern hemisphere that have two eyes which is really rare in planetary scale. Its double eyed vortexes 1800 miles wide, two hurricanes whirl around each other.

It is a mystery that how deep this raging spinning vortex goes but maybe this mega storm eye fall downward into planet’s surface. Down on its surface permanent twilight is all over the planet. The atmosphere huge weight creates an unbelievable pressure. The atmosphere pressure on Venus is about 100 times than Earth, the pressure is so high that no human body can resist.
and it will be flatten within a second. Venus’s dense atmosphere creates one of the strangest
mega storm imaginable. The wind are four miles per hour with the force of a hurricane, the
reason for such a slow winds is because the planet has a very slow rotation rate. It takes Venus
eight months to turn once around with respect to the Sun. The crushing atmosphere pressure
rams gas molecules together so tightly that they feel more like a liquid. Venus’s surface is so
unfriendly that life is impossible on the planet.

3. The Earth

The atmosphere of Earth is mainly composed of nitrogen 78.08% and oxygen 20.95%,
the remaining parts divided into permanent and variable constituent. This constituent consist of
noble gases such as; argon, neon, helium, krypton and xenon and other materials like carbon
dioxide, nitrous oxide, hydrogen, methane and radon. The unstable composition of atmosphere
are water, carbon monoxide and ozone. The pressure at sea level is one atmosphere. Pressure of
Earth atmosphere changes vertically, the temperature changes between 283 kelvin at night to 293
kelvin in day. Earth has four major layer atmosphere and the temperature varies between these
layers. The troposphere, stratosphere, mesosphere. Temperature decreases in two layers
troposphere and mesosphere with height. On the other hand temperature increases in stratosphere
and thermosphere with height and the reason is ultraviolet energy that is being receiving from
molecular oxygen.

The magnetic field around Earth generated by strong electric current which are guiding
by Earth’s metallic core. The magnetosphere’s shape is determined by the area of Earth magnetic
field, also by the solar wind and by the magnetic field between the planets. In the magnetosphere,
sources such as magnetic and electric are much more powerful than gravity and collisions limit a
mix of free ions and electrons from solar wind and Earth’s ionosphere. The life would not exist
without magnetosphere.

4. Mars

Mars is a desert like planet with a very thin atmosphere that is mostly made by carbon
dioxide. The Mars and Earth atmosphere are very different from each other. Mars’s atmosphere
consists of carbon dioxide 95%, nitrogen 2.7%, argon 1.6%, oxygen 0.13%, carbon monoxide,
0.08%, water 0.021% and nitric oxide 0.01%. The gas that dominates Mars’s atmosphere is
carbon dioxide. Scientist found that mars has been outgassed probably 20 times the present
amount of gas in atmosphere, these difference shows that most of carbon dioxide has been lost.
Three theories could have been the reason of this lost; First theory is that Mars may have lost its
carbon dioxide directly into the space, second the carbon dioxide may have outgassed and
combined into planet rock’s material through chemical processing and the third theory is it may
absorbed onto the soil particles. Scientists believe that Mars also outgassed between 10 to 30
times the amounts of nitrogen it has right now.

The main temperature on Mars’s surface changes between 184 kelvin at night to 243
kelvin in day. Like Earth, Mars atmosphere divided to four major layers. The lower atmosphere
which is a warm area, the middle layer which Martian jet stream flow and occur in this region,
thermosphere or upper atmosphere which has a high temperature because of sunlight and
exosphere which starts at 200 km. Magnetic field on Mars unlike Earth is weak. The
measurements for Mars showed that this planet has a large iron core which produces a magnetic
field but it is a weak magnetic field and the reason for that is the planet core turned to solid.
So Mars last layer of atmosphere which is magnetosphere is too small to protect the planet’s atmosphere against solar wind. Therefore dream of living on Mars may be possible but it is definitely harder than it looks like.

5. Jupiter

The chemical composition of planet Jupiter is similar to sun than any other planet in our solar system. The atmosphere on this planet based on ground optical telescopes and Pioneer 10 satellite data is consist of hydrogen 89.8%, helium 10.2%, methane 0.3%, ammonia 0.026% also clouds made of ammonia ice, water ice, ammonium hydro sulfide and other materials like phosphine, sulfide, acetylene, ethane and germanium tetra hydride which are found by Voyager probes.

Jupiter is the largest planet in the solar system, it is a gas giant. This planet rotates extremely rapidly, one day on Jupiter is 10 hours long, and that is the fastest spin of any other planet in solar system. A planet this big can reflect a lot of sunlight event orbits the sun at distance of 800 million kilometers. When people look at Jupiter they do not see its surface they see the top of its clouds and they are strange mix of permanence and change. The atmosphere of Jupiter is banded with multiple stripes running parallel to its equator. The lighter color stripes are called zones and the darker ones belt, they are fairly stable. Though their shape and coloring change over time. Belts and zones circulate around the planet in opposite directions. They form due to convection in Jupiter's atmosphere.

Upwelling air cools and forms ammonia clouds that creates the light colored zones. That air flows to the sides and sinks and sunlight changes the chemistry in clouds forming molecules that color the air yellow, red and brown. This is what causes the darker belts. Turbulence in the regions between zones and belts can create storms, gigantic vortices raging in the clouds. Dozens of then dot the face of Jupiter all the time but there is one to rule them all, the great red spot. A fittingly huge storm for the giant planet. It is actually a colossal hurricane several times larger than entire planet Earth with sustained wind speed of 500 km per hour. Jupiter rapid spin is what keeps red spot circulating and the redness is due to cyanide like molecules that absorb blue light letting redder light pass through. The red spot appears to be shrinking, it was essentially bigger and more elongated few decades ago. It also changes color over time too, having gone from deep red to salmon and then back again and the reason for this changes are still unknown.

Jupiter’s atmosphere is thick, several hundred kilometers deep. The pressure of atmosphere increases with depth. The planet does not have a real surface, the gas gets thicker and hotter and eventually just some sort of smoothly changes into a liquid over a hundred several kilometer range below the clouds. After Jupiter was formed it started to cool by radiating away heat from its upper atmosphere but this can increases the pressure inside the planet and it heats up. The heat works its way out of Jupiter and gets radiated away as infrared light, so in the end the amount of heat Jupiter gives off is more than it receives from Sun. On Earth the weather is powered by the Sun but on Jupiter it gets the energy from the planet itself. Jupiter has a very strong magnetic field to due to metallic hydrogen inside it which coupled with its rapid rotation. Like Earth it has twilight at its poles, as the solar wind funneled down to the cloud tops. Jupiter’s moon also play a role in planet magnetic field and twilight on Jupiter as well.
6. Saturn

Saturn is a gas giant which has a thick atmosphere but it is smaller than Jupiter, like Jupiter it probably has a rocky core several times the mass of Earth surrounded by layers of ice and metallic hydrogen, on top of all that it its atmosphere hundreds of kilometers deep composed of hydrogen 96.3%, helium 3.25%, methane 0.45%, ammonia 0.0125%, Etan 0.0007% and clouds made of ammonia ice and water ice. Saturn spins very fast completing a day in just 10.5 hours. Its rapid rotation and low density means it is really squished. There are two types of clouds on this planet. The upper level is seemed to be made of ammonia and the lower layer is either ammonia hydrosulfide or even water. Saturn bands of clouds are like Jupiter visible but it is a little fainter than the ones on Jupiter and also they are much wider around planet equator. Unlike Jupiter which always has raging and aggressive storms on Saturn happen every 30 Earth years which is equal to one Saturn year.

Large groups of clouds have been spotted on planet's equator many times and now researchers are sure that these are huge storms. Saturn winds are some of the most powerful storm in the solar system with the speed of 1800 kilometer per hour. Scientists recently observed an aggressive hurricane which unlike any other storm this one has a very well defined eyewall which is similar to storms that happen on Earth and that is something that has not seen anywhere in solar planet. The temperature on Saturn is quite bit chilly for humans, the warmest temperature is known to be negative 122 degrees Celsius and the average is known to be around negative 185 degrees.

7. Uranus

Uranus like Jupiter and Saturn does not have hard surface and because of that atmosphere of this planet is very deep. The atmosphere consist of hydrogen 82.5%, helium 15.2%, methane 2.3% and clouds made of ammonia ice, ammonium hydrosulfide and methane ice. It is estimated to be in tens of thousands of kilometers deep and is in three main layers: a troposphere, a stratosphere and thermosphere. The further they go deeper in the atmosphere the temperature of planet increases and the reason is the absorption of ultraviolet and IV radiation that is coming from the Sun which is a little surprising because of distance of Uranus from the Sun. Seasonal variation are extreme in Uranus climate and the reason is the fact that planet has a slope of 98% on its axis as it turns around the Sun. Tilt of the Earth which measures an exact 23%, it is more than enough to guarantee that Earth experiences well defined season. In 1986 Voyager 2 space probe was observing this planet it believed to be a dead world but in fact only southern hemisphere was visible, this was in contrast with other gas giants such as Saturn and Jupiter which are all known to have way more areas that are covered with clouds. Since the flyby of 1986 with more evidence and more information scientist have found that Uranus is not always this way, thanks to new telescope technology specially Hubble space telescope. Scientist have now managed to see more of the northern hemisphere and get more information about the formation of the clouds on this part of planet and they found out the formation is tend to be different from the clouds in the south. In the northern hemisphere clouds are smaller, brighter and sharper but they do not last as long as the clouds in southern hemisphere. Scientists also found dark spots on Uranus which believed to be aggressive storms with the speed of 824 kilometer per hour.

Understanding the seasonal variation in Uranus is difficult for researchers on Earth because it takes 84 years for Uranus to complete a full orbits around the Sun.
8. Neptune

From the images that has been received from Voyager 2 show that there are differences between Uranus’s atmosphere and Neptune's. Its atmosphere is a composition of hydrogen 80%, helium 18%, methane 1.5%, Etan 0.0002% and clouds made of ammonia ice, water ice, ammonia hydrosulfide and methane ice. The temperature in aerosol layer of this cold planet is -346 Fahrenheit (-210 C) which is close to Uranus’s main cloud level in atmosphere temperature, also the effective temperature of Neptune and Uranus were found to be close to this temperature. In Neptune’s troposphere and lower stratosphere temperature said to be about 59 Fahrenheit (15 C) colder than those in Uranus and the reason is Neptune has greater distance from Sun than Uranus. Neptune’s atmosphere looks like to be way warmer than if it would be receiving all the heat from the Sun and that is because this planet has an internal strong heat source unlike Uranus which has the weakest heat source. Neptune's atmosphere found to be similar to Uranus and it seems to have a little temperature change in latitude. It includes big heat capacities for both atmospheres. Also ionosphere in Neptune is super-hot 900 Fahrenheit (482) Celsius and an exosphere that is mainly consist of hydrogen thermal corona, this components are similar to ones in Uranus. Although Neptune’s stronger gravity but colder stratosphere cause lower particle densities in Neptune’s upper atmosphere that are found in clouds of Uranus’s atmosphere in the same of height. Scientist found evidence of methane hydrocarbon recycling in Neptune’s atmosphere and they saw that the methane is likely broken down by the sunlight. After methane got broken down, the result which is hydrocarbons sink into Neptune’s atmosphere. The hydrocarbons also decompose on their way downward and release carbon which with the methane that is in the upper atmosphere recombines. There are also areas such as Great Dark which occurs at 20 degrees south latitude and lesser dark spot about 50 degrees south. Clouds such as methane ice are white and about 31 miles above dark spots but they do not rotate with the spots.

9. Pluto

In August 2006 the International Astronomical Union (IAU) changed the status of Pluto to dwarf planet which only bring the rocky planets of inner solar system and gas giants of outer solar system will be known as planets but this does not mean human can ignore the importance of Pluto which used to be a planet once. “The atmosphere of Pluto is way more dynamic than anyone can imagined” 9. Pluto is freezing cold about 45 kelvin (-380 Fahrenheit) that carbon monoxide, methane and frozen nitrogen on its surface convert directly to gas but very slowly given Pluto unusual orbit. It swung closest to the Sun in 1989 but researchers believe that all gas surrounded Pluto will soon freeze and merge onto surface this can effect in billion years in feature to make this dwarf planet without atmosphere. Scientists surprisingly found that the upper atmosphere is very colder than they expected about 70 kelvin, their anticipation was about 100 kelvin and the reason is that the thin air does not go high enough to go away by the solar wind. There is necessary no nitrogen escaping from upper atmosphere of Pluto. Admits Michael Summers from George Mason University “it is all methane and very little of that is leaving either” 10 but still the reasons for upper cold atmosphere are unclear it could be other compound radiating away heat to space. Researchers found out that Pluto surface pressure has increased in recent years, explanations are mostly because of axial tilt it north pole is tipped 120 degrees downward with respect to “up” in solar system coordinates. It takes 248 years for Pluto to orbit around the Sun, Pluto has one pole always facing the sunlight and the other in the shadow. There
is enough frozen nitrogen and methane in northern hemisphere to keep the atmosphere from collapsing completely.” So not only Pluto’s atmosphere not only disappears but it can get thousands of times denser”9. That is the conclusion of a team of modelers led by Alan Stern New Horizons principal investigator. They also conclude that through those wacky tilt oscillates over million year time with dramatic implications for how much sunlight the surface receives and how much gas gets liberated.

One other aspect about Pluto’s atmosphere is that once sunlight breaks down methane, the molecular fragment quickly recombine to form heavier compounds like acetylene, ethylene and ethane but Pluto does not have enough gravity to hold on to these gases and as this planet gets further from the Sun the atmosphere freezes and turn to solid back down on the surface of Pluto. It is not proven yet but it seems that from Pluto’s dual orbit with its moon Charon that Pluto may have a magnetic field.

Summery

From the composition of planets we can divide the solar system planets into 3 major groups. Mercury and Pluto are the first group because they do not have the normal atmosphere in comparison with other planets because of their distance from the Sun, which Mercury is super close and Pluto is super far. Venus, Mars and Earth are the second group. All of these planets have large amount of nitrogen in their atmosphere, one of the other major elements in Mars and Venus’s atmosphere is carbon dioxide. Jupiter, Saturn, Uranus and Neptune are the third group which they have hydrogen and helium as their major atmospheric element and beside that Uranus and Neptune have methane in their atmospheric structure. Venus and Mercury are the hottest planets in perspective of surface temperature and Uranus, Neptune and Pluto are the coldest. The temperature for other planets differ “between” 140 to 293 kelvin. Also magnetosphere has important role in structure and composition of atmosphere. All of our solar system planets have magnetosphere except Venus which has an induced magnetosphere. Scientists are not sure about Pluto’s magnetosphere yet.

Without doubt in future we will see new explorations in this time of planet hunting. In my opinion that there is life outside of our blue planet and it is only matter of time for humans to find out smart life. We will reach a time that living on Earth will be impossible and there is no other option except leaving this planet and in order for that to happen we have to find our new home outside of this galaxy.
Figures

1. Mercury’s Atmosphere Composition

![Mercury's Atmosphere Composition](image1)

2. Venus’ Atmosphere Composition

![Venus' Atmosphere Composition](image2)
3. Earth’s Atmosphere Composition

4. Mars’ Atmosphere Composition
5. Jupiter’s Atmosphere Composition

6. Saturn’s Atmosphere Composition
7. Uranus’ Atmosphere Composition

![Pie chart showing Uranus' atmosphere composition with H2, He, CH4, and Other categories.]

8. Neptune’s Atmosphere Composition

![Pie chart showing Neptune's atmosphere composition with H2, He, CH4, and Other categories.]

References

2. [Internet]. Available from: http://hyperphysics.phy-astr.gsu.edu/hbase/solar/venusenv.html
7. What You Happen If You Free Fall Into The Planet Uranus. [Internet]. [cited 2016 Apr 21]. Available from: http://wn.com/what_you_happen_if_you_free_fall_into_the_planet_uranus
Milky Way’s Mysterious Black Hole

Nadja Stanisic

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Astronomy 112

Dr. William Sherry
Milky Way’s Mysterious Black Hole

Black holes are quite possibly the most mysterious things that are in outer space, and new things are learned about them every single day. Black holes come in all different sizes, and form in a few different ways. There are black holes at the center of every galaxy, including our galaxy, which is the Milky Way. The Milky Way is where the planets all orbit the sun, and it is also home to the famous black hole Sagittarius A. This black hole is a supermassive black hole, and will eat anything that comes near it.

Before learning about Milky Way’s Black Hole, it is important to understand what the Milky Way is. So what exactly is the Milky Way and how did it even get that name? The Milky Way is one of the hundreds of billions of galaxies in the Universe, in which our solar system is located. It is made up of a bunch of small, clustered stars which seem to look like cloudy patches. The Milky Way rotates once every 200 million years. In our solar system, the Earth orbits the Sun and is surrounded by a vast galaxy of stars, about halfway out from the center of the Galaxy. The Milky Way got its name because the disk of the galaxy appears to be spanning the night sky like a hazy band of white light, and because the sky resembled spilled milk. If you look at figure A, you can see that the milky way looks like a spiral of color and then has some white fog in the middle. It got its name many years ago, translated from the Latin “Via Lactea” which was then later translated from Greek for “Galaxias”. Astronomers for a long time thought the Milky Way was made up of stars, but they couldn’t prove it. It was finally proven in 1610 when Galileo Galilei looked at individual stars across the sky and realized there were many, many more stars in the sky and they are all part of the Milky Way.

We know our Solar System lies near the galactic plane because the Milky Way divides the night sky into two equal hemispheres. Since there is so much gas and dust that is within the galactic disk, the Milky Way has low surface brightness. Because of this, we can’t see the bright galactic center or see what is on the other side of it, but scientists think there is a supermassive black hole that swallows anything that comes near it. The Milky Way looks like a huge whirlpool barred spiral galaxy, with two spiral arms, called Scutum-Centaurus and Carina-Sagittarius. It measures about 120,000 light years across and is 1,000 light years thick. Density waves that orbit around the Milky Way is what causes the spiral arms on it. A period of active star formation is caused by the density waves moving through an area compressing gas and dust around it. Since we are embedded inside the Milky Way galaxy, it was very hard for scientists to see what it actually looks like, so all pictures are artist’s renditions or simply pictures of other spiral galaxies. Just like how it would be hard to know what your house looks like if you have never been outside of it. We just know what other galaxies look like, so we can estimate pretty well what ours looks like.

Astronomers have estimated that there are between 100 and 400 billion stars in the Milky Way, which each of those stars having at least one planet. That being said, this means there are hundreds of billions of planets located in the Milky Way, and many are believed to be the same size and mass as Earth. The Milky Way is constantly losing and producing new stars. The sun is located in a space between the two arms of the Milky Way, called the Orion Arm. Just like any other galaxy, the Milky Way has a super-massive black hole at the heart of the Milky Way. Scientists have proven that every galaxy has a supermassive black hole right in the center of it. It is known as Sagittarius A, having a diameter of 44 million kilometers, which is 4 million times
the mass of the Sun. Also like other galaxies, the Milky Way is surrounded by a vast halo of dark matter. It is not yet determined what exactly dark matter is, but it accounts for 90% of the Milky Way’s mass (which gives it a mysterious halo), and could possibly determine how fast a galaxy rotates. This means that what we can see with the naked eye without a telescope is only 10% of the Milky Way, and the other 90% is just dust and gas matter. It is also believed that dark matter’s mass is what helps keep the galaxy together as it rotates, and not tearing itself apart. The Milky Way is believed to have formed due to small galaxies colliding very early on in the universe, which still happens. Our galaxy is expected to collide with the nearest galaxy to us, Andromeda, in about 3-4 billion years, long after we are gone. When this happens, it will most likely cause an even bigger black hole, but it is far in the future so there is nothing we know for certain yet. If you look at figure B, you can see that the Milky Way galaxy is near the Andromeda galaxy. There is no other galaxy as close as that one. The Milky Way is almost as old as the Universe, whose age is about 13.7 billion years old. The Milky Way has been around for about 13.6 billion years. We are still learning more and more about our Milky Way galaxy every day and find out new information all the time.

So what exactly is a black hole? It is quite possibly one of the most mysterious and fascinating things in space. A black hole is a place or region in space that has extreme density and such a strong gravitational pull that it eats and pulls in anything that comes near it. Black holes do not suck in objects that come near it though, it acts more like a hole, and objects fall into them. They are extremely massive, but only cover a small region in outer space. For example, a black hole could be the size of a small city, but be able to fit multiple mountains inside. Nothing can escape a black hole, not even light, which is the fastest moving thing in outer space. Because of the strong pull black holes have, scientists can’t exactly “see” them, but they know they exist though because of the radiation that is caused as gas and dust is pulled into them. It is said that if a person fell into a black hole, gravity would stretch them out like spaghetti and they would die instantly. There are three layers to a black hole, the outer horizon, and inner event horizon, and the singularity. The outer horizon is the very beginning of the black hole, where light loses its ability to leave. Once something crosses the event horizon, there is no escaping for that object. The singularity of a black hole is the very inner part of it, where all of its mass lies and is concentrated. If you look at figure C, you can see it out on a diagram of the different layers of a black hole. Black holes can be big or small, with the smallest black holes being as small as an atom. There are three main types of black holes, including stellar black holes, intermediate black holes, and supermassive black holes, which is the one our Milky Way Galaxy has.

Stellar black holes are the smaller versions of black holes, but are still very deadly, and are formed by a star falling in on itself. They have masses less than 100 times the mass of the Sun in our solar system. When a small star has burned its way through all of its iron, and energy production stops the star collapses, and a new core will form and become a neutron star or a white dwarf. But when it is a larger star that collapses, it creates a stellar black hole as it falls in pretty much eating itself. Black holes formed this way are pretty small but are very dense, and have a huge amount of gravitational force pulling in any object that comes close. Black holes get bigger by consuming the dust and gas that surrounds it. Scientists know how to figure out the mass of a black hole, by looking at its gravitational pull it has on a nearby star. The Milky Way has a few hundred million stellar black holes inside it, due to large stars collapsing.
An intermediate black hole is a midsize black hole which used to not even exist. Scientists once believed there were only large and small black holes, but now know that medium sized black holes exist and are formed when stars collide in a chain reaction. If there are several intermediate black holes forming in the same region, they could eventually all come together and go into the center of the galaxy and create a supermassive black hole. Not very many of these types of black holes have been found, but in 2014, astronomers found an intermediate black hole in the arm of a spiral galaxy.

A supermassive black hole are the most common black holes found and are sometimes billions of times as massive as the sun. These are the largest black holes there are. These specific black holes are at the center of almost every galaxy in the universe, including the Milky Way. It is not for certain how supermassive black holes are formed, but they may be the result of thousands of tiny black holes coming together. Large gas clouds collapsing together could also be how these massive black holes are mad. Another theory is from a group of stars all falling together. Once supermassive black holes are formed, they grow bigger and bigger in size because they can easily gather dust and gas around them.

Now that you know what the Milky Way is, and what a black hole is, it is time to learn about Milky Way’s mysterious black hole, Sagittarius A, or Sgr A for short. Like all other supermassive black holes, Sagittarius A lies in the center of the Milky Way galaxy. If you look at figure D, you can see that it lies right in the center of the Milky Way. It is 26,000 light years from Earth and is 14 million miles across, meaning it could fit inside the orbit of Mercury. Although it seems huge already, the amount of mass that can fit inside this massive black hole is enormous. It is more than 40,000 Suns in mass, but it is very hard for scientists to see it because it is very foggy and blurry on an X-Ray. Sgr A will try to eat anything that comes close to it, but it doesn’t always destroy everything. The area around the black hole is a good place for new stars to form, so new star formation will sometimes take place right outside of the black hole. But the stars the black hole has eaten, flares of gas spit out after the stars have been eaten. If you look at figure E, you can see what an actual black hole looks like. The blue is the flares of gas that it is spitting out.

The first person to discover that a radio signal was coming from a location at the center of the Milky way was Karl Jansky. It was discovered on February 13 and 15, 1974. The astronomers Bruce Balick and Robert Brown also helped and they used the baseline interferometer to see it, and the name Sagittarius was made because it was an “exciting” name. Over the last 20 years, there has been enough evidence collected by astronomers by observing the motions of gasses and stars to see that there is something in the middle of our galaxy, a black hole. The ionized gas streamers was the first evidence they needed. Using that evidence and the velocities of the gas, astronomers estimated that a mass of 6 million solar masses are just within 10 arc seconds inside the black hole Sagittarius A. It was unbelievable that this much mass could fit inside such a tiny place, but astronomers were successful in proving it. Now, there is high resolution infrared cameras that can capture a black hole by observing a compact cluster of early-type stars. Most of these stars are moving across the sky very fast. UCLA is where the two main groups devoted to tracking these stars are. Andrea Ghez is one of the observers, and she uses the 10-m Keck telescope. Both groups look at the high spatial resolution and track where the stars lie within the cluster. These infrared images are collected once or twice a year usually. Even though these telescopes have large diameters, they sometimes run into air turbulence. The earth’s
Atmosphere can sometimes blur the images in the telescopes because there are so many molecules colliding with each other. Kind of like how you can see the heat waves coming off the ground when the ground is really hot, it is the same in the atmosphere. Astronomers have to use a device that increases the sensitivity of the observations so they don’t run into this problem anymore. The device is called Adaptive Optics (AO) systems and it uses a mirror to mimic the shape of the light wave that is coming in. If you look at figure F, you can see what an actual adaptive optics telescope looks like. It is crazy to see a telescope that looks like that and is actually that big. It corrects any turbulence that might happen, before the data is even recorded.

There are tons of stars zipping around Sagittarius A at speeds up to 3 million miles per hour, and there are things that help keep track of the motion. There are very accurate stellar positions that are made so that you can keep track of the way the stars move in the central cluster near Sagittarius A. By using Kepler’s laws of motion, it is determined that the positions of the bright stars can be used to show the mass that is in correlation with their orbits. By doing this, this is how scientists figured out that Sagittarius A is 4.6 million times the mass of our sun. Radio emissions show that stars are quickly circling around the supermassive black hole in our galaxy. Recent studies show that galaxies near us are not like ours, so it is unclear how the formation of such a large black hole happens. It is a very mysterious subject.

In 2002, there was something unusual spotted in the images the astronomers took in the center of the Milky Way. There seemed to be a gas cloud G2, which is three times the size of earth that was entering the accretion zone in Sagittarius A. Astronomers predicted that its orbit would get closest to the black hole in 2014, and then be completely destroyed eventually. Some astronomers actually thought this cloud could possibly be hiding a dim star, or possibly a stellar mass black hole behind it. It was said that G2 would have multiple encounters with members of the black hole that are orbiting near the Galactic Centre. The rate of accretion for Sagittarius A is actually pretty small, seeing the mass the black hole has. The cloud then approached the black hole, and astronomers described it as a “flop”. There was then an analysis published that stated the cloud might be a dense clump of matter and is acting as a breeze on the disk of matter. It was originally expected that as they hit, it would cause fireworks, but it didn’t.

The Milky Way is an amazing place, and it is crazy to learn about everything that goes on inside of it. It is home to the earth and other planets that orbit the sun, and also a supermassive black hole. There are different types of black holes, specifically three different ones. There is a stellar mass black hole, an intermediate black hole, and a supermassive black hole, which is the one our black hole in the Milky Way is. The black hole in our Milky Way galaxy is called Sagittarius A, and it eats anything that comes near it. Black holes aren’t actually seen, but we know they are there because we see the gravitational pull it has on other planets. I think astronomers are amazing the way they can learn all this by observing other planets and galaxies.

I believe we are going to keep learning more and more about this topic in the future. We learn new things every day, and new discoveries are going to answer current unanswered questions that we have today. Black holes are something we never really want to come in contact with, because if we got close enough, it would just rip us to spaghetti shreds. But it is fascinating to learn about, and know how complicated they are.
FIGURES

Figure A

Figure B

Figure C

Figure D

Figure E

Figure F
References

http://coolcosmos.ipac.caltech.edu/ask/220-What-is-the-Milky-Way-

http://science.nasa.gov/astrophysics/focus-areas/what-are-galaxies/


http://hubblesite.org/explore_astronomy/black_holes/


https://www.e-education.psu.edu/astro801/content/l8_p7.html
Life Cycle of a Star

Brandon Stemkowski

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Abstract

Stars have existed for billions of years and they are still growing in countless numbers. There are more stars in the universe than can possibly be counted. Stars are giant orbs floating in space that give off heat, light, and energy. A star emits light through nuclear burning. Stars grow in temperature as they get older and the higher the temperature, the brighter they become. A star is formed from interstellar gas and dust that is condensed until the energy given off, heats up enough to create the stars. Hydrogen and helium burning is what fuels the stars. A star can take one of two paths during its life, it can become an Average Star or it can become a Massive star based on how much gas and dust surrounds it. If the star becomes an Average Star, then it will grow to become a Red Giant. After the Red Giant uses up all of its hydrogen and helium, it will collapse and create a Planetary Nebula. After a while the gases that surround the nebula will dissipate back into space and a Dwarf star is what will be left over. If the star goes through the Massive Star path, then it will grow to become a Red Super Giant. As this giant comes closer to its end, its mass will tell whether or not it will become a Neutron Star or a Black Hole. If its mass is less than five times that of the sun, then it will supernova and become a Neutron Star. But if its mass is over five times that of the sun then it will supernova and it will become a Black Hole. Black Holes are terrifying rips in space that nothing can escape from if they get too close. No star or even light can escape from a Black Hole. A lot is still unknown about the Black Hole.

Introduction:

For billions of years, stars have existed, expanded, and continue to grow in countless numbers. There are so many stars in fact, that it would take over 6,000 years to count all the ones that lie in the milky way alone. There are an estimated 100 billion stars in this galaxy, yet there are countless other galaxies spread throughout the wide reaches of space (Becker 2015). The sun that everyone sees in the morning sky is a star as well, a yellow star to be precise. A star is basically a floating ball of gas that gives off heat, light, and energy. Stars give off so much light that they can be seen from lightyears away, which is why they can be seen from earth every night. Stars are not all that simple though. Stars are quite complex in their composition, they go through many phases in their lifetime, they usually go out with a bang, and there is more than one way that a star can die.

Basics of a Star:

The difference between star and a planet is that a star like the sun, “emits light produced in its interior by nuclear ‘burning’. Whereas a planet only shines by reflected light” (What 2014). The sun most people are familiar with is actually only an average star when compared to the many millions in the universe. It is a yellow star, it is still in its youth, and it is about an average size. Stars, as they grow older, tend to grow bigger and change in color according to their temperature, which is one way to make a guess at a star’s age. “The sun (like any other star) is a great ball of gas held together by its own gravity” (What 2014). The sun’s interior produces tremendous amounts of radiation and energy, this is the only reason that the sun does not collapse. Gravity is continuously forcing the sun towards its center and it is this radiation and
energy that the sun produces that counteracts and keeps the sun from collapsing upon itself (What 2014).

A star is formed from an interstellar gas cloud as well as interstellar dust. The gases generally consist of helium and hydrogen. As the cloud gets smaller, it loses some of its potential gravitational energy. This is then turns into heat which a star begins to grow from. As the cloud’s density rises, it gets harder for the heat to escape which causes the center of the cloud to heat up quite quickly. If temperatures inside of the cloud reach their high point quickly enough this will cause a nuclear reaction. As the heat rises, the burning of hydrogen and helium in the cloud will cause a star to finally take its place in space (What 2014).

A Growing Star:

As stated above, a star starts out as nothing more than just a giant gas cloud consisting of helium and hydrogen as well as some cosmic dust floating around in space. This is also labeled as the Stellar Nebula in Figure 1. This gas cloud, as it gets smaller, begins to heat up until it reaches a high enough temperature to become a baby star. It then begins to have nuclear reactions due to the high temperatures and high density and this is how it slowly grows over time.

From this point a star can take one of two different paths. It can take the path of an average star such as the sun. Or it can take the path of a massive star which is exactly what it sounds like, a massive star!

Average Star Path:

The surface temperatures of an average star range from around 2000 degrees Celsius and 30,000 degrees Celsius and the color of the star depends upon the temperature. It is quite obvious that the hotter the star the brighter it will get. Here is the range in color according to temperature: less than 3,700 K (Red), 3,700-5,200 K (Orange), 5,200-6,000 K (Yellow), 6,000-7,500 K (Yellowish White), 7,500-10,000 K (White), 10,000-30,000 K (Blue White), above 33,000 K (Blue) (Ventrudo 2008). “The brightest stars have masses 100 times that of the Sun and emit as much light as millions of Suns. They live for less than a million years…The faintest stars are the red dwarfs, less than one-thousandth the brightness of the Sun” (Life 2015).

The smallest mass that a star could reach would be about 80 times the mass of Jupiter or eight percent that of the sun. If it is anything less, then nuclear reactions could not occur. Anything less than this would glow very dimly and would be deemed a brown dwarf or a large planet (Life 2015).

As the star begins to grow older and lose hydrogen at its center, it starts to swell as the sun will eventually do. At this point the star is now called a Red Giant. A Red Giant is “A large bright star with a cool surface” (Life 2015). They have a diameter around 10 to 100 times that of the sun and are very bright because of how big they are. Although their size is enormous compared to the sun, they are actually only about 2,000-3,000 C, a few thousand degrees cooler than an average young star. Very large Red Giants are also known as Super giants which have
diameters that are about 1 thousand times that of the sun and their luminosity is about 1 million times greater (Life 2015).

The Red Giant is actually the last phase before an average star passes away. As the Red Giant “burns through its fuel, it will eventually collapse. The outer layers will be ejected in a shell of gas that will last a few tens of thousands of years before spreading into the vastness of space” (Taylor 2012). Discoverer of the planet Uranus, William Herschel, named the first Planetary Nebula because of its light green tint and round shape, this reminded him of the planet Uranus (Normandin 2004). Even though he named this object after a planet, it actually shares no relationship with planets at all. The process of a Red Giant forming into a Planetary Nebula is actually quite a gentle process especially when compared to the Massive Star Path. Some examples of Planetary Nebula can be seen in Figure 2. As shown in Figure 2, Planetary Nebula come in a range of different shapes, sizes, and colors. Although they may look quite different from one another, they share pretty much the same qualities and compositions.

The final resting stage of low and medium-mass stars is the White Dwarf stage. A White Dwarf is the husk of what was once a star. It lies at the core of a Planetary Nebula and can be best seen when the gasses that once surrounded it dissipate back into the reaches of space. It no longer has any hydrogen fusion reactions occurring so it basically just lies dormant in space. Over time it will very slowly begin to lose its luminosity. White Dwarfs are approximately the same mass of the sun while being around the same radius as the earth (White 2013). Red Dwarfs, which are stars that don’t make it to the red giant stage and simply burn all the hydrogen left in them, also exist. Although, Red Dwarfs take trillions of years to consume all of their fuel, so long that so far no Red Dwarfs have had the time to transform into White Dwarfs (White 2013). As stated by Taylor Nola from Space.com:

“White dwarfs reach this incredible density because they are so collapsed that their electrons are smashed together, forming what is called "degenerate matter." This means that a more massive white dwarf has a smaller radius than its less massive counterpart. Burning stars balance the inward push of gravity with the outward push from fusion, but in a white dwarf, electrons must squeeze tightly together to create that outward-pressing force. As such, having shed much of its mass during the red giant phase, no white dwarf can exceed 1.4 times the mass of the sun” (White 2013).

White Dwarfs are much more complicated than most other star processes when it comes to gravity.

Massive Star Path:

Alternatively, the other path that a growing star can take is the path of a Massive Star. These stars are much bigger and go a much more exciting path than the Average Stars. Massive Stars have “Masses greater than 9 times the sun” (Khan 2010). Massive Stars start out pretty much the same way normal stars do, with a large cloud of hydrogen and helium as well as a little bit of cosmic dust. The cloud is obviously larger than the gas clouds that form a normal star. As the star begins to condense and become more and more dense, the center will ignite to start the star’s life. Even at this stage, since the star is much bigger than your average star, the hydrogen will burn much faster and much hotter (Khan 2010). A Massive Stars lifespan is also actually
significantly shorter than that of a normal star. A normal star will usually survive around “ten or eleven billion years…here [they survive about] ten million years” (Khan 2010). Eventually the helium will form into a helium core with a hydrogen shell, and as more helium is produced, the core becomes denser. The denser the core becomes the more pressure that is put on the hydrogen shell, and in turn, the hydrogen shell releases much more outward energy to increase the radius of the actual star (Khan 2010). This is how an Average star grows to be so much more than average. From this point, the Massive Stars will start its growth process to become a Red Super Giant.

Betelgeuse is the most well-known example of a Red Super Giant. As stated by HyperPhysics.edu, a Red Super Giant’s “Absolute luminosities may reach -10 magnitude compared to +5 for our sun” (Kaufmann 1991). For example, Betelgeuse’s luminosity is almost ten thousand time the Sun’s and its radius is almost 370 times that of the sun (Kaufmann 1991). It is also worth noting that the path of an Average Star is usually following the main sequence, while the path of a Massive Star will usually stray from the main sequence. The main sequence is a continuous band of stars with a range of color rather than brightness on the Hertzsprung-Russel diagram. The Hertzsprung-Russel diagram as well as the main sequence can be seen below in Figure 3.

A Red Super Giant is created when fusion stops and gravity takes over, condensing the star smaller and tighter. As the star contracts the temperatures get much hotter until the helium inside the star is able to fuse into carbon. If the mass of the star is of a great amount, then the burning helium could begin with an explosive flash (Red 2013). The energy produced in this helium fusion is what causes the star to expand outward and grow even larger into a Red Super Giant. Temperatures of a Red Super Giant usually only reach around 2,200 to 3,200 degrees Celsius, which is only a little over half as hot as the sun (Red 2013). Taylor Nola states in her article Red Giant Stars that “This temperature change causes stars to shine in the redder part of the spectrum, leading to the name red giant, though they are often more orangish in appearance”. These Giant stars will only survive for a few thousand years to one billion years before they move to their next phase. And from here the Red Super Giant can take one of two paths. It can become a Neutron Star or it can become a Black Hole. But before it can take either path, the Red Super Giant has to Supernova.

A Supernova is the very last action a star makes before it becomes either a Black Hole or a Neutron Star. A Supernova is really basic, it is nothing more than the explosion of a star and it is the biggest explosion that will ever happen in space. Supernovas happen in every galaxy but it is actually quite difficult to see them in the milky way because of the cosmic dust that blocks the view (May 2015). As Susan May from the NASA website explains, “In 1609, Johannes Kelper discovered the last observed supernova in the Milky Way. NASA’s Chandra telescope discovered the remains of a more recent supernova. It exploded in the Milky Way more than a hundred years ago”’. A Supernova can happen from one of two things. A star can either end its life and explode which is the easy way. The more complicated cause of a Supernova happens in binary star systems. A carbon-oxygen white dwarf begins to steal matter from its companion star until it is overloaded with matter and explodes, causing a Supernova (May 2015). Supernovas actually have a lot to teach about the universe. This is why scientists study them, in-fact one kind of Supernova has taught that the universe is ever expanding at an increasing rate. They also play a key role in distributing elements throughout the universe (May 2015). Finally, after the biggest explosion in the galaxy takes place, it is time to take a look at the results.
Neutron Star:

A Neutron star is the first possible end of a Red Super Giant. A Neutron star is created when the Red Giant undergoes a supernova explosion. As stated on NASA’s website, “This explosion blows off the outer layers of a star into a beautiful supernova remnant. The central region collapses so much that protons and electrons combine to form neutrons. Hence the name neutron star”. Neutron stars can appear as isolated objects or they can appear as binary systems that have planets surrounding them. There are only four known neutron stars that have planets and these binary systems are the only ones with which scientists can measure the mass of these stars (Neutron 2008). The pressure must be at most five solar masses for the neutrons to support a star. Anything more and the star will collapse into a black hole (Cosmos). Neutron stars are extremely dense! In fact, a teaspoon of neutron star material would weigh around a billion tons (Cosmos). Neutron stars rotate super rapidly due to angular momentum and have extremely strong magnetic fields because of the conservation of magnetic flux. A good example of how this works was stated on the Cosmos website as it was stated that the rotation is kind of like an ice skater increasing her spin by bringing her arms closer to her body.

Black Hole:

Even today, Black Holes hold a lot of mystery to astronomers and scientists. They are deep space objects that are very mysterious as well as very dangerous, not even light can escape from a Black Hole. A Black Hole is created when a star that is about five times the mass of our sun, collapses. These stars will use the fusion process to eat up every element on the periodic table; silicon, aluminum, potassium, until it reaches iron. No energy is produced by fusing iron atoms since it is the stellar equivalent of ash (Cain 2013). As Fraser Cain from Universe today explains “in a fraction of a second, the radiation from the star turns off. Without that outward pressure from the radiation, gravity wins out and the star implodes. An entire star’s mass collapses down into a smaller and smaller volume of space…The velocity you would need to escape from the star goes up, until not even light is going fast enough to escape” this is how a Black Hole is formed. One other way you can get a Black Hole is when two Neutron Stars collide. The idea behind Black Holes actually follows from Albert Einstein’s Theory of Relativity. This states that “the strength of a celestial object’s gravitational field is determined by the density of matter it contains: the higher the density, the stronger the gravity” (Clark 2016). Nobody really knows what happens when an object enters a Black Hole but it is speculated that the object once it passes the event horizon, it is crushed out of existence. On a side note, the event horizon is the boundary of no return, once an object crosses the event boundary it will be in the grasp of the Black Hole forever (Clark 2016). Black Holes have zero volume and infinite density, which baffles scientist mainly due to the fact that it seems impossible. Quantum Theories of Gravity are the only thing that could possibly explain Black Holes but for now, they are only theories.

Alternate Method of a Stars Death:

No star can live forever, no matter which path a star takes, where there is life, there will always be death. But imploding into a supernova and becoming a Black Hole, condensing into a
Dwarf Star, or imploding into a Neutron Star are not the only methods that a star can end its life. There is actually one last way that a star can die, and that is Black Holes. To reiterate how dangerous and terrifying Black Holes can be, they are the only thing in the solar system that can destroy a star without the star going through its life cycle. Anything that gets too close to a Black Hole is immediately caught and eventually destroyed, even stars. Refer to figure 4 for an example of a Black Hole eating a star.

Conclusion:

When you get down to it, stars are some very complicated alien objects. But they can be quite easily simplified. If you want to get down to the very basics of stars, they are large glowing orbs in the sky that are the result of solar gas and interstellar dust. They age just like every living thing does, and they die just as every living thing does. I would not go as far to say that we know every little thing there is to know about stars but we know an awful lot. Stars, when looked at closely, can be ripped apart into many small pieces that we can observe. We know the temperature of stars, the luminosity, the radius, the diameter, how they are born, and the many different ways in which they can die. But we still have many questions. How did all the stars come about? Where did they all come from? How many stars are there in the universe? These are questions that can only be answered with time. As I have demonstrated, stars go through many different phases and can take a variety of paths throughout their lives. Stars come from solar gas and dust. Depending on the amount of gas and dust they are surrounded with, they will either take the path of an Average Star or they will take the path of a Massive Star. If they take the path of an Average star, then they will grow to become a Red Giant until their death when they become a Planetary Nebula, and then eventually become a Dwarf Star. But if they take the path of the Massive Star, then they will proceed to becoming a Red Super Giant, and after it’s death, it will supernova and become either a Black Hole or a Neutron Star. It’s quite simple when you want it to be. But as you can see, stars are quite complex in their composition, they go through many phases in their lifetime, they usually go out with a bang, and there is more than one way that a star can die. In time, more will be discovered about these massive space objects, but for now all we can do is appreciate them in all their magnificent glory.
Figures

Figure 1

http://www.schoolsobservatory.org.uk/astro/stars/lifecycle

![Life Cycle of a Star](http://www.schoolsobservatory.org.uk/astro/stars/lifecycle)

Figure 2


![Planetary Nebula](http://www.universetoday.com/wp-content/uploads/2013/10/plneb.jpeg)
Figure 3
http://www.universetoday.com/wp-content/uploads/2013/05/hr-diagram-schematic.jpg

Figure 4
http://www.nasa.gov/sites/default/files/cygx1_ill.jpg
References


Electricity in the Human Body

Kindle Stewart

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Dr. Durandet
ABSTRACT:

Electricity is all around us in today’s world, but it may appear in some aspects that most individuals may not recognize. It is quite easy to see evidence of electricity in appliance like a dishwasher, because they are plugged into an energy source. Many people would not think of the human body as a circuit that conducts electricity because it does not have to be charged by an outlet. However, the human body gets energy by the calories it receives from the food it consumes and breaks down, and uses that chemical energy to perform everyday functions that we typically do not even think about like breathing or maintaining heart rhythm.

When most people hear the word “electricity”, they think of wires, appliances, and light switches. According to GCSE Physics, the physics definition of electricity is the flow of electrons from positive to negative in terms of a circuit with a power source. The flow of electricity throughout a circuit is measured in amps, while the amount of “power” being put into the circuit is measured in volts (i.e. the electrons flowing from a battery). It is rather easy to correlate electricity with a hair dryer, washing machine, or flashlight since the result is something that can be measured quite easily when the appliance turns on or light is produced. However; few people would list the human body among things that use electricity. Just like the hairdryer receives energy from an outlet and a flashlight from a battery, the human body receives energy from the food it consumes. The unit for energy is a calorie, which would explain why the food we eat is measured in them. Within the human body lies a circuit, and this circuit is composed of the central and autonomic nervous systems that both send and receive electrical impulses and signals throughout the body. The central nervous system consists of the brain and spinal cord, and is responsible for sending sensory information to the brain and then taking the appropriate action based upon that information. The central nervous systems has millions of cells that are used specifically for transporting information from one area of the body to another, these cells are called neurons. There are sensory neurons and motor neurons in the central nervous system that help control movement and feeling in throughout the body. According to Dr. Tom Virden, this information is transmitted between neurons by an electrical signal or impulse. For example, when your hand is placed on a hot stove, the sensory neurons in your hand send an electrical signal to your brain, which then sends an electrical impulse to the motor neurons to pull your hand away from the hot stove.

How do the neurons send this electrical signal? Well, Figure 1 shows the membrane of a neuron. This membrane consists of tiny hole-like structures, known as gates, that regulate what may enter the cell as well as what may leave. Mainly, the gates let ions (positively or negative charged particles) in and out of the cell. Dr. Tom Virden explains how neurons generate an electric charge and release that charge as a signal to other parts of the human body. The cells in the body contain sodium on the inside and are generally surrounded by potassium on the outside. When a neuron is resting, it is pumping out the positively charged sodium ions and in turn, letting in potassium ions. During this process, the neuron is letting in less positive charge than it is getting rid of, making the surrounding area more positive on the outside of the neuron and the inside more negative. As long as the neuron holds a negative charge in this particular instance, this is referred to as “resting potential”, and the neuron holds a charge of -70 millivolts. When the neuron is receiving information from a particular source, however, there is a slight voltage jump in the positive direction that brings the charge of the neuron from -70 millivolts to -55
millivolts; this is known as depolarization. Once at -55 millivolts, the neuron is at what is referred to as the “threshold stage” where the neuron becomes active. The gates of the membrane of the neuron open and sodium ions rush into the cell. While this is happening, the cell is becoming more and more positive, so the potassium ions begin to pour out because of repulsion between the two types of ions. This process increases the voltage of the neuron from -55 millivolts to +30 millivolts, a rather large voltage increase of 85 millivolts from such a small cell. This voltage increase creates an electrical signal that will discharge to other neurons down a nerve until it reaches its intended destination. This is easier visualized in Figure 2. Now that potassium ions are surrounding the neuron and sodium ions are within it, the process repeats as it begins to pump out sodium ions and take in potassium at the same time.

The autonomic nervous system is different from the central nervous system as it controls bodily functions on its own and does not require a thought to do them. The autonomic nervous system is in charge of things like specific organ functions such as digestion, breathing, and your heart beat. The heart is a very complex organ that sustains the entire body by supplying blood and oxygen to every part by blood vessels. One of the most fascinating things about the heart is its rhythm, known as the heart beat or pulse. The heart “beats” from a series of electrical impulses that cause the muscle tissue to contract, causing the blood to leave the heart and travel through the blood vessels to other parts of the body. Once again, the electricity of the heart is maintained by neurons that carry those signals down nerves to other neurons that will then cause the heart to contract in that specific area.

Scott from The Medicine Journal explained how neurons of the heart work together to communicate and effectively maintain the “ideal rhythm” of the heart. The Sinoatrial (SA) node plays a major role when it comes to the electricity within and around the heart. The SA node is made of groups of neurons that work together to generate the perfect rhythm, which would explain why they are referred to as “the pacemakers of the heart”. In reference to the heart in particular, cells are surrounded by and contain electrolytes. While an electrolyte is a rather commonly used term, rarely is it explained. Common electrolytes in the human body are sodium, potassium, chloride, magnesium, calcium, and phosphorus; with potassium typically found within cells while sodium and calcium generally reside outside of the cell. The membranes of these cells in the SA node form a layer of protection, so to speak, between the interior of the cell and the exterior that is surrounded by electrolytes. The blood pressure of the body causes the sodium ions to enter the cell while causing potassium to leave. Since more sodium is coming in than potassium is leaving, the cell contains a highly positive interior and is therefore considered to be positively charged. Does this mean that a higher blood pressure in an individual would cause their SA nodes to be more positively charged? Actually, no, it wouldn’t. This is because that when the cell reaches a determined peak of positive charge, the calcium channels of the cell open within the membrane. Now that there is no longer a barrier between the interior and exterior of the cell, there is a higher concentration of positive ions within the cell than there is outside of it; this creates a potential. Once the potential is high enough, there is enough “energy” to travel across neurons down nerves and into nearby muscle tissue causing the heart to contract. This process of releasing the charge of the SA node cells to cause the heart to beat is known as discharging.

It only takes a few milliamps to throw off the electrical rhythm of the heart. What happens if the heart rhythm is off, even slightly? Figure 3 lists some particular values for the amount of amps a person can handle without sustaining significant consequences as well as more permanent ones. The table shows these values as milliamps, which is 1/1000 of an amp, a very
small value. This shows that it does not take much completely throw off the natural electrical rhythm of the heart and have dire results. Figure 3 shows that around 60 milliamps is the threshold for the heart to endure and maintain normal fibrillation (rhythm), and higher than that will cause someone to have an non normal rhythm, referred to as defibrillation (defib). However, an imperfect heart rhythm does not mean that all hope is lost for the individual. There is a machine called an Automated External Defibrillator (AED) that will deliver a “shock” of current to the individual in hopes of regaining normal rhythm.

According to Boston Scientific, most typical AEDs are composed of three parts: “a high-voltage source, a capacitor, and switches.” A capacitor is part of a circuit that holds charge for a certain period of time, this is one of the key components to an AED as it is able to release that charge all at one time and does not require much time to regain that charge. The switches in the AED are also quite important, as they prevent it from releasing the charge until it is needed. Without a switch, the AED would be constantly releasing charge, and therefore would be ineffective at restoring normal heart rhythm. The AED is a very useful tool in the medical field, but it cannot restore the electrical current in the heart once it has stopped beating. This is why there is cardio-pulmonary resuscitation (CPR). Performing CPR helps move blood to the heart, letting the sodium ions leave the cells and the potassium to enter. CPR is performed in hopes that enough SA nodes will discharge and cause the heart to contract at the appropriate time to regain rhythm with (hopefully) no damage to the brain or any other internal organs.

CONCLUSION:

The research I conducted really helped me understand that physics is actually everywhere, even within me. Applying physics to a field I am genuinely interested in really helped me acquire more knowledge about electricity and be able to dig deep for this paper. The most fascinating part about my research was finding out how our bodies generate electricity in the first place. In biology I always learned about neurons send electrical signals to the brain and receive them as well, but I had no idea how that actually worked. As for the heart, I knew that a normal heart rhythm had to be maintained to stay alive, but I didn’t really know the actual physics. I was unaware of how that electricity was generated in the heart to make it contract on queue, but now that I have done research on how everything in the body comes together and functions as a single unit by electricity, I understand how sensitive it is to outside electrical interference.

Another thing I learned from performing this research was how Automated External Defibrillators worked. In class we briefly mentioned AEDs and how they held charge like a capacitor, so I really wanted to discuss them on a deeper level and get into the details of the electrical physics aspects. Overall, I learned that electricity in the human body is a really spectacular concept to wrap your mind around. How everything is able to communicate to other parts of the body is incredible, that a simple impulse of electricity from the brain can move your hand away from a hot object or how a discharge of electricity can make your heart beat flawlessly and in sync. Performing this research has helped me gain a much deeper understanding of the physics behind the electricity in the human body.

I have always seen physics as just a “stepping stone” class for me, but I really think that it has been one of the best classes I could have taken. In both parts of physics courses I’ve taken, I’ve been able to visualize what is happening, whether it be on a macro or microscale, and
determine an outcome. This research has helped me visualize exactly what happens to the cells of the body as electrolytes and ions move through them, causing electric potential to do whatever that cell is designed to do.
FIGURES:

1. This is a diagram of a neuron. The cell membrane consists of tiny “flaps” that open or close in order to control the intake or release of particles.

2. This figure demonstrates the neuron on the left transferring an electrical signal to the neuron on the right. This process will continue down a chain of several neurons.
3. This table shows the amount of current required to cause different reaction in the human body and compares it to that of a common fuse breaker in a circuit typically found in a home.

<table>
<thead>
<tr>
<th>Current (mA)</th>
<th>Effect</th>
</tr>
</thead>
<tbody>
<tr>
<td>1</td>
<td>Barely perceptible</td>
</tr>
<tr>
<td>16</td>
<td>Maximum current an average man can grasp and &quot;let go&quot;</td>
</tr>
<tr>
<td>20</td>
<td>Paralysis of respiratory muscles</td>
</tr>
<tr>
<td>100</td>
<td>Ventricular fibrillation threshold</td>
</tr>
<tr>
<td>2 A</td>
<td>Cardiac standstill and internal organ damage</td>
</tr>
<tr>
<td>15/20 A</td>
<td>Common fuse breaker opens circuit†</td>
</tr>
</tbody>
</table>

*From NIOSH.† Conid with 20 mA of current can be fatal. As a frame of reference, common household circuit breaker may be rated at 15, 20, and 30 A.
REFERENCES:


Praise the Sun! A Scientific Examination of the Sun

Max Swift

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AST 112 12666
Dr. William Sherry
Abstract

This paper reviews the history of the Sun and the Sun’s cultural and scientific impact on Earth and humanity. It will also go in-depth on the Sun’s anatomy and the chemical reaction that creates the Sun’s glowing and always present fire. Specifically, the Sun is composed of the core, radiative zone, convection zone, photosphere, chromosphere, and corona which make up the anatomy and work to ember the Sun’s glow. To further gain appreciation of the Sun, the paper will compare the Sun to other stars in the universe by examining the Sun’s distance, size, temperature, and age. Finally, this paper will use modern information to properly determine when and how the Sun will ultimately die. All of these factors will hopefully grant accurate insight into the science of the Sun.

Introduction

What accurately symbolizes life? Food and water quickly come to mind but there is something else that is incredibly potent and a harbinger of life. This engrossing object is the ubiquitous Sun. In the grand scheme of the expansive universe, the Sun is a regular star in a sea of millions but to humanity the Sun is easily the most significant star to ever exist. Neanderthals and even modern scientists have stared at the Sun in amazement and captivation. Clever marketers seem to love to use the Sun as a logo or marketing tool while everyday citizens enjoy absorbing the Sun’s pleasant and warm gentle rays.

The cultural and scientific effect the Sun has continuously shined on humanity is uncontrollable. The Sun’s bright light gives the life on Earth the strength to survive and thrive, but what is the Sun made of and how does the Sun continually spawn this aura of flames? Understanding this requires investigation into the Sun’s layers such as the core, radiative zone, convection zone, photosphere, chromosphere, and corona. What makes the Sun similar yet so different than the other millions of stars in the vast universe? Examining the Sun’s distance, size, temperature, and age should create a platform for fair comparison between other stars. Finally and most importantly, when and how will the independent and influential Sun die off? Scientifically answering and exploring these questions will ideally generate positive insight and appreciation of Earth’s best ally, the Sun.

Anatomy

Although the Sun appears to be a simple sphere, it is composed of many unique and complex parts. Just like its friendly star brothers and sisters, the Sun is a giant mass of neutrons, protons, and other subatomic particles [2]. Similar to an onion or everyone’s favorite ogre Shrek, the Sun has several layers. The core, radiative zone, convection zone, photosphere, chromosphere, and corona are all the significant layers that in unison form the Sun.

The power of the Sun comes from the powerful nuclear reactions occurring at the inner layer known as the core [2]. The main goal of the Sun’s core is creating energy with thermonuclear reactions which then creates intense temperatures deep inside the Sun’s core [1]. On a molecular level, hydrogen atoms fuse together to form helium atoms because of the incredible pressure in the core which then creates energy [2]. The temperature around the core steams up to about 15 million degrees Celsius [2]. If the mitochondria is the powerhouse of the cell, then the core is easily the powerhouse of the Sun.
The next layer from the core is the radiative zone. In this layer, energy leisurely moves out taking around 170,000 years to radiate through the entirety of this layer [1]. Both the core and radiative zone pack gas particles so compactly that the overall gas, or plasma, can hardly travel [2]. In these layers, solar energy must move by radiation [2].

Creeping away from the innermost regions, is the layer known as the convection zone. It is in the convection zone, that energy continues its long trek outwards towards the surface through convection currents of the heated and cooled gas [1]. Unlike the core and radiative zone, the convection zone is more spacious triggering gas particles to collide less and to collide softer which causes the temperature to drop [2]. The convection zone operates like a middle man between the inner zones creating the power and the outer zones exposing the power. The name convection zone most likely comes from the fact that energy in this layer travels towards the surface through convection currents of the heated and cooled gas [1].

After the convection zone, comes the photosphere which is the Sun’s familiar surface. The photosphere is a very thin layer of magnetically charged plasma that has a decreased temperature of 6,000 degrees Celsius [2]. In total, the photosphere is 60 miles thick and is what the average viewer notices when exposed to the Sun [2]. More scientifically, the photosphere is rampant with magnetic activity with some regions being centers of positive polarity while other regions are centers of negative polarity [2]. Magnetic field lines link these regions of positive and negative polarity and when these regions shift the field lines break and reform which can produce violent storms [2]. Imagining a hunk of flaming energy erupt from the sun into one of these storms must seem terrifying yet eerily beautiful simultaneously. The photosphere is also home to sunspots [2]. Sunspots are regions where temperatures are cooler because of strong magnetic activity, and these regions thus appear as dark spots because of how they contrast against the intensely bright photosphere [2]. Even the Sun has blemishes.

Creeping out from the photosphere, is the chromosphere. Similar to the photosphere the chromosphere is a very thin layer; however, it holds magnetic field lines which hold back electrically charged solar plasma [1]. Solar flares, coronal holes, coronal mass ejections, and other occurrences which blast hot plasma outwards all happen on the surface of the chromosphere due to the magnetic activity [2]. These plasma features are called prominences and can sometimes even fling material away from the Sun [1]. The chromosphere is where the Sun truly shows off its theatrical and uncontrollable power and ferocity. Regarding temperature, the inner edge fluctuates around 100,000 degrees Celsius while the outer perimeter hits temperatures around several million degrees [2].

Finally, there is the corona which is the outermost region and branches off millions of miles into space [2]. Inside the corona, are ionized elements which glow in the x-ray and extreme ultraviolet wavelengths [1]. The temperature in the corona is continually changing which makes the power of a breeze of solar particles that blows out in all directions also known as a solar wind change as well [2]. The corona completes the basic anatomy of the Sun and gives a basic scientific understanding of how the Sun operates and produces its glow. Figure 1 shows and labels all covered layers from core to corona.

The Sun Compared to Other Stars

When comparing the Sun to other stars that inhabit the universe, the most visible difference is visibility itself. From Earth, the Sun is close, in fact, it is very close. Most people know that Sun is the closest star to Earth, and in numbers, it is only 93 million miles away [3].
Alpha Centauri, three stars bound together by gravity consisting of Alpha Centauri A and B and Proxima Centauri, is a triple-star system that is nearest the Sun [3]. The Alpha Centauri A and B are 4.37 light years away from the Sun while Proxima Centauri is the closest star besides the Sun at 4.24 light years away [3]. This puts into perspective how truly close the Sun is to Earth and creates one of the Sun’s most distinguishing features from its star peers. Besides distance, other factors that should be considered to gain a better insight on the Sun are size, temperature, and age.

The Sun is constantly being described as massive or large, but to truly understand its size proper numbers should be viewed first. There is only a 6.2 mile difference between the equatorial diameter and polar diameter [4]. The Sun’s mean radius is 434,450 miles and this forces the diameter to be around 864,938 miles [4]. The overall circumference of the Sun is around 2,713,406 miles meaning that 109 Earths could be lined up along the face of the Sun [4]. When examining size, volume should also be properly accounted for. The Sun’s volume is $1.4 \times 10^{27}$ cubic meters which means 1.3 million Earths could potentially fit inside the Sun [4]. The immense volume of the Sun creates an even more intense mass. The Sun’s mass clocks in at around $1.989 \times 10^{30}$ kilograms which when compared to Earth is 333,000 times the Earth’s scrawny mass [4]. These numbers are huge and easily dominates the solar system. However, in truth, the Sun is an average sized star [5]. The smallest stars are less than $1/10^{th}$ the size of the Sun while the largest stars are more than 100 times more massive than the Sun [5]. The Sun is characterized as a G-type main-sequence star or a yellow dwarf [4]. Compared to the red giant known as Betelgeuse, the Sun is totally outclassed as Betelgeuse is about 700 times bigger [4]. Although the Sun demonstrates impressive numbers, it is honestly just an average star in regards to its size and mass.

One of the defining features of any star is the brilliant heat and temperatures they generate. From Earth, the heat from the Sun feels perfect except for perhaps in Arizona. Interestingly, to determine the temperature of a star the feature that provides the most evidence for this is color. For simple analyzing, astronomers have invented a classification scheme with seven categories known as OBAFGKM with each letter being subdivided into 10 subclasses [6]. Stars with the classification of O0 have temperatures of 40,000 K and are the color blue [6]. Stars with the classification of B0 have temperatures of 20,000 K and are the color light blue [6]. Stars with the classification of A0 have temperatures of 10,000 K and are the color white [6]. Stars with the classification of F0 have temperatures of 7,500 K and are the color yellow-white [6]. As stated above, the Sun is classified as a G-type star or yellow dwarf [4] and stars with the classification of G0 have temperatures of 5,500 K and are the color yellow [6]. Below G0, there is K0 which have temperatures of 4,000 K and are the color orange and there is also M0 which have temperatures of 3,000k and are the color red [6]. Viewing these classifications, the Sun achieves temperatures of 5,500 K while at the top are blue stars which reach 40,000 K [6]. Once again, the Sun achieves simple mediocrity compared to the galaxy’s other stars, or put more appropriately, the Sun generates average temperatures.

Finally, the Sun’s age should be evaluated to determine if it has any seniority over fellow stars. The first mystery in this process is how astronomers determine star age in the first place. One way astronomers determine a star’s age is by examining the star’s spectrum, luminosity, and motion through space [7]. Once this information is carefully collected, astronomers can create a star profile and compare the star to prior models which display what appearances stars should have at various points of their evolution [7]. Since this method relies on the models it can be
inaccurate, so there is a new technique which has been created [7]. This newer technique of determining star age is called gyrochronology and is based off a star’s rotational speed [7]. The speed of a star’s rotation is constantly fluctuating and is reliant on a star’s age and color [7]. So astronomers that know a star’s color and rotation speed, can accurately determine a star’s age with an uncertainty of only 15% [7]. After using all of these fancy techniques, what have astronomers approximated the Sun’s age to be? In total, the Sun has recently blown out the candles for its 4.6 billion year birthday [7]. From humanity’s perspective, 4.6 billion years is an insane amount of time that is almost unfathomable. Surprisingly, the Sun is still relatively young since it will take around 7 billion years until it turns into a red giant [7]. Astronomers using NASA’s Hubble Space Telescope have found a star which may be around 14.5 billion years old [8]. However, this does create an interesting predicament since the universe itself has an age of 13.8 billion years [8]. Ultimately, in regards to age, the Sun once again is plainly average. It is not the youngest nor the oldest in the universe and it appears to be partially close to hitting its midlife.

Overall, the Sun appears very average. Its most unique feature is how it sits in the solar system perfectly complimenting Earth and providing the proper amount of heat. With size, temperature, and age the Sun does not hold any amazing records and is simply a regular star. Fortunately, this regularity is probably beneficial as it provides Earth with a dependable life source.

When and How the Sun Will Die

The beauty and also distress that arise in life often are byproducts of time. The dimension of time is ultimately temporal, so life is constantly being born yet also dying. Stars are no exception to this concept. Every star experiences a birth and also a death. The death of a star that is the most significant for humanity is obviously the Sun.

It will take 7 billion or even 8 billion long years before the Sun transforms into a red giant [9]. As a red giant nearing death, the Sun will eat up and evaporate the Earth [9]. To understand how stars reach this state it is important to examine a star’s humble beginning. In the beginning, a star is simply a large agglomeration of gas composed mostly of hydrogen and some helium [9]. The gas, when put into one place, starts to collapse on itself under the weight [9]. This causes the proto-star to be pressured which in turn causes the gas to heat up [9]. The gas gets continually heated up until there is so much heat that electrons get stripped off the atoms and the gas is charged [9]. Helium is then formed because the hydrogen atoms which contain a single proton fuse with the other hydrogen atoms which creates helium with two protons and two neutrons [9]. Afterwards, this fusion lets out energy in the shape of light and heat [9]. This then creates outward pressure and the gas finally ceases from collapsing any more [9]. This long recipe is how a new star is born.

Interestingly, the process for birthing a star follows a similar path as to how a star dies. The process stars go through requires a sizeable supply of hydrogen. Luckily, stars have enough hydrogen to keep the process continually occurring for billions of years [9]. However, there is a point when all of the hydrogen has finally turned into helium [9]. When this occurs, the Sun will have trouble generating the same amount of energy as before, and unfortunately, begin collapsing on itself [9]. Even the helium core will begin collapsing on itself and this causes a release of energy because of the increased pressure [9]. This release of energy causes the Sun to have an increase in brightness but also causes the Sun to become a red giant [9]. It is a red giant.
because the surface temperature becomes lower than a star such as the Sun, but it is also much bigger [9]. The day the Sun becomes a red giant will not be a happy day for the unprepared solar system. When the Sun’s outer core hydrogen finally depletes, the Sun will experience collapsing again [9]. The last pieces of helium will eventually change into heavier elements; however, there no longer is radiant energy to keep the Sun pushed up against its weight [9]. The core of the Sun will decrease and turn into a white dwarf [9]. Once the core collapses, the Sun’s outer layers of the atmosphere will be promptly separated and left behind which will result in a magnificent planetary nebula [9]. Figure 2 shows an image displaying a planetary nebula. The appearance of a planetary nebula looks eerie yet also strangely captivating.

The idea of the Sun eventually dying off is a depressing and scary thought, but it is a natural process that will be difficult to avoid. As stated above, it will take the Sun 7 billion or even 8 billion years before turning into a red giant [9]. With such a giant period of time, the Sun’s demise is an event which people should not realistically become worried or upset about. The day the Sun does finally die off it must be an amazing spectacle. Watching the powerful and unmoving Sun collapse on itself and branch out engulfing parts of the solar system, should be history’s most theatrical and magnificent event ever.

Conclusion

It is truly amazing how a gaseous ball out in the distance has created and impacted so much. At first glance, the Sun appears a simple sphere but the above research shows that it is much more complex. The Sun’s anatomy is composed of an intricate layer system which respectfully includes the core, radiative zone, convection zone, photosphere, chromosphere, and corona. Each layer has a designed purpose and contributes towards the Sun’s structural integrity. When comparing the Sun to other stars, it appears very average and one of its most personal features is its perfect placement in the solar system. It is a G-type main-sequence star or yellow dwarf with a respectful radius of 434,450 miles and a branching diameter of 864,938 miles [4]. The Sun also creates steamy temperatures around 5,500 K [6] and is only 4.6 billion years old [7]. The end of the Sun’s life is eventually and inevitably coming but certainly not soon. It will be around 7 billion or even 8 billion years before the Sun becomes a red giant [9]. After that, the Sun will collapse into a white dwarf and finally create a planetary nebula [9].

It is easy to see why the Sun has so much significance as a cultural and scientific mascot. It truly represents a strong, dependent backbone for the Earth to gently orbit around and flourish from. The Sun’s bright light literally and figuratively makes all other stars difficult to view. Most people love the Sun and it is easy to understand why. It gives people the warmth and light of life but also the innate ability to inspire and motivate others. When the Sun is out, is when the day begins and everyone’s thoughts and ideas have a platform to become realities. The Sun has developed into more than just a star.

Hopefully in the near future, astronomers can acquire an even more accurate and closer examination of the Sun and learn about more parts of the Sun’s anatomy to further connect the dots on how the Sun operates. Astronomers finding and viewing other stars in the universe should also be beneficial because it offers more examples to compare to the Sun to gain better insight on the Sun and even the universe. Finally, may be the future has astronomers and other scientists learning and brainstorming ideas on how to proactively preserve the dying Sun and create more accurate and realistic simulations of the Sun’s impactful and magnificent death. Whatever the future holds, the Sun will be there brilliantly shining on.
Figure 1

Figure 2

Cited References


6 University of Nebraska-Lincoln. Spectral classification of stars [internet]. Astronomy Education at the University of Nebraska-Lincoln; [cited 2016 April 19] Available from http://astro.unl.edu/naap/hr/hr_background1.html


Wormholes

Harrison Taylor

11/19/15

Physics 112

Dr. Durandet
Abstract

Wormholes are a theoretical and elusive topic that has sparked much debate. Should they exist, wormholes would provide a fast route for transportation through space, though potentially very dangerous if the right precautions are not taken. These bridges could even hold the key to time travel, and even travel between universes. This report will explore what wormholes are, wormhole theories, such as quantum entanglement and time travel, as well as a relatively recent study in which the illusion of wormhole based travel by a magnetic field was observed. Einstein’s theory of general relativity will also be discussed as it plays an integral role in the concept of wormholes.

Introduction

Wormholes have been used in many science fiction stories and have been the subject of many scientific studies. To be able to utilize a wormhole and have a way to travel billions of light years through space in a short amount of time would be very helpful in the pursuit of knowledge about the universe. Not only is it theorized that they would allow for long distance space travel, but it is also thought that they could be bridges to other universes, as well as a means for time travel. It is an exciting time in quantum physics to be able to work towards, not only discovering new possibilities as to how a wormhole could exist and behave, but also how to marry quantum theory with classical theory, which is what is has been causing problems with a lot of wormhole hypotheses.

What Are Wormholes

A wormhole can be described as a bridge through space. A bend in the fabric of space and time that would allow for matter to enter at one end, travel through the wormhole, and come out the other end at some other point in space that could have potentially taken light years to reach via conventional methods of space travel, or even a way to travel from one universe to a completely different universe, as well as time travel. “A wormhole is a theoretical passage through space-time that could create shortcuts for long journeys across the universe” (Redd 2015). There are, however, some dangers that trying to travel through a wormhole could present, such as breakdown of the wormhole itself and very intense radiation. “But be wary: wormholes bring with them the dangers of sudden collapse, high radiation and dangerous contact with exotic matter” (Redd 2015). Wormholes can, however, theoretically be stabilized by something called exotic matter. “Exotic matter…contains negative energy density and a large negative pressure.” (Redd 2015). It is theorized that a wormhole could be made stable enough to send humans through if enough exotic matter were inside the bridge. However, it is thought that sending “normal” matter, such as humans or spacecraft through a wormhole might bring about the possibility of destabilizing the wormhole and causing it to collapse.

To be technical, wormholes aren’t the real name of this theoretical phenomena. They are actually called Einstein-Rosen bridges. This is due to the fact that Einstein’s theory of relatively is used to predict the existence of wormholes. “Well, scientifically, they aren’t called wormholes
at all…Theory of Relativity is what allowed for their existence” (Denagamage 2014). A simple way of imagining the behavior of a wormhole, is to take a sheet of paper and stretch it so that it is taught at every end, then place some object on the paper to make it bend. This is essentially what objects in space are thought to do to the fabric of space-time. “…imagine that you have a sheet that is stretched taught…the gravity of objects bends space-time around them” (Denagamage 2014). However, wormholes don’t just slightly bend space-time. In the analogy of the piece of paper, the paper would be completely folded over, the top part of the paper being one area in space, and the bottom part being another area in space, and these areas can be light-years apart, refer to Figure 1. If we imagine that the two parts of the paper have gravitational fields, those fields would be acting in opposite directions, both pulling inwards to the area between the two parts of the paper. Should some object come into the area of influence of one of these gravitational forces, they would be pushed into the middle and eventually end up on the other side, so if this paper is space-time, that object would have just travelled, potentially, thousands, if not billions of light years in a very short time (Denagamage 2014).

To further understand wormholes, one must also understand the theory behind black holes and the idea of white holes. Black holes are, basically, an area in space that has so much mass concentrated in one area that even objects that are travelling at the speed of light are unable to escape from its gravitational pull. This means that it is impossible, even for light, to escape from the pull of a black hole once it reaches the black hole’s event horizon. “In particular, a black hole has something called an ‘event horizon…but you can't get back out.” (Bunn 1995). Along with black holes, there is the idea of a white hole, which is essentially the antithesis of a black hole. Where black holes pull in matter, white holes expel matter. “A white hole is a black hole running backwards in time…white holes spit them out.” (Hamilton 2001). White holes come into play when talking about a Schwarzhchild wormhole, which is comprised of a black hole at one end, a white hole at the other end, and a wormhole that connects the two, allowing for travel from one area to another, or one universe to another (Hamilton 2001). What is interesting about white holes and black holes is that at certain points in time, it would be impossible to distinguish one from the other because one will absorb the same amount of radiation as the other releases. This is interesting because it brings up the question of what happens if there is a white hole somewhere in space that isn’t connected to a black hole. A white hole that is isolated cannot absorb any Hawking radiation that is released from a black hole. Since the white hole cannot perform this task, it is thought that it would actually let off pseudo-Hawking radiation and explode (Hsu 2011). This leads to a flaw in this hypothetical phenomena, and that is that a white hole breaks the second law of thermodynamics, and therefore cannot exist as an isolated object, as well as it cannot actually ever go out of equilibrium with a black hole. This makes the idea that a wormhole being a bridge between two different universes by acting as a connection between a black hole and a white hole very unlikely.

**Einstein’s Theory of General Relativity**

Before getting into general relativity, it is helpful to understand the theory of special relativity, as that was Einstein’s first step along the way to his theory of general relativity. This principle basically states the difference in perception of time one person would observe relative to another person as they both travel at constant, but different speeds. This is the concept of time
dilation, which is that someone moving at a very fast speed, like in a vessel in space, would observe time as passing slower than someone who is on Earth. This is also where Einstein’s very famous formula, $E = mc^2$, came from. The basic explanation of this formula is that there is a direct relationship between the speed of an object and it’s mass. That is, the mass of an object will increase as its velocity increases. These two aspects of his theory were the association of space and time, and mass and energy.

After theorizing special relativity, Einstein began further work to relativize physics. “The fuller development of his goal of relativizing physics came with his general theory of relativity.” (Norton 2015). His theory of general relativity was the realization of that goal, and was also the last theory in this field that is considered non-quantum; it has caused a problem in that there hasn’t been a way to use it in tandem with quantum theory as of yet. This theory was very revolutionary because it dealt with forces, like gravity, in a way that they had never been considered before. “It treats a force by means of geometry and eventually leads to startling notions: black holes…” (Norton 2015). The main idea of this theory, in keeping with treating forces geometrically, is that the gravitational force of a celestial body literally bends the fabric of space-time around it. This differs from Newtonian theory, which states that the reason that the Earth orbits the sun is that the gravitational pull of the sun keeps the Earth from moving in a straight line and causes it to move in an elliptical pattern. According to Einstein, however, the reason the Earth moves around the sun as it does is because the sun causes the space around it to bend and warp. “In Einstein's theory, the presence of the sun disturbs--that is, curves--the very fabric of space and time…winding around the sun's worldline in spacetime.” (Norton 2015). Therefore, the Earth is actually moving in a straight line through space, that space has just been curved to make the Earth move around the sun, rather than away from it, refer to Figures 2 and 3. This theory applies to wormholes in that they also bend the fabric of space-time so that, referring back to the paper analogy, it is essentially folded one end over the other.

**Quantum Entanglement**

Moving from classical theory into quantum theory and mechanics, the theory of quantum entanglement refers to two particles that both occupy different states at the same time (Chu 2013). A simple, yet paradoxical example is that these two particles could both spin clockwise and anti-clockwise at the same time. However, it is not definite what state one of the particles is in until it is measured. Once one particle is measured, the other takes on the corresponding state, and this relationship is maintained even over billions of light years of distance (Chu 2013). This is interesting because it appears that these two particles communicate with each other at what appears to be faster than the speed of light. Wormholes have been suggested to be the reason as to why they are able to do this. “Now an MIT physicist has found that, looked at through the lens of string theory…wormhole connecting the pair.” (Chu 2013). However, since the particles seem to communicate faster than the speed of light, this conflicts with Einstein’s theory. However this also brings about a way to potentially bring quantum theory and the theory of general relativity together. To do so, the way we view space-time will need to be revised according to quantum entanglement, that is, that the universe is not molded by gravity, but by something more elementary.
This idea came from Julian Sonner, a physicist at MIT’s Laboratory for Nuclear Science and Center for Theoretical Physics. Julian was following up on an experiment that suggested a wormhole being created from two black holes that were experiencing quantum entanglement being pulled apart, and decided to transition that concept over to quarks. “…Sonner has sought to tackle this idea at the level of quarks…Once extracted, these particles are considered entangled.” (Chu 2013). Sonner created entangled quarks and then began to pull them apart, just as the black holes in the experiment he was inspired by were pulled apart. The result was the same, a wormhole was formed between the two quarks. This could explain how two particles or celestial bodies could communicate with each other so fast over great distances. This experiment also implied some contradictions to Einstein’s theory of general relativity, as previously stated. “…implying that the creation of quarks simultaneously creates a wormhole…such as that between pairs of particles strung together by tunneling wormholes.” (Chu 2013). Sonner’s findings seem to suggest that gravity, as well as the bending of space-time, may actually be a product of quantum entanglement.

**Time Travel with Wormholes**

Einstein’s theory of relativity shows that time travel is, theoretically, possible so long as someone or something is able to move at a fast enough velocity. Relativity shows that if a space vessel moves fast enough, the crew on board would only experience a few days or weeks of time, but because they were moving at some fraction of the speed of light, the universe around them would have experienced, potentially, thousands of years. "Since moving clocks run slow in relativity…external universe depending on how fast they were traveling.” (Friedman). This is considered time travel as a crew could leave Earth, travel at that constant speed for a week in a straight line, decelerate, turn around and come back to Earth, and it could be thousands of years later on Earth despite it only being a week or two for the crew, but this does not utilize a wormhole. However, if some technology were created that would allow for a wormhole to be created and stabilized to the point of traversal, the vessel could do the same maneuver as previously mentioned, but this time, it would have a connection back to Earth with the device to create a wormhole. The ship would stop above the Earth, thousands or millions of years later from when it left, but the wormhole created between the device on the vessel and the device on Earth would allow for the crew to come back to the year they left on, and allow for travel to the future (Friedman). This also shows that it is impossible to travel backwards in time to before the moment that the wormhole connection was first made between the ship and the facility on Earth, which answers the question of why nobody from the future has travelled back in time to visit us at any point in history so far.

**Wormhole Based Travel Illusion**

In August of 2015, a paper was published about an experiment that was conducted by a group of researchers based out of the Autonomous University of Barcelona. What they did is they created a small sphere that was made out of ferromagnetic and superconductive metamaterials. This sphere was separated into two layers, the surface being made of ferromagnetic material, and the second layer was made with the superconductive material. The
innermost area of the sphere had ferromagnetic material wound into a spiral shape to act as a coil to allow for current to pass through. They then submerged the sphere in liquid nitrogen. They then sent a current into the coil they made to provide a magnetic field. They monitored the magnetic field as it entered the sphere to see how it behaved. What they saw, however, was that the magnetic field seemed to not travel through the sphere at all. Instead, it merely entered at one end of the sphere, and came out at the other end without any in between. “When the magnetic field entered the sphere at one end…there was no trace of the magnetic field.” (O’Callaghan 2015). What occurred was, the magnetic field did actually travel through the sphere, however it was rendered invisible by the materials the sphere was made from. There superconductive layer and the ferromagnetic layer caused the attraction and the repulsion of the induced magnetic field to cancel out, which made it invisible to instrument detection. “…the attraction and repulsion of the magnetic field was cancelled out, making it undetectable.” (O’Callaghan 2015). What is really exciting about the results of this experiment is that, if this concept could be taken and applied to technology like MRI machines, the machines could become much more efficient. If an MRI were able to work without producing any magnetic interference, there is potential that a person could have multiple MRIs done on multiple body parts and all the images would be clear of any interference or distortion since the magnetic fields would be invisible and non-detectable by the other machines.

Conclusion

Wormholes are a very exciting and mysterious prospect. These bridges through space are a very interesting, theoretical phenomena and can be rather hard to wrap one’s head around, especially with the relatively sparse and sometimes conflicting information out there. They could potentially allow for travel across billions of light-years of distance through space in a very short amount of time. They could even be used to allow for people to travel forward in time. Even more difficult to grasp is that it is thought that wormholes could act as bridges between universes. All that is very hard to grasp and even harder to test since we don’t have an actual wormhole to test these hypotheses on and some of these hypotheses even break some of the laws of physics that have been in place for many years now.

I believe that it is time to change our view of space-time, as was mentioned in the section on quantum entanglement. The fact that classical theory and quantum theory clash so much has been holding this subject of study back, and if no way to work around these incompatible theories can be found, it would make sense to review and critique what we already know. Einstein was a very brilliant physicist and made it possible for us to have gotten to where we are today in our knowledge of physics, but that doesn’t mean he couldn’t be wrong on certain parts of one of his theories. If we can finally merge the two areas of theory, it could potentially lead to new breakthroughs in our understanding of space-time and the universe as a whole.

Since all the information on wormholes is only partly theoretical and a lot more hypothetical, it is hard to really see where the future for this area of study will lead. I would like to see it lead to the confirmed existence and utilization of wormholes and eventually even used as a means of time travel. However, there is so much unknown in this subject that I really cannot say that I can see where it is headed. I would like for the merging of classical and quantum
theory to become a reality so that we can see what new information and hypotheses come from that merging. It is a very interesting and exciting area of study and we should all look forward to where future studies and researchers will take it.
Figures

Figure 1- Space-Time Folding Like Paper (Redd 2015).

Figure 2- Newtonian Gravitational Theory (Norton 2015).

Figure 3 - Einstein's Gravitational Theory (Norton 2015)


Downhill longboarding is a competitive sport where riders suit up in protective gear at the top of a long hill, or even a mountain, and race to see who reaches the bottom first. Riders need both a high amount of skill and a comprehensive knowledge about the physics behind the sport to be successful. In order to attain the highest speed, the rider must understand how their potential energy is converted into speed and what factors, such as friction, work against them. They can then choose maneuvers and gear to counter these factors and even use these factors to their advantage. Downhill riders need both speed and stability, and by using physics, they can determine how to emphasize these components while giving them a competitive edge.

Skateboarding began in Oahu, Hawaii in 1959 when surfers needed a way to train when there were no waves. It was named sidewalk surfing and the boards were simple wood planks attached to metal roller skate wheels. In the 70s, urethane wheels were developed which made the sport safer and improved the overall ride of the skateboard. In the 90s, skateboarding gained further popularity through celebrities, such as Tony Hawk, and distance skateboarding became a means of transportation. This is where longboarding became separate from skateboarding and eventually evolved its own subcategories, one of which was downhill longboarding. This category emphasizes speed and control as the rider races down hills. It is becoming a serious sport with professional riders sponsored by companies such as ABEC 11, Longboard Girl’s Crew, and Landyachtz. There are even international races that require qualifying races such as Pike’s Peak in Colorado and the Kozakov Challenge in the Czech Republic. No longer is longboarding just a means of transportation, but a serious sport that continues to gain popularity.

Although one can learn to ride a longboard without knowing physics, understanding the physics behind the ride can only help the rider. This is especially true for downhill races where tactics, such as speed control, not only matter to the rider for competitive reasons, but are a huge safety concern when competitors are easily achieving speeds of over 30 mph. There are many simple maneuvers performed by longboarders that help them not only win the race, but do so safely. All of these maneuvers can be explained by physics.

The most basic physics component to keep in mind about downhill longboarding is that, potential energy, which is “the capacity for an object to do work based on its position” is equal to kinetic energy, which is the energy of an object in motion (ignoring all other factors). Downhill races start at the top of the hill, which is where the potential energy is at its greatest amount and the kinetic energy is equal to zero, because the rider is not moving. At the top, the potential energy is equal to the mass of the rider multiplied by the gravitational constant multiplied by the height of the hill, resulting in the following equation:

\[ PE = mgh \]

Again, when all other factors are ignored, potential and kinetic energy are conservative forces, meaning there is no net loss of energy when the energy is converted from potential to kinetic. At the bottom of the hill, when the rider has achieved maximum speed, all of the potential energy has been converted to kinetic energy. The kinetic energy is equal to the mass of the rider, times the velocity squared, all divided by two, demonstrated by the following equation:

\[ KE = \frac{mv^2}{2} \]

When looking at this from a longboarding perspective, the conversion of the potential energy into kinetic energy is what causes the rider to accelerate down the hill (see figure 1). The rider at the highest hill with the greatest mass (greatest amount of potential energy) has the greatest potential to achieve the fastest speed at the bottom of the hill (kinetic energy).
Unfortunately, in the case of longboarding, the mechanical forces are not conservative because friction causes energy to be lost as heat. Friction is defined as an opposing force of an object’s motion due to the resistance of their surfaces and can be calculated by multiplying the object’s normal force by the coefficient of either static or kinetic friction. The normal force is the force applied back to the object by the surface to which it is applying force and is equal to the product of its mass and the gravitational constant (on a flat surface). Static friction occurs when the object is not in motion; in order for the object to move, a force greater than the static frictional force must be applied. The friction caused by an object sliding along a surface is called kinetic friction.

\[
F_{\text{kinetic friction}} = \mu_k N
\]
\[
F_{\text{static friction}} = \mu_s N
\]

When observing the whole situation, the potential energy being converted into kinetic energy while energy is being lost through friction simultaneously.

Although friction takes energy away from total potential energy, it is an important component when the rider needs to slow down, or brake. There are two ways to brake: foot braking or sliding. Foot braking is when the rider drags their foot across the street to create friction and slow themselves down. This method cannot be used for downhill races because it would be highly unstable and unsafe. Sliding is the alternative method, and is essentially when the rider kicks the board out and the wheels slide perpendicular to the direction of motion (see figure 2).

The friction in this case is a little more complex. When the longboard wheels are rolling down the hill, static friction is present because, although the wheels are in motion, they maintain a point of contact with the road so they do not slip. This force does not take energy away from the system, but there is kinetic friction present at the axle, plus the wheel is flexing from the force, which dissipates some energy. This is referred to as rolling friction. In order to execute a slide, the rider must push the board out with a force greater than the static frictional force. Then, when the wheels are dragging across the street rather than rolling, it creates kinetic friction between the wheels and the street and drastically slows the rider down until they either come to a stop, or they continue riding down the hill. During a slide, the wheels are indeed dragging across the street, but they also continue to roll, so both kinetic and static friction are present. There is also the option for the rider to use sliding gloves, which have plastic pucks attached to the palms, and place their hand on the ground while they slide (see figure 3). The plastic sliding across the ground supplies additional kinetic friction, and the motion itself provides more stability for sliding at high speeds.

Another component of friction present is known as drag, which is the force opposing the forward force of an object due to air friction. The air passing over the rider’s body creates drag and slows them down. In order to increase speed, the rider assumes a tucking position, which reduces the surface area coming in contact with the air, therefore reducing the drag. If the rider wishes to use drag to slow themselves down, they stand up and put their arms wide out. This increases surface area, which increases the force of drag. Tucking is especially important because not only does it help maintain speed, but it increases the riders stability by lowering the center of gravity. It is also be used to momentarily increase speed so other riders can be passed during a race.

The most critical moments during a race are when the riders navigate turns because there is the highest chance slowing down and losing the race, or even serious injury due to crashes. The rider must understand how friction and centripetal force operate so they can respond and
maneuver their board appropriately to stay safe as well as retain the highest amount of energy to maintain their speed. According to Newton’s first law, an object in motion will stay in motion unless it is acted upon by an outside force. The longboard will continue to travel straight unless the rider forces it to turn, which they do by leaning left or right. Throughout the turn, as long as the wheels are not sliding across the ground, only static friction is present. In order for the rider to complete the turn, the frictional force has to balance the centripetal force.

Centripetal force is defined as the center-seeking force of an object in motion in a circular path. The magnitude of this force can be calculated by multiplying the mass of the object with the square of its velocity, all divided by the length of the radius of the path of motion:

\[ F_c = \frac{mv^2}{r} \]

For the frictional and centripetal forces to balance, they must be equal, giving:

\[ F_{f,s} = F_c \]

Substituting in the values for each equation:

\[ \mu_s N = \frac{mv^2}{r} \]

When the coefficient of static friction between the particular polyurethane of the wheels and the asphalt is known (\( \mu_s \)), as well as the radius of the turn and the mass of the rider and the board, the maximum velocity at which the rider can make the turn without sliding can be calculated.

\[ \mu_s Nr = m\nu^2 \]

\[ \frac{\mu_s Nr}{m} = v^2 \]

\[ \sqrt{\frac{\mu_s Nr}{m}} = v \]

If the final equation is observed, it can be seen that the variable for the radius of the turn is located in the numerator of the fraction. The value of the numerator is the main determining factor of the speed because the radius can be chosen and adjusted by the rider, while the other variables cannot be manipulated in a given situation. In order for the numerator to be large, the rider must choose the path of the turn that has the largest radius. Following the inside of the road will provide the smallest radius, so the rider’s speed will be the slowest if they choose this path. Using the outside of the turn provides a larger radius, so the rider will be able to go a little faster, but there is a third option that allows for the largest radius. If the rider begins the turn on the outside of the road, hugs the corner of the turn, then returns to the outside, the largest radius is achieved and consequently the highest speed (see figure 4). Not only does choosing the largest radius allow for the highest speed, but it makes it easier for the rider to physically complete the turn. When looking at the equation for centripetal force alone, the radius is located in the denominator of the fraction. When the denominator is large compared to the numerator, it gives a smaller value for the centripetal force, meaning there is less force pulling the rider towards the center of the turn, making it easier for them to balance and maneuver.

There are points in time during the race where the rider may exceed this speed, which would cause the wheels to slip and begin sliding across the asphalt. Sliding around a turn is comparable to a car drifting around a turn and it allows the rider to complete the turn at high speeds without crashing. There is another option where the rider initiates a straight slide just
before going into the turn, which releases energy as heat between the wheels and the ground, drastically reducing the speed of the rider. Which method the rider chooses depends on the skill of the rider and the circumstances of each particular turn, such as how sharp the turn is, and where other riders, cars, or obstacles are.

Although the maneuvers during the race are important if the rider wants to win, the setup or build of the longboard itself can have a huge influence on the performance of the rider. Special considerations are taken into account when choosing a deck, wheels, trucks, and even protective gear. Each item has certain features that can either give the rider an advantage or disadvantage when the purpose is to set up a board for downhill racing. The aim is to have a stable board that can achieve and handle high speeds.

Starting with the wheels, the size of the wheels will affect the speed of the board. The velocity of the wheel is equal to its angular velocity (represented as \( \omega \)) times the radius where the angular velocity is equal to its frequency in revolutions per second multiplied by \( 2\pi \):

\[
\begin{align*}
\nu &= \omega r \\
\omega &= 2\pi f
\end{align*}
\]

Mathematically, when the value for \( r \) (the radius of the wheel) increases, the velocity will also increase because they are directly proportional. Although they may achieve the fastest speed, they are heavier and will accelerate slower. Longboard wheels are measured in millimeters and average in the 70-80 mm range.\(^{14}\)

The hardness of the wheel is another factor to take into consideration. Wheels are labeled with a durometer, which is the measure of hardness that consists of a number followed by a letter, which is “a” for the polyurethane that the wheels are made of. The durometers of longboard wheels range from 75a to 99a, with 75a the softest and 99a a very hard wheel (see figure 5). A softer wheel provides a lot of grip, but produces more friction while the harder wheel has less grip and less friction.\(^{14}\) Although it is much easier to initiate slides with harder wheels, downhill riders will choose a softer wheel, usually in the range between 78a-83a.\(^{14}\) At higher speeds, it is more important to have good grip. It is easier to initiate a slide at higher speeds, and if the wheels are too hard, they could loose traction when the rider did not intend for them to, possibly resulting in a crash.

When considering a deck and a pair of trucks, the goal is to have a stable board with a low center of gravity. The center of gravity is the point on the object where the majority of the weight is located. In order for the object to remain upright, its center of balance cannot lean over boundary of its base (see figure 6). Taller objects have a higher center of gravity while shorter objects have a lower center of gravity.\(^8\) If the two objects have the same size base, the taller object will topple over first when it begins to lean because its center of gravity will cross over the boundary of the base first. The taller object’s stability can be greatly increased by expanding the size of the base.\(^8\) In the case of longboarding, a human is a relatively tall and thin object, having a center of gravity near the hips. It is easy for the rider to lean over the base of the longboard and lose their balance.

There are different styles of longboard decks to choose from that have different advantages and disadvantages. For a dropthrough deck, the trucks are attached through a hole on each end of the deck (see figure 7). These decks are lowest to the ground, which gives the rider the lowest center of gravity, making it the most stable type of board for high speeds.\(^{15}\) These boards are recommended for beginners because of how stable they are. There is also a top mount deck, where the trucks are attached directly to the bottom of the board (see figure 8). These boards are higher off the ground, making the rider’s center of gravity higher and decreasing the
stability of the board. These boards are capable of maintaining a higher grip at high speeds than
the drop-through deck, but the require a greater amount of skill on the part of the rider to
maneuver and remain stable.\textsuperscript{15} There are also drop-platform decks, which have the trucks
mounted to the bottom of the board, yet the board itself dips down and brings the rider closer to
the ground. These work for downhill boards, but are not preferred because they do not allow as
much grip as a top mount, plus the board usually has unique concaves meant for another style of
longboarding, known as freeriding.\textsuperscript{15}

The truck, which is what connects the wheels to the board, is another vital component of
the longboard that needs to be chosen with thought. There are options for the height, the kingpin
location, and the angle of the baseplate. There is always a trade off between the stability and
responsiveness of trucks. For the height, a higher truck raises the center of gravity, therefore
reducing stability, yet it increases the leverage over the pivot axis, which increases the ability to
turn sharply.\textsuperscript{16} When the height is lowered, the center of gravity is also lowered, which increases
stability, especially at high speeds, but decreases turning ability since there is less leverage over
the axis.\textsuperscript{16} Downhill longboards should have lower trucks to emphasize stability at high speeds,
even though some turning ability is sacrificed.

The kingpin of the truck is the central piece that holds the truck together. A standard
kingpin faces the center of the deck and provides a rigid feel and sharp turning response. These
types of trucks are commonly found on traditional skateboards.\textsuperscript{16} Reverse kingpins, which point
away from the center of the board, have a “smoother and more fluid turning feel” (see figure
8).\textsuperscript{16} There are even double kingpin trucks which have extra fluid turning. Downhill longboards
will almost always have a reverse kingpin truck because the ease of turning over the standard
kingpin is favorable, while double kingpins turn too easily and become unstable at high speeds.

The baseplate is the metal piece that attaches directly to the deck itself, and the angle of
the baseplate can be manipulated to acquire desired characteristics (see figure 8). A higher-
angled baseplate increases the angle between the axis of the truck and the pivot point. This
provides more leverage per lean, making it less stable with a more dramatic amount of turn.\textsuperscript{16} A
lower-angled baseplate provides less leverage, reducing the amount of turn but increasing
stability.\textsuperscript{16} Angles range from 35° to 50°, with the lower range of angles being favored by
downhill riders. Advanced riders can mix angles, choosing a lower angle for the rear and higher
angle for the front, which would still allow more turn, yet keep the rear more stable.\textsuperscript{16}

When a rider intends to partake in downhill longboarding, or races in particular, safety is
a huge concern because crashing at such high speeds can result in serious road rash, broken
bones, and head injuries. So while protective gear is mandatory, downhill riders choose gear that
has aerodynamic efficiency to maintain speed and assist acceleration. Wind resistance consists
of air pressure drag and skin friction.\textsuperscript{17} When a rider accelerates down a hill, they disturb the air
flow so it breaks and flows around the rider. This creates a high pressure zone in front of the
rider with a low pressure zone behind them, consequently pulling the rider backwards and
slowing them down.\textsuperscript{17} Skin friction occurs when the air directly contacts the skin and creates
friction that slows the rider down.\textsuperscript{17} In order to reduce these effects, the rider can choose
protective gear that helps streamline their profile as well as assuming positions that reduce drag,
such as the tuck.

When buying a helmet for downhill longboarding, riders can choose between a traditional
skateboarding helmet or a downhill specific helmet. Traditional helmets are bulkier, have an
open face, and the top of the helmet has a flatter shape. Downhill specific helmets have full face
protection with an angled visor and are streamlined by the top being rounded and the sides being
narrower (see figure 9). The full-faced helmets are preferred because not only do they provide additional protection, their streamlined features break the air more easily and produce less drag. These types of helmets cost hundreds of dollars, so the rider has to choose if the advantages are worth the high cost.

In order to reduce skin friction, tight-fitting clothing can be chosen over loose clothing, but this does not provide much extra protection. Racing leathers are available, which are form-fitting leather body suits that provide more protection than regular clothing but also reduce skin friction and drag because they are form-fitting (see figure 9). These suits can easily cost over a thousand dollars, but are worth it for professional racers who regularly put themselves at high risk.

The world record speed on a longboard was achieved by Mischo Erban at an undisclosed location in Colorado in 2010. He clocked in at 80.83 mph, monitored by the International Gravity Sports Association. He is known as one of the best downhill longboarders of the time, as he was the defending IGSA World Cup Series Downhill Skateboarding Champion in 2010. His gear and his board set-up were chosen to help him achieve such high speeds. He chose a GMR drop-through longboard deck, which lowered his center of gravity. His Seismic 85mm Speed Vent wheels had a large radius and come only with a durometer less than 80a, showing that a large radius and grippy wheel support high speeds well. His choice of truck was a Ronin reverse kingpin, which provides fluid turning, although the angle and height was not specified. For safety gear, he was wearing a streamlined downhill helmet with a full-faced visor and racing leathers, which reduced his drag and skin friction. He further reduced his drag by assuming a very low tucking position. All of his gear was chosen to favor stability, grip, and to reduce air resistance, which, combined with his skill and experience as a downhill competitor, helped him achieve the world record speed on a longboard (see figure 9).

Overall, downhill longboarding is a dangerous, adrenaline-pumping sport that requires a high amount of skill and nerves of steel. When the rider understands the fundamental physics behind the sport, they will be able to make better predictions about their speed and motion and be able to know what maneuvers will be appropriate for certain situations. They can calculate the maximum speed at which they can turn without having to initiate a slide, which they can then practice recognizing what that speed feels like physically so they know when they need to slide to complete a turn safely. They can calculate what their speed will be at the bottom of a hill and determine if it is possible to complete that particular run with their current skill set. It is also important that they understand the physics of their board and use this knowledge to build a set up that will give them a competitive edge.

As a longboarder myself, I have much respect for downhill riders because through personal experience, it takes a great amount of self-control and physical skill to maintain the stability of a longboard at such high speeds. When this stress is combined with the stress of other riders in a competition, it is crucial to have control over the mind as well and to be able to keep calm and focus. Downhill longboarding is gaining popularity as the technology and gear are improving to make the sport safer. I hope to see this become a major sport, possibly even an Olympic event similar to luge, and gain a higher respect from all athletes and nonathletes alike.
Figures

Figure 1. Conversion of potential to kinetic energy

Figure 2. A longboarder sliding perpendicular to his direction of motion. (http://ffden-2.phys.uaf.edu/211_fall2013.web.dir/craig_lematta/Project%20template/Slide5.htm)

Figure 3. A homemade slide glove
Figure 4. R1 (orange) shows that path on the inside of the road results in the smallest radius, giving the slowest speed. R2 (yellow) shows the path on the outside is intermediate, while R3 (pink) shows that going from the outside, hugging the corner, then returning to the outside results in the largest radius and the greatest possible speed.

Figure 5. 70mm 90a wheels. This is a harder wheel meant for sliding rather than downhill racing.

Figure 6. The object topples over when the center of balance crosses over the boundary of its base. A wider base increases stability.
Figures (continued)

Figure 7. A dropthrough deck; the truck is connected through a hole in the board.

Figure 8. This is a reverse kingpin truck, pointing away from the center of the board, connected to a topmount board. The baseplate angle is also demonstrated.

Figure 9. Mischo Erban sporting his dropthrough board, downhill helmet, and racing leathers.
(www.dailymail.co.uk)
References

Electromyostimulation, Electromyography, and the Human Condition

Trung Tran

PHY 112
Dr. Durandet
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ABSTRACT

There is a direct correlation between isometric muscle strength and electrical activity associated with such strength. This has been a major focus of study since the early 1950s. Although a consensus has not been confounded, the utilization of electromyography as a metric unit for analyzing the various possibilities for determining force is both challenging and competitively fascinating. It has been challenging due to the mere complexity and inherent variability of physiological and biological signals and has been fascinating to make such quantitative and qualitative measures of strength while ostensibly performing a gesture through electrodes on an individual. This study utilizes methodologies in order to collaborate various calibration processes, which include but are not limited to, electrode arrays and PCA, signal processing using multivariate statistical techniques.

Electromyostimulation, as well as electromyography, have gained incredible insights due to the now advanced use of EMS gadgets, which are sold to the public. Although broadly ineffective, this has increased the awareness of the product and this has led to many advances in research on the subject. EMS has the potential to offer many things that may lead to the benefit to any individual seeking to improve their muscle mass and physical capacities.

INTRODUCTION

It is estimated that the average human will walk over 100,000 miles in a lifetime (1). As humans, we spend a large portion of our time moving, from our day-to-day actions of tending to your home and self, to the hard work and consistency that goes into achieving a physical goal such as a marathon. Whether voluntary or passive, our lives are made up of the movements we make. (2)

Understanding how the muscle works in normal movements, as well as during exercise is important for understanding how the human body functions as a unit. (2) Movement is far more complex than our bodies have led us to believe: it is made up of muscle force, joint position, and electrical signal all of which must work in perfect unison. As such, this makes it exceedingly difficult to recreate. This difficulty in maintaining accurate measurement has lead to the use of biomechanics and its reliance on calculations and theoretical models. (2)

Biomechanics is a field of science that has been utilized to understand human movement and the muscular forces that affect it. It is a combined application of understanding biological systems, mathematics, the physics of mechanics, and electromyography that make up biomechanics. Electromyography is a particularly important piece of the puzzle, as it is tangible and not a mathematical assumption. It is this that puzzle piece that is the method of recording electrical activity within a muscle, as well as the physiological processes that occur during muscle contractions. Unfortunately, due to the aforementioned variability, there is a general assumption regarding electromyography and muscle forces, which has led to a number of suppositions in regards to muscle physiology. (2)
Theory of Electromyography and Electromyostimulation

Electromyography has been utilized as an analytic method of assessing muscle health and the motor neurons that control them. It then records the electrical expression of the neurons within the contracting muscle, where the electrical expression is representative of the “ionic flow across the membrane of muscle fibers” (2). This ionic flow will then spread to the adjacent tissues where electrodes attached to the proximal skin will sense them. This is meant to discern whether the electrical signals produced by the skeletal muscles are functioning properly. This ability to detect electrical signals has allowed it to be utilized in the medical field under a large slew of disciplines from psychology to kinesiology to biomedical engineering and research. (10) Unfortunately due to the complicated nature of the signal there are a number of complications that include the properties of the muscles being analyzed, as well as the machines utilized to observe it. (2)

Theoretically electromyostimulation is meant to innervate motor units to achieve maximum contraction, which will allow for a number of beneficial effects. In fact, electromyostimulation has gained a substantial amount of recognition over the last several years. It has been successfully used in a number of fields in both athletic and non athletic peoples spanning from training and aesthetics to the rehabilitation of injuries. Referencing its use in training and athletics it has been found effective in conjunction with sports training, and as a tool in staving off injury. Its use in rehabilitation has also been prosperous, having been found to act preemptively in individuals with mobility restraints such as paralysis, as well as to test cognitive and muscular function. (10)

Electromyostimulation and Muscle Strength

Muscle strength is defined as the muscles ability to produce tension. This begins in the nerve and then transfers via pathways through the central nervous system. After it hits the central nervous system it recruits motor neurons, which then stimulate muscle fibers, which then produces the action and provides an example of muscle strength. (3)

The motor unit is the functional muscle of the muscle. It is the primary motor unit upon which all others are innervated from. Within the motor unit are muscle fibers, which can number in the thousands. Muscles that are meant for specific and detail-oriented tasks, but require little power are typically lower in number of muscle fibers, than those that require a lot of power and strength. The contraction of the muscle fibers are powered by motor neurons, which will in turn trigger the finer workings of the muscle, causing the entirety of the muscle to contract. (2) Different sorts of muscles, require different types of mechanisms, meaning that smaller muscle, such as those in the foot or hand require are controlled in a different way than larger muscles, such as the ones in legs or arms: they require different schemes of recruitment and activation. (2) One way that this is seen is in muscle fiber types I and II, which are activated in different ratios and quantities depending on the task.

More than just more than just the type of fibers being activated though affects Muscle Strength. The elastic and contractile structure can also play a part in the
determination. The contractile and elastic portion of the muscle has an effect on the
muscle length, which is the primary determinant for general muscle strength. When the
muscle stretches to its greatest length, that is when it is able to generate maximum
tension, allowing force to produced in great quantities. (2)

Electromyostimulation can affect muscle strength. It can be seen that
electromyostimulation can cause an overstimulation of mid-region muscles, which can
then cause an increase in muscle size. (5) It was seen in the study that acute bout of
electromyostimulation was able to cause response at the molecular level. This response
has effects that initiated hypertrophy in the quadriceps of an individual with no physical
ailments. Ultrasonography was able to provide imaging of the individual muscle within
the quadriceps. Due to this it was seen that the three vastii muscles increased, while those
in the rectus femoris did not. These changes were viewed over the course of an eight-
week protocol of electromyostimulation. This resulted in seeing that under the correct
circumstances both neural and muscular adaptation, in strength and otherwise, could be
seen. It also indicated that the hypertrophy might have been dependent on selectivity. (5)

Electromyostimulation and the Normal Body

Electromyostimulation has often been utilized as a methodology of strength
training in untrained individuals. Reports show a positive increase in maximal voluntary
contraction (MVC), or the maximum force that an individual can produce in an isometric
exercise, following sessions in which electromyostimulation has been used in large
muscle groups such as the quadriceps femoris muscle. Despite this evident increase in
muscle strength it is not entirely understood where the improvement has evolved. (5)

In addition to strength training and changes on a muscular platform,
electromyostimulation has been studied as a method of effecting neural factors. This is
especially evident in programs and studies where the electromyostimulation lasts for less
than four weeks. (5) In a study that tested the efficacy of plantar flexor MVC after
electromyostimulation and electromyography activity over the course of four weeks, it
was found that there was a positive correlation between the increase of muscle activation
and the stimulation. In terms of neural effectors, it was found that there was once more a
positive correlation between the stimulation of the quadriceps and activation of neural
factors and brain regions. There was also a neurophysiological phenomenon viewed
where an increase of strength in the untrained limb was viewed after an increase in
strength was had in the opposite limb. This neurophysiological phenomenon made it
evident that electromyostimulation affected not only the skeletal muscular system, but
also the neural system. (5)

Electromyostimulation has also been seen in media as a terms of aesthetic
enhancement. Recently, there has been a widespread use of bands with embedded
electrodes that are meant to increase the size and appearance of abdominal muscles, as
well as reduce fat. (3) This can also be seen in “whole body” versions, called whole body
electromyostimulation, in order to pursue these same visual goals of reducing fat and
increasing muscle mass. (4) In some studies, results from these sorts of studies on body
composition have proven positive, however there is conflicting data. In some studies it
remains unclear whether or not there is a significant effect of energy expenditure after electromyostimulation is applied.

In one recent study, 19 men who were moderately trained were assigned a low intensity resistance exercise regimen, one in conjunction with electromyostimulation (85Hz) and one without. After each week on the regimen both parties were asked to perform tasks as a means of measuring progress and at the conclusion of the study indirect calorimetry was recorded. It was found in these cases that there was a moderate increase in calories burned during exercise when performed in conjunction with the whole body electromyostimulation. (5) Unfortunately, it seems as though these results could be clouded due to the inaccuracy of indirect calorimetry in above steady state conditions. (4) Regardless of this possible increase in energy expenditure it remains to be seen whether or not this is a viable option for the general public, in terms of weight loss or daily functions. (4)

Electromyostimulation in Abnormal Populations

Electromyostimulation has also undergone a number of studies in populations of people who are not able-bodied adults: the elderly and postmenopausal women have also been researched. In one study that was conducted the effect of whole body electromyostimulation was recorded over the course of a full year. The goal was to see whether appendicular muscle mass and abdominal fat mass in those who were lean, but abdominally obese was reduced in a significant way by whole body electromyostimulation, particularly in those who are inhibited from exercising in traditional way. (6)

46 lean, sedentary, elderly women (70+) of abdominal obesity were selected for the aforementioned study. These women were then categorized into a control group and a group who would be subjected to whole body electromyostimulation. Both the control and experimental group would undergo a certain number of active minutes per day. In the experimental group an 85-hertz impulse of whole body electromyostimulation was incorporated in conjunction with the active minutes. Dual energy x-ray absorptiometry (DEXA) was completed to test the body composition of the limbs and abdomen, while strength of leg extensors was measured by isometrically via force plates. (6) After the 12-month period of experimentation was complete it was found that in both muscle strength and muscle mass, those positive results were seen in favor of the whole body electromyostimulation group, as opposed to the control group. (6)

In a secondary study, 30 postmenopausal women of a moderately younger age, who had previously undergone three years of training in a prior study were given a similar exercise regimen as the previously mentioned study. These women were asked to complete 20 minutes sessions of whole body electromyostimulation in the experimental group. The study was maintained for 14 weeks, in which positive results were seen in both the control and experimental group. (7) A definite decrease in weight was seen in individuals who completed the exercise regimen with or without the whole body electromyostimulation, and there was no statistical significance seen between the two. In
contrast when the skinfold and waist and hip circumferences were retested at the end of
the study evident positive strides in reduction had been made in the experimental group,
while there was no significant alteration found from the basal values in the control group.
(7) In terms of strength gained throughout the program, there was a minimally larger
increase in those who underwent the whole body electromyostimulation, as opposed to
those who were in the control group, but it was difficult to say whether it was of
statistical significance or not. (7) It is likely that the lack of statistical significance in
strength and weight could be contributed to the shorter nature of this study. The
approximate two-month time frame did not allow for larger contributions to be had for
either of these groups. From both of these studies it is evident that there is an advantage
in the body composition and health of elderly individuals who undergo a moderate
exercise regimen in conjunction with whole body electrostimulation. There were noted
improvements in the utilization of electromyostimulation in a number of regards, and it is
worth further studies to determine just how effective it is.

Electromyostimulation: Training and Rehabilitation

Participants followed the treatments on a 15-session course, every weekday from
Monday to Friday, for three consecutive weeks. The EMS workouts and the isokinetic
workouts were both founded using different protocols. The protocol enabling the EMS
workouts included a cathode electrode, which was placed 5 centimeters distally above the
patella apex, while the anode was situated above the femoral nerve trunk proximally. As
a result, this ignited a halbwellenstrom wave that was correlated with a frequency of
fifty hertz. Although the galvanic setting was set to zero, the electrical stimulation pattern
was set to generate a maximum tetanic contraction within the participant’s pain threshold.
The faradic setting was determined based on the tolerance of the patient. The treatment
consisted of 10-second contractions followed by a 50 second rest time at a ratio of 10
repetitions per treatment.

The isokinetic workouts had a setting of zero and were carried out with the knee
at 45 degrees of flexion, as the subjects reported this as the most comfortable and
convenient position for treatment tolerance. In these workouts, there was a single
repetition completed for every single second of tetanic contraction. Throughout four
consecutive Fridays, girth and power measurements were reported accordingly, and on
the final three Fridays, the same measurements were taken at a minimum of 2 hours post-
workout session. Notable, the workout times differed for each day; there was also one
subject who had only been treated for 14 sessions.

It is worth noting that with these EMS workouts, treatment sessions that were
conducted individually produced the maximum number of faradic current that could be
physically tolerated rose substantially. There was however, a general decrease in the
number of the faradic current that could be tolerated by a participant was observed
following maximum tolerance treatments, due to muscle soreness. The isokinetic
participants observed minimal muscle soreness, even after following treatment sessions.
The size of the measurements of girth proved to have no direct relation to the power and strength of a group of muscle, as these measurements lacked to provide a definite pattern, neither increasing nor decreasing.

The study showed direct correlations in several other aspects. An excess of subcutaneous fat in an athlete would decrease the girth measurement with exercise, while increasing with exercise in an athlete with little fat as the myofibril cross-sectional area increases proportionally. Similarly, if the concentration of subcutaneous fat is reduced at the same rate the cross-sectional area of the myofibrils increases, the measurements of girth will remain unaltered with exercise.

CONCLUSIONS

Recent research on electrostimulation and electromyography have provided a better understanding and more effective ways of improving one’s muscle strength and performance, not only in athletes such as basketball players, but also in average healthy individuals. Current findings suggest that using electrostimulation along with various workout training has helped athlete to improve their maximal performance. In order to stimulate the fast-switch fibers, athletes are required to perform intense plyometric exercises, which could extensive tension on the joints and nervous system. Therefore, EMS has shown promising results that would act as a very beneficial supplement resource to help athletes reach that same goal without the risks associated with maximum training in order to achieve similar results. Throughout this study, the relationship between the signals observed in the electromyography and the muscle strength via biomechanical models. This study introduced a common, but effective, approach to adjust to the utilization of the intensity of the EMG, which were essentially derived from the Hill model.

The objective of these related studies were to diagnose the effect of a month long electromyostimulation training program on the knee extensor strength and the vertical jump performance of 10 basketball players. The electromyostimulation sessions were carried out in intervals 3 times weekly, with each session being comprised of 48 contractions individually. Testing for results and measurements were conducted prior and post training programs involving electromyostimulation, as well as after the four weeks of normal basketball training. Isokinetic strength had seen a dramatic increase by week four, p<0.05, at both high concentric and eccentric velocities. Contrarily, this was not the case for lower concentric velocities. Isometric strength also observed an electromyostimulation training increase at the two angles adjacent to the training angle. Although squat jump also showed an increase by 14% by the fourth week, the counter movement jump remained unchanged. Isokinetic, as well as isometric and squat jump performance were all maintained through week 8, however the performance of the countermovement jump showed a significant increase of 17%. The application of electromyography to measure the force necessary requires different parameters of calibration processes for each individual. This standardization was obtained by controlling enough of the variables that directly affected the relationship of the force and electromyography, or the EMG. This study therefore provides conclusive evidence that
there exists a direct correlation between isometric muscle strength and electrical activity associated with such strength, a vital factor for any athlete.

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CITED REFERENCES


The Electrophysiology of Visual Circuits in Drophila

By Kelsey Wallace

November 19, 2015

PHY 112: General Physics II

Dr. Casey Durandet
The following work is to demonstrate the historical and technical applications that have been used to further the understanding the most complex organ: the brain. To explain the processes used to experiment in electrophysiology the building blocks of the vast network that commands the brain must be understood. The basic parts of the brain as well as what entities guide the navigation of the electrical path are utilized to expand upon deeper concepts. When the properties of neurons are understood the physics behind the brain can be explored as well as the physical techniques that are used to discern what electrical events drive the brain to perform higher executive functions.

The human brain is a vast network of parts, subparts, connections and neurons that construct one of the most confound structures. Within the past decade, the collaborations between biologists, neuroscientists, physicists and engineers has allowed for the structure of the brain to continue to be decoded.

The science of the brain is a construct of biology, psychology, physics, and neuroscience coming together as a network. The brain is created of three distinct parts known as the cerebrum, cerebellum and the brain stem\(^1\). The cerebrum is constructed of two hemispheres that network information from one side of the brain to another via the corpus callosum\(^1,2\). The brainstem is considered the primitive brain since autonomic processing is controlled at these regions such as the pons, medulla and the midbrain\(^1\). Without this region of the brain, all autonomic processing would be lost and life would not be sustained without these connections. The final portion of the brain is the cerebellum, which controls balance and other functions such as emotion and language\(^1\).

The smaller aspects that make up the connections of the brain are known as neurons. Neurons, although vast in number only make up about 10% of brain cells and most reside in the cerebral cortex\(^1\). The most important concept of these structures is that they can have over thousands of different connections\(^3\). The structure of the neuron allows for an action potential to propagate to other neurons within the brain. Action potentials are electrical signals conducted in the neuron that allows for muscle movement, learning and other integral processes that allow for humans to function\(^5\). The parts of the neuron allow for it to conduct its job in a more economical fashion. The following constitute as the main parts of the neuron: dendrite, cell body, axon, and the myelin sheath\(^4\). Axons are the portion of the neuron that transmit information to other neurons throughout the brain and are also viewed as a conducting fiber\(^3\). The process of conduction is a process that is seen in biology, engineering and physics. Conductors allow electric current to flow freely and allow little resistance to charge flowing through them\(^6\). By this process the axon is able to transmit information far and wide to other neurons. The axon is insulated by myelin, which allows for the signal of an action potential that is being conducted to travel at a faster transmission that it would without the myelin\(^3\). Myelin is an example of an electrical insulator, which has a high resistance to the flow of current\(^6\). The myelin sheath is an important component to the neuron structure. When the myelin sheath is dysfunctional it can cause many neurological disorders such as multiple schelorsis\(^7\). The cell body is the powerhouse that conducts all signals through the neuron itself and is in charge of housing the nucleus\(^3\). All of these components setup the structure of a cell that can function both chemically and electrically.

The physical structure of a neuron allows for both chemical and electrical processes to occur in synchronicity. The study of physics is used to understand the
concepts of neurons and their electrical properties such as voltage, electric charge, voltage gradient and nerve impulses. Voltage is the measure of potential energy between separated electrical charges of opposite signs and in the case of potential difference the higher the voltage the greater difference of charge. The most important property that is discussed is electrical charge of ions in the membrane. In order to have Action Potentials occur in the neural system, small molecular structures known as ions create a voltage gradient within the neural membrane. The ions that are involved in this change of gradient are sodium, potassium, chloride and calcium, which have distinct charges of either positive or negative charges. Ions flow between membranes and allow for the change of a more positive or more negative membrane. For example, when an influx of sodium (a positive charged ion) allows for a decrease in electrical potential causing a depolarization of the cell itself known as an excitatory postsynaptic potential. On the other side of the spectrum negative charged ions both chloride and potassium cause an increase in the resting electrical potential causing hyperpolarization known as an inhibitory postsynaptic potential. Both of these electrical changes are portions that occur during an action potential and allow for a neuronal communication between multiple neurons. The signals that are conveyed through the neural cells are converted from electrical to chemical to electrical, which causes cells to fire, become inhibited or not allow the cell to fire. This process allows for the network of neurons to communicate across the brain and allow for learning and memory to occur. The electrical properties of the brain allow for scientists to further understand the circuits of the brain.

The brain is essentially a complex circuit board that allows for functionality. In the sense of physics a circuit is a close system that allows electricity to flow through it. The connections that occur via neurons are viewed as a large circuit that allows for information to be conveyed across the vast networks. The electrical signals that travel across the system of the brain form special circuits that control bodily functions like learning, memory, and even sleep. Although neurons are the fundamental units that must work together to connect different regions of the brain they cannot do so by isolation. The circuit of the brain allows neurons to be connected and send information from one neuron to the next and these connections cover multiple different regions that can translate information from different stimuli.

The importance of circuitry has become well known in the media as specialists of all backgrounds have come together to analyze the circuit of the brain through the Human Connectome Project. This project is mapping the human brain by specific pathways, circuits and functions of neural connectivity. By conducting this research, scientists are attempting to further understand the precise connection of specific neurons as well as what occurs in the brain with different disorders of the brain. Specific studies have been conducted to understand what occurs in the circuits of the brain with individuals that suffer autism spectrum disorder. The findings have concluded that individuals with autism have reduced connectivity within the circuits of the brain that promote social behavior.

Models of the brain are utilized in the scientific community by using insect models to construct circuits that are less complex. Other studies are attempting to mimic the brain by producing a model of its connections by building a circuit that can stimulate complex functions such as learning and storing memories. This device uses artificial electronics to mimic the neuronal connections and uses a voltage driven circuit to...
exchange signals in the network of connections\textsuperscript{17}. The importance behind this study is to understand plasticity in the brain in which certain stimuli cause weakening or strength in certain synapses\textsuperscript{17}. The study led by Shriram Ramanathan uses units that act like synapses by using a transistor whose current depends on the resistivity of the channel\textsuperscript{17}. Ohms Law determined that voltage is equal to current times resistance in which resistance is voltage over electric current\textsuperscript{18}. By understanding Ohm’s Law this study was able to show that electronic devices can learn that two stimuli are linked and store memories of the stimuli received\textsuperscript{17}. The issue of models that can replicate the structure of the brain allow for scientists to further understand how humans tick.

The brain is essential an electric component, which science can use to its advantage when attempting to study it. The study of electrophysiology is one such way that the brain can be study. Electrophysiology is the study of electrical properties in tissues of the body\textsuperscript{19}. Voltage and electric current can be measured within small increments of tissue such as a single neuron/ion channel as well as the heart\textsuperscript{19}. With the understanding of physics, neuroscientists can utilize the laws of current and voltage to further decode what is occurring certain stimuli are introduced that determine a specific behavior or when the brain is simply at rest. The study of electrophysiology has spanned across many decades and centuries to contribute to what scientists are able to accomplish today. Science has had to evolve both its techniques and instruments to obtain the findings of the brain that are now known.

The history behind electrophysiology dates back to the 1660’s when Jan Swammerdam was able to trigger a muscle contraction in a frog leg\textsuperscript{20}. Although Swammerdam was able to monitor the movements he was not able to further his understanding as to what caused such a signal to occur within the tissue\textsuperscript{20}. The individual that was able to connect the dots was famous scientist Isaac Newton. Isaac Newton was able to identify that a nerve signal was of an electrical nature but it was not until Luigi Galvani that electrical excitation was determined to be the cause of the nerve movement\textsuperscript{20}. With such electrifying findings this theory was expanded upon and given new life.

To further understand what is occurring within the a muscle of an animal, scientists such as Hermann von Helmholtz was able to determine the speed of a nerve impulse by measuring when a contraction occurred in the muscle and the delay of the electrical stimulus to the muscle\textsuperscript{20}. In order to measure such small impulses, Helmholtz had to record the velocity of the action potential, which was determined to be 25-40 m/s\textsuperscript{20}. This measurement could not have been correct since electric current travels at a higher speed, with this issue came a solution from the mind of Julius Bernstein\textsuperscript{20}. The differential rheotome was a structure that could record fast electrical propagations and show the first recordings of a nerve cell at rest and when an action potential was occurring in an excitatory nerve cell\textsuperscript{20, 21}. Many other theories were produced through this time frame that explained what was occurring in a nerve cell. The theories that further the study of electrophysiology were that of ions and their positive and negative charges and how they were responsible for certain phases that occur within an action potential\textsuperscript{20}. The advancements made upon the theories allowed for smaller molecules to be tested rather than an entire nerve.

John Z. Young was the first to introduce a neuron that could be tested electrically by using a squid axon and recording the impedance by the use of electrodes\textsuperscript{20}. The
The concept of impedance is another form of Ohms Law where \( I = V/Z \) and \( I = V/R \), where \( Z \) is the impedance which is equal to \( R \) the resistance. Throughout the discipline of electrophysiology electrodes are commonly used and are made of a metal that can conduct electricity with one receiving current from electrons (negatively charged particles) and the other releasing electrons. The development of specific electrodes such as intracellular electrodes which can pierce through the membrane of a neuron, allowed for more precise electrical readings of action potentials. Alan Hodgkin and Andrew Huxley were the scientists that created intracellular electrodes that could record a precise action potential and an understanding of the resting voltage potential of a neuron.

Techniques to further the study of electrophysiology have been used to determine what is occurring within certain sections of the brain or a specific ion channel. The following are techniques that are used today within the field: Single channel patch clamping, whole cell clamping, as well as current and voltage clamping. The single channel patch clamp technique records from an electrode that is placed into a single ion channel where ions are allowed to pass through the restrictive membrane. By this finding there was proof that ion channels were responsible for neuronal excitability, which can be seen when the channels open and close to allow negative and positive charged particles to flow through. Whole cell clamping allows for a sum of the currents flowing through all ion channels in the cell to be measured which can give an in depth analysis of what one neuron is experiencing.

The following techniques were used in research revolving around a fruit fly and its ventral nerve cord and occipital lobe. Current clamp is used to measure the voltage difference across a membrane while positive or negative current is continuously being injected into the cell. What can be looked at by this technique is the basics of Ohm’s Law \( V=IR \) which is used to determine the relationship between current and the input resistance. The application of a specific drug can be used to further understand the properties of the drug and if there is a change in the voltage response when the applied current is initiated usually resulting in a different ionic conductance. The final technique is voltage clamp recordings that inject a certain amount of current into a neuronal cell and to attempt to keep a constant voltage, by doing so a direct measure of ionic current can be calculated as well as the conductance of the cell. Each study can be utilized to understand the brain and its electrical and chemical nature.

As in many science forms, neuroscience and electrophysiology use computational models and animal models to allow for a more basic set up rather than the original such as the complex human brain. Simple models such as the drosophila (fruit fly) can display mechanistic insight into how circuits within a more complex system behave and why they behave this way. Neural circuitry may seem like an impossible endeavor but it would help to reveal which neural circuits are fundamental to a behavior and which are specializations. Fruit flies, though seem insignificant to nature can be used in research to crack difficult questions within the studies of genetics, physics, and chemistry. The use of fruit flies in research has been an ongoing venture for the past one hundred years. With a small organism the genes and mutations that occur within it are easy to identify which helps with understanding how different types of mutations affect specific genes. The important application that has been branched between flies and humans is the mutation in Notch, which is the gene that causes leukemia in humans. The Notch gene is a mutation that causes malformation in the wings of a fly and has been bridged to
human genetics. To understand this genetic mutation is a smaller form can allow for further understanding in how this mutation occurs in humans as well as what can be done to alter the mutation.

Neural circuitry can be used to map the relationship between neural activity and the behavior that occurs. Currently scientists are working to circuit-crack the brain of the fly. Five things must occur in order to solve a neural circuit: detect a specific behavior and the mechanism of the circuit, detect the neurons that are involved in causing the behavior, determine what causes the activity in the neurons and how the neuronal signals are utilized through the system, discover the mechanisms that cause the neural transformations and why are these transformations used in producing the desired behavior. The fly is used for circuit cracking based on its limited number of neurons and well a specific neuron that can be identified in each fly. The fly has cognitive abilities that are behavioral in nature and can use an algorithm to understand the neural circuits. For example, in a recent study flies were shown to decimate between dangerous objects like a predator and an object that would make a good landing site, the results calculated by the algorithm computed that there were opposing circuits that allowed for the fly to illicit a specific reaction. From this research as well as the understanding of the human brain, behaviors do not always have a single circuit. Multiple circuits that act in a parallel fashion convey most behaviors. A parallel circuit is a circuit that has two or more paths for electricity to flow and in the case of a specific behavior one path can be disturbed or not function which will not disrupt the behavior since it the stimulus still have a direct path throughout the circuit. The difficulty of identifying neurons is a set back when using the fly model but there are methods that are used that are attempting to identify specific neurons. To assign function to neurons studies have attempted to inactive or stimulate neurons to reveal circuit function and whether the neurons being tested are necessary and sufficient for the behavior to occur. If a neuron is necessary for the behavior to occur it must be active in order for the behavior to be carried out but if the neuron is sufficient for the behavior to occur only that neuron alone has to be activated to produce the behavior. Once this has been determined the question to what circuit is involved can be addressed. Just as algorithms can be used to determine how a behavior occurs mathematical models have been used to determine the circuitry. Although these models have not been successful within the fly model other models have been used to show motor patterns that occur in a basic circuit. The difficulty of cracking any circuit basic or complex is that neurons can respond to some stimuli while others don’t which causes a basic connection to become a complex process of multiple different neurons that provide the stimuli to have the behavior occur.

A class focused primarily of electrophysiology allowed for the following research to occur. To record from the ventral nerve cord of a sarcophaga bullata (fruit fly) one must secure the fly so that its abdominal area can be utilized. To record neurons in the ventral nerve cord of the fly certain equipment is used to collect the data that is needed. The electrodes used for this experiment were made of silver wire, which is plated with a sodium chloride solution so that the electrode can be oxidized when connected to the battery. The electrode must be oxidized in order to have a neutral substance so that the voltage and current that is run through the wire is grounded. To record neural activity a system known as a Picoscope is used to record the neural activity without current being transferred through the system while the AM system is used to record the current from
the ventral nerve cord. To access the ventral nerve cord a scalpel is used to cut threw the abdomen either from the ventral or dorsal side. This area is very delicate since the esophagus of the fly and the ventral nerve cord are parallel to one another and they both look similar under the microscope. If the esophagus is cut this can cause death in the specimen. Once the ventral nerve cord is exposed the reference electrode is placed in the fly’s abdomen while the measuring electrode is placed on the VNC. The Picoscope system picks up electrical activity in the VNC and can later be analyzed in the program known as MatLab. To determine whether the electrodes are placed within the VNC correctly the volume on the speakers is used to resolve if background noise at 60Hz is the cause of the readings or if it is truly neural activity. Matlab can be used to perform spike sorting, which can measure the electrical activity from distinct neurons within a system as well as detect which neurons are producing which frequencies. By measuring the electrical activity of the VNC it can be used to further understand a complex structure such as the spinal cord. Electrophysiology is now able to use circuits of nerves to further understand how the spinal cord works and how to provide further treatment to those with spinal cord injuries. For example, researchers have succeeded in using electrical stimulation to help paralyzed individuals flex their toes, ankles and knees. Electrical stimulation sends pulse of current down the spine that mimics the neural signals from the brain to initiate movement.

Other experiments can record neuronal activity through other regions of the brain. In the sarcophaga bullata (blow fly) the occipital lobe can be penetrated in order to view specific neuron activity. Multiple electrodes are used in order to record any activity. Silver wire electrodes are used as well as a glass electrode that is prepped in a micropipette puller. This puller creates a sharp glass electrode that can puncture through the optic lobe of the fly. The Picoscope system is used to record any neuron activity when the electrodes are placed properly. The speaker system can detect the correct activity as it sounds like a strange buzzing fading in and out. This sound is described in electrophysiology as a neuronal cell dying since its electrical impulses are slowly fading away.

The brain is essential a large complex circuit that scientists have been attempting to further understand. As someone who has been involved in course work that is determined to explore the convoluted functions of the brain, I truly have been given a unique insight into how science and technology have advanced its understanding within this matter. In order to participate and conduct research in electrophysiology and record neuronal function from the central nerve cord as well as the occipital lobe of the fly I had to be taught multiple disciplines that could explain the brain electrically, physically and molecularly. The data that was received from the equipment that was used allowed for students to experience what action potentials look like as well as what they sound like. So often a construct that is taught like an action potential is very similar to qualities of electrical phenomena in which it can be difficult to visually show the properties since both are at a level where we must believe the prospects of the phenomena are occurring without it being seen.

I believe that the use of models within experiments have allowed for the understanding of complex systems. Animal models have been used for decades in order to explain why certain behaviors occur in the brain. Most often flies and rats are used since they are a less complex organism and can be manipulated to view cellular
mutations. Along with animal models being a more simplistic model they also allow for experiments that could never be conducted on humans. In such cases animal models can still fall victim to ethical dilemma. Ethics has been a long-standing guideline in any scientific community when it must engage in the use of animal models as well as human models. I do believe that this is why computational models are becoming more of a necessity since such systems do not fall onto ethical guidelines.

As the world continues to unlock the secrets of the brain I believe that the findings will prove useful in understanding different mental disorders, the concept of pain and consciousness, as well as repairing brain damage and spinal cord injuries. Mental disorders could become more controlled with different therapies that may target the underlying problem rather than the symptoms. To understand the process of what is consciousness and how one is human may allow a more uniformed branch between religion and science to occur. The concept of pain, brain damage and spinal cord injuries could be finalized and used to repair damage within the central nervous system as well as the peripheral nervous system. With the collaboration of physicists, chemists, biologists, and engineers the future of neuroscience is going proliferate throughout the technological age.
Image I. The parts of the brain as well as the limbic system and basal ganglia

Image II. A structure of the neuron
Image I. Electrical Impulse within a Neuron
Image IV. White matter fiber architecture from the Human Connectome Project\textsuperscript{14}
Image 5. Recording of a voltage channel producing an action potential in a giant squid axon.
Image 6. Image of a drophila neuron connecting across the antennal lobe

Image 7. Spike sorting from the recording on neurons in the Ventral Nerve Cord of a Drosophila
Image 8. Fly Optic Lobe and the Circuit Connections that have been determined through electrophysiology.
Citations

18. Resistance and Resistivity. Retrieved November 16, 2015, from hyperphysics.phy-astr.gsu.edu/hbase/electric/resis.html


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Once More, From the Beginning: Star Edition

Chelsey Wilburn

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Dr. William Sherry
Abstract

Stars are a very complex body. They start the same way, complex. A cloud full of gas and dust, called a Nebula, starts to collapse. Once this happens, atoms start to attract each other and their mass starts to accumulate. That mass grows and becomes a protostar. This protostar will evolve to be in the Main Sequence. From there it carries on through a common cycle.

Important Information

Nebulae

Nebulae have been an accepted concept for quite a while. They concept was observed in 150 AD by Ptolemy on several occasions. He was found, on many occasions, nebulous areas near stars that were nebulous. But the areas themselves did not show the signs of being actual stars.

Later, the Orion Nebula was discovered by two different astronomers within 8 years of each other. Considering this was in the 15th century that is incredibly impressive. These two men were Nicolas-Claude Fabri de Peiresc I'm 1610, and Johann Baptist Cysat in 1618. Both men discovered the Orion Nebula by use of telescopes, no ordinary feat. Unfortunately, the first detailed study of the Orion Nebula was by a man named Christian Huygens. This wasn't performed until 1659, forty-one years after Cysat. He believed that he was the first to discover the Orion Nebula which is untrue.

After Cysat’s work though, more and more people started finding evidence and instances where other nebulae were. Finally, in 1715, a list of nebulae was published to the public eye. This list included six nebulae. After this initial publication, many other astronomers followed suit and the list started growing more and more. This gave a pathway towards actually understanding what nebulae did.

There are several different kinds of nebulae. There are emission nebulae, planetary nebulae, reflection nebulae, dark nebulae, supernova remnants, and diffuse nebulae. Each type pf nebulae is caused by something different. For instance, planetary nebulae are caused by stars, a lot like our Sun, which did not supernova and just fluffed off its outermost layers into space. This leaves behind a white dwarf and a planetary nebula. Another example is diffuse nebulae. These are the nebulae where stars are born. Diffuse nebulae are composed of left over material from population two and population three stars. These materials are left over from when these stars have exploded. There is also material from the Big Bang. These materials that have been left over include hydrogen gas, helium gas, and other heavier gases like lithium, oxygen, and sulfur. These gases are what combine with dust to make stars. Figure 2 is a great example of a diffuse nebula which was producing stars at that time.

Different Kinds of Population Stars

Populations of stars are the classifications of stars that exist in regards to what they are made of. In the previous paragraph, it was mentioned that the remnants from exploded population two and population three stars. This will give some insight to what that means. There are population one, population two, and now there are population three stars as well.
Population one stars are classified as metal rich stars. They are the heaviest and are composed of heavily recycled material. Population two stars are considered metal poor. They are composed of much lighter material. Finally, population three stars are a new kind of classification. They are composed of material that has not yet been recycled from other stars. These stars are composed of material directly from the Big Bang. Population three stars make the metals which are seen to make up population two stars.

All of this information means that diffuse nebulae are composed of all almost brand new materials, and very light materials. They are composed of almost no metals, which is very good for star making.

The H-R Diagram

A very important way of organizing different stars using multiple different categories is by using an H-R Diagram. H-R stands for Hertzsprung-Russell. These two men both created a different diagram. Separate they were decent, but together they had an incredible diagram that covered almost all categories a star could have at a quick glance and organized them in a slightly and easy to follow manner. Follow along with Figure 3 to see what a practical application looks like.

H-R Diagrams are described as a scatter plot graph. Depending on the different characteristics of the star, the star will get placed in certain areas. They are categorized by luminosity, spectral type, surface temperature, the radius of the star, and the color. These categories can help determine a star’s lifetime, expectancy and actual age, and what sequence it is in. There is the white dwarfs, giants, super giants, and the main sequence.

H-R diagrams are also very useful in determining what type of star cluster one is looking at. This is because with an H-R diagram, it is easy to track the evolution of stars in the same cluster. This, in turn, makes it almost effortless to track what age the cluster is and what the future evolution will because stars in the same cluster are all generally in the same age range.

Main Sequence

The main sequence is the most common place for stars to be. They are acting normal and are not abnormally large, luminous, or hot. Stars in the main sequence are stars that are changing hydrogen into helium in their cores for energy, for the most part. Our star, the Sun, is in the main sequence and does the same thing. This is incredibly important for determining what a star will do at the end of its life and how long it will live.

There is a general rule that, on the main sequence, the brighter (or more luminous) a star is, then it will always be bigger than a less luminous star which is also on the main sequence. This is because there is a certain amount of balance in gravity that must be maintained. If a star has a high mass, then the core temperature must, in turn, also be high. If a star’s core temperature starts to grow, then the star must grow as well. This balance is shown throughout the entire main sequence. That is what makes it what it is. The relationship between mass and temperature to maintain the balance of gravity is what keeps stars in the main sequence.
**White Dwarfs vs. Super Nova**

It is a common misconception among people that white dwarfs are different from run of the mill star. This is most definitely not the case. White dwarfs are very small, extremely bright, and extremely hot stars, but they actually come from regular stars. There are very few that are known right now. Our own star, the Sun, is a star that will become a white dwarf when it reached the end of its life.

Many people know the term super nova, it means a large explosion. In this case, supernovas come from stars just like white dwarfs. The main difference though is that white dwarfs come from a relatively peaceful process, and supernovas are very violent explosions that leave nothing behind.

**Star Production**

**Before the Beginning**

Stars are very complex, and they start the same way. But before they may begin to form, there must be the ideal conditions for them to start in.

Diffuse nebulae are the ideal, actually the only, place for stars to form. That is because of all of the low level gases inside of them; like hydrogen, helium, and other slightly higher gases. These are the perfect ingredient to make stars. There must be ample amounts of this material because the following reaction happens rather quickly, and new stars need a lot of fuel and a lot mass to pull into itself.

**The Process**

It all starts with a ripple, a tidal force in the universe. These tidal forces are caused by a couple of different things. Supernova, black hole eruptions, and normal tidal flow of the universe is what causes these things to bump into each other. It is not a gentle love tap either, it is a rude, hard shove.

This causes atoms in the diffuse nebula to crash into each other. Once these atoms crash into each other, it could cause the atoms to collapse under their gravitational weight. This means that the hit each other with such a force that they collapse into each other and create one particle of bigger mass. Refer to figure 1 to see the process.

This continues the happen and the particle will start to attract other particles into itself. Eventually it will start to slow down. But as soon as the pressure supporting the large particle is less than the pressure of gravity, there is a second large collapse. This collapse happens incredibly fast. The gas particle will start to heat up, then it becomes what is known as a protostar.

**Protostar**

A protostar is basically just a baby star. It is what will be known as the star when it reaches full maturity. The protostar’s center starts to heat up, this is what is known as the core.
As the core heats up, it starts to burn, or convert, hydrogen to helium (this is known as nuclear fusion). The star starts to become a main sequence star. See Figure 1 to visualize the process of going from matter to protostar.

Becoming a main sequence star means reaching a stage of equilibrium. As mentioned earlier, main sequence stars are almost perfectly balanced between temperature and size. But it is not an easy task to become so stable. That is part of the reason it takes so long for a star to become a fully formed star. “It could take a star like our Sun 50 million years to fully form,” says NASA. It can official takes anywhere from one million to fifty million years before a star can go from protostar to full on main sequence stage stardom. All it takes is a little patience because the core has to reach a high enough temperature for stable nuclear fusion and the size has to be big enough to be classified as a star.

Star Clusters

Star clusters are pretty self-explanatory in the names themselves. They are clusters of stars that coexist in a specific area of space. There are two different kinds of star clusters, globular and open. The open clusters are much younger and more spread out than the globular clusters.

Clusters are believed to exist just because of the fact of how stars are made actually. It is very uncommon for just one area in a diffuse nebula to collapse in on itself and create a protostar, more often than not it will create multiple, instead of just one. This causes stars to form in close proximity to each other, all around the same.

Star clusters are a useful way to determine how stars will die. That is because, once all of the stars from one or many star clusters is placed on an H-R diagram, a pattern will start to show up. You will be able to easily track where the other stars will go because they are all roughly the same age.

Star clusters might not seem like it, but they are useful and important patterns in the universe.

Possible Outcomes

After a star is completely formed it carries on its life to have a couple of different outcomes. As a quick overview, a star has a couple of outcomes as it reaches the end of its life. These include becoming a white dwarf or ending as a supernova.

White Dwarfs

The more common outcome, though, is when the outer layers fluff off and become planetary nebulae as they float through space. The outer layers will float away until the still hot core is exposed and becomes extremely bright and very small. It also becomes a little unstable.
These are what is known as white dwarfs. For the most part, a normal white dwarf will fade into darkness and become a cold lump of mass, known as a black dwarf.

**Supernova**

A supernova is a supernova 9, 16, is an extremely violent explosion that a white dwarf goes through when the white dwarf is near another star and starts to attract that star’s matter. As the white dwarf attracts the other star’s matter the white dwarf will start going through mini explosions to dispel the acquired matter once again. This will happen in cycles. Once the white dwarf is completely out of fuel, it will sometimes collapse in on itself, causing a massive explosion. This leaves nothing but a black hole. This can badly affect the stars and planets around this supernova. A supernova happens when the core no longer has fuel to fuse, causing the core to go through a quick collapse. The sudden collapse causes a large explosion which can open a black hole. No one is quite sure about black holes so there is not a definite answer as to why this happens.

**Conclusion**

Stars are very complex beings. It is not everyday that a new star, or star cluster is born. With all of the right circumstances and ingredients, it happens quickly… well quickly in terms of the rest of the universe. This is such an important concept because there would be so much left undiscovered if scientists and astronomers didn’t know how stars worked.

I chose my topic because of its importance I feel that it has in the astronomy field. I think it is fascinating all the different things that have to happen in order for one star to be born. It just makes me think about how lucky we were to have the Sun made for us by the same process, but with just the right amount of everything. The right amount of heat, the right amount of light, the right amount of distance from us. All of these things are so important that even if they changed just a tiny bit, our world wouldn’t be able to exist.

The formation of stars has always been a fascinating topic for me. I thank the opportunity given to me to research this topic, write it, and get it published in the Symposium for other people to read. I feel that everyone could benefit from learning how our star, the Sun, was formed. Stars are such fragile beings. Yet they seem so big and indestructible. It seems crazy to think that they need any kind of help from the universe to make a new one.
Figures

Figure 1: This is a diagram showing how the force of gravity is pressing on itself in a molecular cloud, causing clumps, and then protostars to appear.


Figure 2: picture of star formation in a nebula much the one neas Orion’s belt. It shows the start of stars starting to form.


Figure 3: This is an excellent representation of what an H-R Diagram looks like. It has all of the criteria and has some plotted “stars” on there.

References

The Physics of Firearms

Ross Yalch
PHY112
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Professor Michael Swingler
Abstract

All you hear when you pull the trigger of a firearm is a loud bang, but what is really happening? What are the mechanics, chemistry, and physics behind the process of launching a small projectile out of a firearm? To start this discussion, first one must understand the most common mechanisms and mechanics behind the operation of a firearm. After that, a chemical reaction and transfer of energy is what puts the bullet in motion. While traveling down, and after the projectile leaves the barrel, what is happening? Finally, how have scientific developments over the past century improved the accuracy and technology behind firearms?

Discussion

Since China invented black powder in the 13th century, humans have been perfecting the firearm to be more lethal and more accurate. The early firearms involved simply packing a projectile into a bamboo tube, and igniting black powder behind it. At the time this may have been innovative, but this method is now not only primitive, but unsafe. To fully appreciate firearms and the power they hold, first one must understand how far they have come. In 1520 a German man by the name of August Kotter invented the rifled barrel. Rifling is the spiral grooves cut into the barrel of a gun in order to give a bullet rotation while traveling through the air. (See figure 1)

Although invented during the 14th century, this game-changing invention did not become a mass produced commodity until the 1850’s and 60’s. It was around this time that the change from single shot rifles to modern ammunition was made. This meant that instead of individually loading each shot, taking up to three minutes, one was able to simply load a prepackaged casing and bullet into their weapon. Due to the increase in speed and efficiency of this invention it made older firearms obsolete and completely changed the way we look at modern firearms. (2013, Supica)

The change in how ammunition was loaded into a firearm greatly changed the mechanics behind how these weapons actually operated. The most common name for modern ammunition is cartridge. For now, a cartridge will simply be looked at as the holder of the projectile. Before a weapon is fired, a cartridge is placed in what is called the chamber. The chamber is a thick metal hole at the base of the barrel, which is tightly sealed in order to ensure when a cartage is ignited as much kinetic energy as possible is forced down the barrel pushing the bullet out the end. (see figure 2) After a cartridge is placed into the chamber one must “cock” the hammer back on the weapon. Some hammers are visible on the outside of firearms and some are internal, and not visible. (see figure 3)
As the first law of thermodynamics states, energy cannot be created or destroyed, simply transferred. (NASA 2015) As your finger transfers kinetic energy into the trigger it simultaneously releases the hammer, then transferring the potential energy of spring tension into kinetic energy, and slamming the hammer into the back of the firing pin. The firing pin is a large pin-shaped object used to poke the end of the cartridge. In most cases, the firing pin will poke the center of the cartridge in an area called the primer. (see figure 3) This area then begins a chemical reaction inside the cartridge.

The transfer of energy from the firing pin to the cartridge sets off a small explosion inside of the cartridge. Inside of the cartridge is a specific amount of black powder, which holds high amounts of potential energy. When the primer is ignited, it simultaneously ignites the black powder converting its energy into kinetic energy. At the very end of the casing, and beginning of the projectile, is a small crimp that keeps the cartridge held into place. This crimp is tight enough to hold the object in place but not enough to hold it in when the black powder is ignited. As George Kingsley Zipf stated, objects will always take the path of least effort or resistance. (Chang 2013) When the explosion inside the cartridge occurs, this releases the tension of the crimp on the projectile, repelling it down the barrel and toward its target.

According Brian Benini, a mechanical engineer with Sinterfire ammunitions, the efficiency of a firearm is broken down as such. Two percent of energy is lost to barrel friction. This is due to the twist of the barrel, and the bullet bouncing around in the barrel. Heat transferred into gasses to push the bullet down the barrel is about 34 percent. Heat transfer from these gasses and the bullet rubbing against the barrel transfers to about 30 percent loss in energy to heat in the barrel. This means that about 33 percent of energy directly transmits to forward acceleration of the projectile. Also, the average, unspent powder left over after the chemical reaction is less than one percent.

As the bullet travels forward there is much pressure at the nose of the bullet in the form of a shockwave due to the fact that it is traveling faster than the speed of sound. This force exerts a force directly backwards, reducing the bullets velocity. Also, this force tends to push the nose from traveling perfectly straight. As a bullet travels through the barrel, the bullet slightly rubs along the inside. The rifling in the barrel then causes the bullet to rotate like a spiral thrown on a football. Rifling is measured by what is called twist. The equation for twist is below.

\[ \text{Twist} = \frac{L}{D_{\text{Bore}}} \]

$L$ represents the twist length to rotate the projectile one rotation, and $D_{\text{Bore}}$ is the diameter of the hole or bore of the barrel. When a bullet passes through the air on
its way toward its target, it passes through a strong aerodynamic pressure field. Without twist, a bullet would simply tumble through the air. This would greatly change the direction of travel making it very inaccurate. This gyroscopic motion helps the bullet cut through the air and keeps it traveling in a straight line. The average spin on a bullet, as it leaves the barrel, is around 24,000 rpm. (Schueman 2015)

The distance a bullet can travel at accurately has a lot to do with its shape. Rifle bullets are elongated and come to a sharp point. This is because their purpose is to travel far, fast and accurate. Pistol ammunition is much shorter and flatter at the tip. Pistol ammunition is used at close ranges and is flatter in order to transfer more of its energy over a larger area. Bullet weight and velocity have a lot to do with how much kinetic energy is transferred into a target. To calculate kinetic energy one can use the following equation.

\[ KE = \frac{1}{2} M V^2 \]

When you increase the mass of a bullet, without increasing the amount of powder to push the bullet out of the barrel, you lower the velocity. The amount of powder can be increased, but then the size and mass of the bullet must be decreased. Most bullets have a lead core with a copper jacket around the outside. These two metals are very dense, and give you the ability to have a smaller object with a large mass.

As soon as the bullet leaves the barrel of the gun, gravity, drag and wind force act on it. Because most sites are mounted above a weapon, the barrel must point up to meet these sites at a given point. When the weapon is fired, the bullet travels in an upward motion. There are two points in which the bullet and the sites are perfectly aligned. (see figure 4) The only force pulling the bullet toward the ground is gravity. The higher the muzzle velocity, the less time the bullet is subjected to gravity. This makes the trajectory path more shallow and gives the bullet a better chance at going further, more accurately. As drag, air resistance and wind force act on a bullet, the velocity decreases, therefore making the bullet drop faster.

Drag is determined by bullet speed, ballistic coefficient and air density. The speed of the bullet is simply the velocity at which it is traveling at any given time. The ballistic coefficient is calculated with the following equation.

\[ BC = \frac{M}{(C_d \times A)} = \frac{p \times l}{C_d} \]

- \( M \) = Mass of projectile
- \( C_d \) = drag Coefficient
- \( A \) = Cross section Area
- \( p \) = average density
$l = \text{body length}$

Drag coefficient is the way to measure the ability to overcome air resistance during flight. Air density has multiple factors, which include altitude, humidity temperature, and pressure. As these increase and decrease the air becomes denser or less dense. This is different in every location, making it impossible to site in a weapons system once, and have it be accurate everywhere. (Klatt 2016)

The force of wind also changes the flight path of your bullet. When shooting with the wind behind the direction of travel, your bullet will maintain its velocity for longer, therefore making it appear to be shooting higher. Shooting into the wind will slow the bullet down, and will allow gravity to pull the bullet closer to the ground. A left or right wind is much harder to determine and will push your bullet to the side. There are multiple ways to correct this. The two main corrections are adjusting your optic system to compensate for the wind, and the other simply involves shooting to the left or right of the target point.

**Conclusion**

A firearm, in its most basic form, is simply the transfer of energy from your finger, all the way to the target. Large amounts of potential energy via springs and chemicals instantly change into kinetic energy that launched a small object with enough force to kill. With further engineering and technology not only can weapons be more accurate and powerful, but safer.

Throughout my research, it was interesting to learn how both simple and complex physics principles determine the motion of a bullet. Although I already had a pretty large understanding of the mechanics of firearms, it was the science behind ballistics that intrigued me. I had always known that if you wanted to shoot further, you had to aim higher, but this was brought to a new light. While in the Marine Corps we always sited in our rifles at 36 yards. This was because the bullet was at the same height then as it would be at 300 yards. At 36 yards, the bullet was still traveling up, but by 300 it had begun to drop.

The future of ballistics is already at our doorstep. There is a company out of United States called Tracking Point. It is more an optics company than a firearms company. They use GPS, and up to date weather information to do all ballistics calculations automatically. This system works by placing a red dot on the object that you wish to shoot. Once the object is marked, the computer system auto adjusts the optic system for the environment and range at which you are shooting. You then pull the trigger. When you pull the trigger a second dot appears through the scope and waits to fire until both dots are aligned. This ensures unbelievable accuracy, and practically eliminates human error, which is something everyone can be excited about.
Figures

1) Figure 1 is a picture of barrel riffling versus a smooth rifle such as a shotgun. https://www.hunter-ed.com/images/drawings/bores.jpg

2) Figure 2 contains a picture of the chamber of a gun. This is designed to tightly seal in the burning of powder, and forcing all kinetic energy down the barrel of the gun. Due to the large amount of force put on this area, the chamber must be built very strong. http://www.opensourcegun.com/blog/author/glockman1727ak4774/
3) Figure 3 is a diagram of a cartage, showing its different components.

4) Figure 4 is a diagram of bullet trajectory. As shown, the bullet and sites are lined up at two locations. One close, and one far.

http://steamcommunity.com/sharedfiles/?id=538064538

*Not to Scale - Exaggerated for Effect*
References


Benini Brian. Sinterfire Ammunitions. Engineer and Project Manager. 200 Industrial Park Rd. Kersey, PA 15846. (814) 885-6672


Do Compression Swimsuits Make a Winning Difference?

Nancy Zanello
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Physics 112
Dr. Casey Durandet, PhD
Swimming is one of the few sports where the body is immersed in an aqueous medium. Swimmers do not face the torment of gravity on the muscles but this does not make competitive swimming an easy sport. Instead, the swimmer has to overcome viscosity and multiple forms of hydrodynamic drag. Every muscle in the body is engaged in fighting these forces in hopes of beating not the adjacent swimmer but the clock. In a sport where one one-thousandth of a second can make the difference between first and last place, a swimmer has to use every advantage possible. Some of these advances have come in the form of engineered swimsuits, with claims of optimizing body position, enhancing muscle stability, and drag reduction.

Any competitive swimmer will say swim technique is the most integral part of their strategy. A dynamic technique that is reconstructed with every race throughout a swimmer’s career. Competitive swimmers have two goals every time they enter the water, drag reduction and maximizing propulsion. Technological advances have created relied upon tools such as video analysis, which is used to view and evaluate stroke technique and body position in and on the surface of water. No two swimmers are alike, each with different mass, strengths, and abilities but small changes to the angle of a hand or foot, a swimmer can maximize propulsive force, allowing a drop of milliseconds. A millisecond may seem insignificant to most but in an objective sport where time is the determining factor, milliseconds will make the difference whether or not a swimmer medals.

Biomechanics in swimming have also advanced. Compression suits, once used by physical therapists and runners, have made their way into swimming. Boasting streamlining and improvement of hydrodynamics, swimmers are turning to these suits for their advantageous claims. USA Swimming claims these suits are used to test “…how effectively swimmers are using oxygen and using it for energy [1].” Claims of reducing the forces impeding swimmers is a lofty goal. A small study published in the International Journal of Sports Physiology and Performance conducted on runners, found no difference in performance except for two participants who showed significant improvement. Stickford, who conducted the study, believed this to be the result of the placebo effect [2]. A similar study conducted on cyclists found the perceived effects was significant during recovery which was positively affected with use of compression garments [2].

Swimming takes place in a different medium than most sports, water. The swimmer has to overcome viscosity, the resistance to flow, of water as well as the viscosity of air. There are factors that affect the viscosity of water. USA Swimming specifies the ideal temperature of water and air for competitive swimming; 82 degrees or lower for water and ideally air temperature for competitive swimming should be 78-80 degrees; never exceeding the water temperature by 2 degrees. The organization states the more aerobic the activity, the lower the air temperature needs to be [3]. Attention to such detail affects performance. Using Table C-2 [Fig. 1] of Wiley Online, the density of water at 80 degrees F is 1.934 slug/ft³. The dynamic viscosity, fluid resistance to flow, the force at which the internal friction is overcome is $1.799 \times 10^{-5}$ lb*s/ft². Kinematic viscosity differs from dynamic viscosity, it is the ratio of viscous force to the density of water which is $0.930 \times 10^{-5}$ ft²/s [4]. Air at the same temperature is $3.86 \times 10^{-7}$ lb*s/ft² and $1.69 \times 10^{-4}$ ft²/s respectively [5]. Water is 50 times more viscous than air and resistance is greatly increased when compared to air. Viscosity affects flow which in turn affects the swimmers on multiple levels. Initially the water in the pool is still, once the swimmer begins to accelerate, waves are propagated. The swimmer in the center of the pool has less turbulent flow.
to deal with than the outside swimmers. The Reynolds number is the method used to classify whether a flow is laminar or turbulent. \( N' = \frac{\rho v L}{\eta} \), where \( L \) is the length of the object, \( \rho \) the fluid density, \( \eta \) its viscosity, and \( v \) the object’s speed in the fluid. According to Wright [6] “If \( N'R \) is less than about 1, flow around the object can be laminar. The transition to turbulent flow occurs when the Reynolds number is between 1 and about 10. The smoothness of the surface will determine if there can be a turbulent wake behind the object with some laminar flow over its surface. For Reynolds numbers between 10 and \( 10^6 \), the flow may oscillate between the laminar and turbulent. When the Reynolds number is greater than \( 10^6 \), the flow is entirely turbulent. (See Figure 2 and 3.) In addition to waves created by swimmers, the waves reflected from the walls increase turbulence in the water, again increasing resistance, affecting flow and swimmer momentum.

Water, as a Newtonian fluid, resists flow. A swimmer applies all of Newton’s laws when racing. Henderson of The Physics Classroom, explains Newton’s first Law of motion states: an object at rest stays at rest and an object in motion remains in motion with the same speed and in the same direction unless acted upon by an unbalanced force [7]. Using this law, an object, the swimmer, has to overcome static force. The resistance when pushing off the platform is the static force. \( F_s = \mu_s F_n \). \( F_s \) is static friction, \( \mu_s \) is the coefficient of static friction, and \( F_n \) is the normal force. Once off the platform, the swimmer must use the momentum created to accelerate in the aqueous medium. This is where Newton’s 2nd Law applies [7]. \( F = ma \), \( F \) is force, \( m \) is the mass of an object and \( a \) is the acceleration of that object. If two swimmers of the same mass are standing on a platform, the swimmer applying the greater net force will have the higher rate of acceleration with which to move in water. The greater the magnitude of force applied, the longer the distance swam in a shorter time frame. Once in motion the swimmer uses the dynamic force to continue to displace the medium to propel him or herself forward. Newton’s 3rd law, a conservation of energy, states that for every action there is an equal and opposite reaction [7]. The action – reaction describes the momentum created by the swimmer applied to the water. When a swimmer pushes water back toward the feet, the momentum propels the swimmer forward. Momentum created must be maintained to keep the swimmer moving forward. Momentum depends on two variables, \( p = mv \), where \( p \) is momentum, \( m \) is the mass of the object, and \( v \) is the velocity of the object. By forcing water back the result is the swimmer moving forward and the force used to move the water backwards, is the same force that will push back at the swimmer. When the force applied to an object acts in the same direction as the object’s motion, the object will consequently speed up [7].

The work done by the swimmer, can be derived by \( w = F \cos \theta \), \( F \) is the force applied to the water and \( d \) is the distance the water is moved. The water moving with backward momentum acquires this kinetic energy. The kinetic energy equation is \( \frac{1}{2}mv^2 \), where \( m \) is the mass and \( v \) is the velocity squared. Work is done when the water is moving in the direction of the force [8]. A large percentage of work performed by the swimmer is converted into kinetic energy of the water displaced, accelerating the mass and increasing its speed [8]. At the competitive level where swimmer velocity is increased, the pressure in the surrounding water is affected and creates a pressure gradient similar to that of a boat moving through water (Fig. 3). Both work and kinetic energy measure the amount of motion of a body. Ideally, in a calm pool, no waves are created and momentum and energy are conserved. Competition pools are not ideal, involving eight swimmers, each working to create momentum and working to overcome turbulent flow, positive acceleration, and conservation of energy. Swimmers prefer to be in the center lanes.
instead of the outside lanes because without understanding the mechanics they feel the forces against them. The higher the velocity of the swimmer the higher the wavelength of the vortices created by each stroke and kick. In addition to the disturbances created, as a swimmer accelerates in the water, a trough is created around the swimmer\'s body leading to a greater use and consumption of energy by the swimmer. Rolling of shoulders and slight movement of the body will reduce wave drag. By minimizing the cross sectional area in contact with water, smaller waves are created decreasing wave interference. (See Figure 4). Swimmers in the center lanes have only their own trough and waves to contend with. The swimmers on the outer lanes have four times as many waves to work through plus the addition of reflected waves, in addition to their own trough to swim out of. Creating a simulation of swimmer body position in water, Marinho, Barbosa, Rouboa & Silva, (2011) found a greater amount of the swimmers kinetic energy is lost as potential energy of wave formation [9].

Frictional drag created by the constant collision of water molecules with themselves and the swimmer, will slow a swimmer down. Frictional drag is required to propel the swimmer forward. Pressure caused by drag occurs as the swimmer accelerates through the water, building around the swimmer\'s head. Water here pushes back at the swimmer\’s head at a greater gradient than the pressure at the swimmer\’s feet, thereby creating another turbulent flow in addition to the swimmer\'s arm and leg movement. The pressure around the head creates a “separation point” [9]. The pressure of swimming at the surface is different above the swimmer, in air, and below the swimmer, in water. As can be seen, the amount of forces working against the swimmer are what make it such an arduous sport. Despite the difficulty, the forces against the swimmer can be reduced at the start upon leaping off the platform.

Drag, specifically hydrodynamic drag, involves drag of friction, of pressure, and wave drag, are the greatest obstacles faced by swimmers because it is amplified with the different medium. With a rudimentary application of physics, it was generally accepted to reduce drag forces the hand had to be perpendicular to the direction of motion. Surface drag depends on the smoothness of a surface led to many swimmers shaving their body hair. The idea is the smoother the surface the less turbulent the flow. Another form of drag is related to buoyancy and the body position of a swimmer in the water (Fig. 3). Drag is calculated by $F_D = C_D \rho A(v^2/2)$. In swimming, all aspects of body positioning and swimsuit attempt to reduce drag as much as possible. Through study and evolution of the sport, swimmers used physics and applied lift forces to maximize movement. Lift is the net force acting perpendicular to the relative movement of fluid, calculated by this equation: $F_L = C_L \rho A(v^2/2)$. Bernoulli\’s principle explains how lift is applied to swimming (Fig. 4). $P +1/2 \rho v^2 + \rho gh = constant$, where $P$ is the absolute pressure, $\rho$ is the fluid density, $v$ is the velocity of the fluid, $g$ is the acceleration due to gravity and $h$ is the height above some arbitrary reference point.

To minimize drag and create lift, swimmers streamline their body as they enter the water. Pressure is force per unit area (square meters). Bernoulli\’s equation states that in a streamline fluid flow, the greater the speed of the flow, the less the static pressure; and the opposite is true and this was tested and confirmed by NASA [14]. By reducing the amount of area facing the direction of motion, drag can be reduced. To reduce hydrodynamic drag the swimmer assumes the following position, prone position, arms fully extended above (or in front) of their head (Fig. 5). The favorable body position of swimmer gliding are prone with arms extended in front of the head is more favorable than prone with arms aside the trunk, as computed by Marinho, Barbosa,
Rouboa & Silva (2011). Swimmers adjust limbs and body position to reduce the cross sectional area, leading to an increased momentum and increased propulsion while under water. Time spent completely submerged is also regulated because of increased speed under water than on the surface due to wave drag. A techniques used by swimmers is “sculling” (Fig. 6). By slightly cupping their hand, fully extending the arm, a swimmer creates lift similar to that of an airplane wing, created by the different pressures above and below the swimmer. Full extension of limbs and body decreases the wave and drag resistance. The pull creates the propulsion force needed to accelerate forward using the physical forces to one’s advantage.

Bouyancy is another factor swimmers deal with. Archimede’s principle states: an object in a fluid medium will experience a buoyant force equal to the volume of fluid which is displaced whether or not the object is fully submerged. This principle is closely related to density. Rushall (2007) explains that once a person is submerged in water, the center of mass is not always the same as the center of buoyancy. The human body is comprised of bone, muscle, fat, air in the lungs and other structures, as well as fluids. Every swimmer trains to have low fat content and maximize strength while remain lean not bulky. The proportions of fat to muscle in a swimmers physical make-up will determine the specific gravity of that swimmer, the ability to float, and the characteristics of floating. The center of buoyancy is “the point in which the buoyant force acts” (Rushall, 2007). Most people tend to be bottom heavy, which is why when submerged in water the legs begin to sink and the body position rotates until the centers of gravity and buoyancy are aligned vertically (See Figure 7). Swimmers prefer a horizontal position in water and those who attain it expend less energy to swim. Having a buoyant rear, decreases the drag and wave resistance created by the shape of the human body. Body composition varying from one swimmer to the next contributes to all the forces and techniques applied in swimming. The main point contradictory to common thought is a swimmer will move faster submerged, under water than they would above it. Rushall (2007) stated the displacement of water uses the highest amount of energy and by the reasoning resistance is proportionally reduced by any amount of water that goes over the top of the swimmer.

Human shape is not designed for swimming, which gives swimmers a negative acceleration. Even streamlining, progress ends quickly due to the maximum velocity limit. As such, the best method but nearly impossible to attain is that of a “planning hull” [10]. Steve Haakes, a sports engineer explains why humans are not adapted for water [11]. By studying fish, and the undulatory method of swimming, it is obvious that a wave essentially runs the length of the body to a fin, the wave reaching maximum amplitude at the end of the caudal fin [12]. Flow velocity of a fish is the vector sum of the initial forward velocity and the fin movement direction. A perpendicular force, lift, is angled forward and parallel to oncoming flow creating a thrust. (See figure 8.) Positive thrust is created in both upward and downward movement. The tail (legs in humans) optimizes propulsion and various fins such as pectoral fins (arms in humans) provide lift and maneuvering capability. Different fins provide different functions aiding the fish. Dorsal fins provide stability and pelvic fins provide more lift. Unfortunately there are no similar equivalents in humans.

Greatly disadvantage in water, swimmers turned to expert help of scientists to design technologically advanced swim suits boasting drag reduction and streamlining of swimmers. The 2008 Olympics in Beijing was an example of where high tech suits resulted in record breaking swim times. Swimmers were highly impressed by the results, the Federation Internationale de
Natation (FINA), the governing body of swimming, was greatly disturbed, banning the suits for future competitive use.

Compression suits claim to "...positively impact performance, including drag reduction, assistance with recovery and increased circulation". According to US Master Swimming, compression suits target large muscle groups aiding in blood flow, warming muscle, reducing muscle vibration and support body alignment. By shaping the body to decrease drag, the swimmer will optimize the forces applied to propel through the water at a greater speed with less effort. The design behind the suit, flat stitching and the suit closer to the body by essentially, taking the suit a second layer of skin, drag will be decreased. A FINA guideline is that suits remain permeable yet water repellent which reduces drag. The highest levels of friction occur at the water molecules closest to the body. One suit boasts higher compression on the legs, fully bonded seams, kinetic taping, ab activators, and laser cut straps. The manufacturers claimed to Morrison in 2012, that their suit “compressed a swimmers body into a streamlined tube and trapped air, adding buoyancy and reducing drag”. This company utilized an internal team of 19 supplemented by hydrodynamic experts, aircraft engineers, nano textile producers in addition to experts in kinesiology, biomechanics, fluid dynamics and even a sports psychologist, indicating the level of commitment and science applied to their design. Compression suits are not a fad or purely psychological effect (Morrison, 2012). The suit is designed to constrict the stomach the least and the chest, buttocks and hips the most, attempting to mold swimmers into an unblemished tube [13]. The swimsuit boasts significant results when measured against a standard suit. The suit reduces passive drag by 16.6 percent and active drag, the resistance at the surface, by 5.2 percent, and improves oxygen economy by 11 percent.” [13], meaning a swimmer can maintain the same speed while consuming less energy. With the assistance of NASA’s Aeronautics Research Mission Directorate, which focuses primarily on improving flight efficiency by way of fluid dynamics, the forces of pressure and viscous drag, which are the same for bodies moving through air as for bodies moving through water, were identified [16]. Viscous drag was approximately 25 percent of the total retarding force in swimmers [14]. Lockney of NASA, stated the outcome was a suit that reduced skin friction drag by 24 percent (Fig. 8). These suits are designed specifically to compress the body but not to restrict breathing or inhibit movement.

Despite the records broken at the 2008 Olympics, opponents claim this phenomenon was not the result of increased technology on the physiological state of the swimmer. Instead it was the result of the psychological state of the swimmer that led to improved times. Of the testing trials that occurred with the suits by elite swimmers worldwide, perhaps it was a bit of both, if the swimmer believed the compression suit increased performance, then the swimmer had an improved performance. As Strickford found in her study, the placebo affect can affect physiological state [2]. Phillips interviewed numerous scientists with similar results as Strickland, although none of the studies were conducted on swimmers. Phillips found that “the mechanisms behind compression are still unclear, but warming, proprioception, and assistance in the mechanical removal of waste products are all common theories” [15]. The only conclusive evidence is anecdotal by the Olympic swimmers. It may not be clear if compression suits do positively contribute to faster event times, at the very least, we can see the swimmers and FINA believe the suits aided in the record setting times. One can say with certainty that the technologically advanced suits are not harmful.
Figure 1: [online library.wiley.com/doi/10.1002/9781118131473.app3/pdf](online library.wiley.com/doi/10.1002/9781118131473.app3/pdf)

Table C.2
Physical properties of water (U.S. customary units)

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<th>Temperature T (°F)</th>
<th>Specific Weight γ (lb/ft³)</th>
<th>Density ρ (slug/ft³)</th>
<th>Dynamic Viscosity μ (x 10⁻⁵ lb·s/ft²)</th>
<th>Kinematic Viscosity ν (x 10⁻⁵ ft²/s)</th>
<th>Surface Tension σ (lb/ft)</th>
<th>Modulus of Elasticity E (10^3 lb/in²)</th>
<th>Vapor Pressure p_v (lb/in²)</th>
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*At atmospheric pressure.

*In contact with the air.

Figure 2: Figure 12.18 (a) Motion of this sphere to the right is equivalent to fluid flow to the left. Here the flow is laminar with $N\chi R$ less than 1. There is a force, called viscous drag.
Figure 3:

Figure 4: http://illumin.usc.edu/assets/media/637/ii7_142_swimming_fig1.jpg

Figure 1: A Ship's Wave Formation
Figure 5:

Small drag in streamlined position

Large drag in unstreamlined position

Figure 6: http://msnbcmedia1.msn.com/j/MSNBC/Sections/NEWS/SwimmingStrokes.grid-6x2.jpg
Figure 7:  http://coachsci.sdsu.edu/swim/bullets/float36.htm

**Figure 2.** The roles of the centers of buoyancy and gravity and how they determine the angle at which a swimmer floats.

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Figure 8:

**The physics of swimming**

Jul 19, 2012
REFERENCES:
6. Wright State University. FLUID DYNAMICS AND ITS BIOLOGICAL AND MEDICAL APPLICATIONS Internet: http://www.wright.edu/~guy.vandegrift/openstaxphysics/chaps/12%20Fluid%20dynamics.pdf
12. Kreuger, P. “Undulatory Swimming” Internet: http://lyle.smu.edu/propulsion/Pages/undulatory.htm

How Does
the Earth's Magnetic
Field Affect its Surface?

Nadia Zia
November 19, 2015
PHY112
Dr. Durandet
Abstract

Magnetic field of the Earth is generated by moving and flowing the molten metals inside the Earth’s core. Materials inside the core are in liquid due to the high temperature of this region. The magnetic field lines of the Earth protect the Earth’s atmosphere from the solar wind, which are the movement of energized, charged particles such as electrons and protons from the Sun to the solar system. The Earth’s magnetic field is getting weaker. One reason is because by having many times reversal magnetic field every 450,000 years; it is possible that the Earth’s magnetic force is getting weaker. Another reason is by erupting lava in some parts of the Earth; the molten metals erupt and convert to solid metals. So, there is no liquid in order to create enough magnetic fields by flowing and churning them. If this magnetic field did not exist, the atmosphere would be substantially reduced and without magnetic field, there would be no atmosphere and life on Earth.

Introduction

History of studying Magnetism

According to Ronald Merrill, the author of, “Our magnetic earth”, in Greek mythology, a mountain in Magnesia had a power to pull the iron nails out of walker’s shoes. The mountain was probably consisted of iron ore and ferric oxide, which we call them magnetite now. Some philosophers began studying on magnets and magnetism in 5th century BC, such as Diogenes but in the book, “Physics around us” which is written by Ernest Henley and Gregory Dash, they state, serious study performed by William Gilbert in 16th century CE. Gilbert in his book, “De Magnete”, scorned the meta-physicians for not being practiced in the subjects of Nature, such as magnetic experiment but, before Gilbert’s research, people used magnetic compass for navigation. This compass is a small magnet that is free to a pivot and tends to align along the earth’s rotation axis. One end of the magnet shows North Pole and the other end shows South Pole.

What is a magnetic field?

Magnetic field is a region which is created by moving charged particles or charged ions in a space. It is also produced around a magnetic material, such as Iron, Cobalt, and Nickel in a dipole field lines.(Henley and Dash, 2012) (Figure1.1)

Magnetic field is driven by a magnetic motive force, mmf. Mmf is analogous to the emf of an electrical circuit. Mmf is referred to a permanent magnet or a current-carrying wire. The strength of magnetic field is proportional to the mmf production and the magnetic susceptibility,
or $\mu$ of the medium. Therefore, the magnetic field is induced when iron or any magnetism material getting close to a current-carrying wire or inside the current-carrying coil. 3

The magnetic susceptibility of all material depends on their temperature. At high temperature, some materials are not strongly affected by an external magnetic field, such as Iron. When the Iron gets cooled, the attraction gets increased. Magnetic field is driven by a magnetic force, which is analogous to the emf of an electrical field. So, the field strength of a current-carrying wire is increased when Iron or any other magnetic material is getting close to the wire. 3, 9

What is Planetary Magnetic Field?

There is a magnetic field inside and outside some planets, which is called Planetary Magnetic Field. This magnetic field resembles that planets have a giant bar of magnet in the center of themselves. Most planet is consisted of inner core, outer core, mantle, and crust (Figure 1.2). The core is composed of molten metals, which are magnetic material and magnetic field is created by moving and churning these liquids at a planet’s core. 6 (Luhman and Russell, 1996)

How the Earth Magnetic Field is generated?

As Luhman and Russell published an article in “Encyclopedia of Planetary Sciences”, they declare that magnetic field of the Earth is generated because of moving and flowing the liquid materials inside the Earth’s core. Materials inside the core are in liquid due to the high temperature of this region. The magnetic susceptibility of all materials is temperature dependent. At high temperature, iron is not strongly affected by an external magnetic field. When iron gets cool, the attraction increases gradually. So, at a definite temperature, which is called ferromagnetic Curie temperature, or $T_C$. 6

The Earth’s core high internal temperature is produced by combination of the heat due to radioactivity and the internal friction due to tidal motion of liquid materials. These dense liquids, which are mainly composed of iron and nickel, flow outward, toward the Earth surface and the Earth’s rotation. They carry electrical currents; therefore, creating magnetic field, which is analogues to the magnetic field around a bar magnet. Magnetic field is present around some planets.3

The magnitude of the magnetic field is measured in units of ampere per meter, tesla, or milli tesla. Researchers examined this magnitude of the magnetic field at the Earth's surface and claim it is from 25 to 65 microteslas or 0.25 to 0.65 gauss. 5

How is the Shape of Magnetic Fields in the Earth?

The dipole of the magnetic field is originated from the center of the planet and aligned along to its rotation axis (Henley and Dash). This magnetic field is extending up to 1, 05,600 km around it. Magnetic field lines mostly exit from Southern hemisphere and enter in the Northern
hemisphere. Spherical bundle of the magnetic field lines are in the core of the Earth. Earth’s magnetic field is in geocentric axial. But, this magnetic field has reversed many times throughout the planet's history. (Figure 1.3)

In an article "ESA Launches Satellites to Study Earth's Weakening Magnetic Field", the author says that this every 450,000 years the magnetic field reversal is being occured, and the last reversal was occurred about 800,000 years ago. They presume that another magnetic field reversal could be occurred in future.8

The magnetic field is generated by dynamo effect in the Earth. This dynamo effect, which is the result from of movement and circulation of the planet’s liquid core, which is rich in Iron ions in the Earth.3

Content

How does the Earth’s Magnetic Field Effect on the Earth’s Surface?

The flowing molten metals inside the core of the Earth is the driving force for the continental drift. The continents have been drifting during Earth’s history and between two continents, there is a sea floor. By pulling apart of two adjacent continents, the fresh magma is flowing in by some centimeters per century. But, the Earth’s magnetic field has change the direction of this movement remarkably. These changes has been recorded in the sea floor and the result has shown that the bands of magnetization are pointing in different directions. These bands of magnetization are induced by the Earth’s magnetic field. This Earth’s magnetic field, is frozen into the lava during eruption and then cools below the ferromagnetic temperature.9

How does the Earth’s Magnetic Field Effect on the solar wind?

Solar wind is a movement of energized, charged particles such as electrons and protons from The Sun to the solar system. These particles are deflected outwards by the magnetic field lines of the Earth and magnetic fields protect the Earth’s atmosphere from the solar wind. The magnetic field deflects the particles off to the north and south poles. Solar wind is first deflected in an imaginary surface, which is called the bow shock. (G. Hulot and etal, 2009). This bow shock creates in from of the magnetic field of the Earth and it is similar to the wave that happens in from of a boat when the boat moves pass through the water.3

The solar wind is also the reason of breaking up the molecular hydrogen in water. The loss of water in planet Mars is because this planet does not have a magnetic field in order to protect itself from the solar wind. As water comes to the space of the Mars, molecular hydrogen of water are broken down into the space and form water vapor in its atmosphere. This is why Mars has little water vapor.3 (G. Hulot and etal, 2009)

This solar wind contributes to the changes in the space which is surrounded the planet’s surface. They intense the energized particles and form clouds of these particles. These clouds are called coronal mass ejections and reach to the surface of the earth in 3 to 4 days.7 Magnetic fields
of the earth deflect these clouds, but some of these particles penetrate through the magnetic field; specially, more in the north and the south poles of the earth. They change the shape and lines of the magnetic field of the Earth.7

If this magnetosphere did not exist, the atmosphere would be substantially reduced because of bombardment of solar wind particles. Without magnetic field, there would be no atmosphere and life on Earth. By changing the magnetic field, the Earth is no longer able to guard against the deadly solar radiation.4 (Henley and Dash)

By increasing of penetration of solar wind or energized particles through the Earth’s magnetosphere, the Earth starts growing warmer because solar wind is the Sun’s byproduct, which is created in the Sun’s internal processes during making sunspots. But this theory is negligible in comparison with global warming, gradual increasing temperature of the Earth's atmosphere due to accumulation of industrial gases.1

Does The Earth’s magnetic field is getting weaker?

The Earth’s magnetic field is getting weaker. This data was collected by a European Space Agency (ESA) satellite array called Swarm2 (Dickerson, 2014). According to Kelly Dickerson, who is the writer of “Earth's Magnetic Field Is Weakening 10 Times Faster Now”, she claims, there are some big and weak spots in the magnetic field, which is 370,000 miles above the planet's surface. It has sprung up over the Western Hemisphere according to the magnetometers onboard the Swarm satellites.2

There is no evidence that why exactly the Earth’s magnetic field is getting weaker. But, one reason is reversal in magnetic field many times throughout the planet's history. It causes the magnetic field gets faded. Another reason is by erupting lava I some parts of the Earth, the molten metals erupt and convert to solid metals. So, there is no liquid in order to create enough magnetic field by flowing and churning them.7 (Thompson, 2008)

More changes to the magnetic field of the Earth happen at lower altitudes closer to the ground. These changes can produce many problems with electrical equipment, GPS signals, and power systems on the surface of the earth, such as disruption on navigation systems, communication systems, or on satellites. Furthermore, changing magnetic field can effect on animals as well.7 (Thompson, 2008)

How does the Earth’s Magnetic Field Effect on Animals?

Animals of all sizes have the ability to sense Earth’s magnetic field. Even tiny animals such as single-cell animals have this ability to orient along with this lines of magnetic fields. Some animals can sense the magnetic inclination while other can sense the horizontal component of the Earth’s magnetic field.5

There are many electoreceptors, which they acts like detectors, exist in the head of some marine animals such as sharks and rays. They are called ampullae of Lorenzini and they are used when the animal moves its head. An electric current is produced in its moving loop by the magnetic field of the Earth. This induced current is used for navigation and orientation. If the
magnetic field of the Earth is changed or weaken, they would no longer find their ways in long migrations.\(^5\) (Merrill, 2010).

There are some termites in Northern Territories of Australia which are called magnetic termites. Their scientific name is *Amitermes laurenis* and build their mounds in order to overcome flood in monsoon season by using Earth’s magnetic field. As Ronald Merrill, in his book, “*Our Magnetic Earth*” claims, “Since the termite workers are blind, they are not using crypto chrome to sense the magnetic field. They must employ other mechanism that some bees do.” Then, he cited James Gould, Kischvink, and Ken Deffeyes that they concluded bees use magnetite particles in their abdomens for finding their ways to go back to the hive. So, by changing Earth’s magnetic field, these animals would be in trouble.\(^5\)

The Earth’s magnetic field also has affection on the immigration of the birds. They navigate by this sensing this magnetic field. Their brain is like a biological compass and it is tuned to the direction of the magnetic field. They sense the both the direction and the strength of the magnetic field. If the magnetic field is changed or weaken, they are no longer to be able to recognize their way or orient themselves.\(^5\)

Some scientists believe that there is some reaction in the eyes of the birds which are sensitive to the magnetic signals. Merrill cited William Cochran’s research about pigeons in his book. He says that homing pigeons require blue light along with olfactory sense to feel the magnetic field. He adds result of Cochran’s research which was involved use of sunset to calibrate the magnetic compass during migration. They changed the magnetic field by 90 degree at sunset and when sunset was completed, the imposed magnetic field was removed. The birds flew in their path by 90 degree relative to a control group. There are probably other mechanisms for the birds to sense the Earth’s magnetic field, which need to be discovered.\(^5\)

How does the Earth’s Magnetic Field Effect on the Climate?

Magnetic field also effects on changing climate; specially, in high altitudes. The authors of “Scientists Discover How Events in Space Effect Climate on Earth” in *Washington blog*, claim that Changes in the Earth’s magnetic field causes some parts of the Earth amount of cooling, such as North America, but causes amount of warming over other parts of the Earth up to 12 degrees. Also, Courtilot, Gallet, Le Mouël, Fluteau, and Genevey in there Article, “Are there connections between the Earth's magnetic field and climate?” declare, that magnetic field of the Earth’s internal can at times have a significant influence on climate itself. They claim that this affection is possibly through the low-cloud connection or cosmic-ray at times of extremal tilt of the dipole.\(^1\)

How does the Earth’s Magnetic Field Effect on Human beings?

By moving blood in circulatory system and by having motion of internal organs, such as heart beating, a person can generate electrical fields and currents. The Earth’s magnetic field can effect on the speed of the flowing blood in major arteries and veins. The possible consequence is arrhythmia, which is abnormal rhythm of heart beat.\(^5\) (Merrill, 2010)
Some people believe if they sleep perpendicular to the lines of the magnetic field and their body cut the geomagnetic field at right angles, they sleep better. In the other words, they tend to put their bed aligned in north-south direction. But, the Earth's Magnetic field is not that strong enough to influence it. There is no reason for such an effect to occur.

**Conclusion**

The Earth’s magnetic field has some effect on the Earth’s crust or surface. One of these affection is changing the directional of the continents drifting. This change in direction of the movement is occurring due to changing the direction of the Earth’s magnetic field. This Earth’s magnetic field, is frozen into the lava during eruption and then cools below the ferromagnetic temperature. These magnetic field lines of the Earth protect the Earth’s atmosphere from the solar wind, which are the movement of energized, charged particles such as electrons and protons from The Sun to the solar system. They change the shape and lines of the magnetic field of the Earth. If this magnetic field did not exist, the atmosphere would be substantially reduced because of bombardment of solar wind particles. The solar wind is also the reason of breaking up the molecular hydrogen in water. The loss of water in planet Mars is because this planet does not have a magnetic field in order to protect itself from the solar wind. As water comes to the space of the Mars, molecular hydrogen of water are broken down into the space and form water vapor in its atmosphere. This is why Mars has little water vapor.

In addition to changing the Earth’s magnetic field, it is weakening by time. There are some reasons for this theory. One reason is by having many times reversal magnetic field every 450,000 years; it is possible to decrease its magnetic force. Another reason is by erupting lava in some parts of the Earth; the molten metals erupt and convert to solid metals. So, there is no liquid in order to create enough magnetic fields by flowing and churning them. (Thompson, 2008) Without magnetic field, there would be no atmosphere and life on Earth. By changing the magnetic field, the Earth is no longer able to guard against these deadly solar radiation. More changes to the magnetic field of the Earth happens at lower altitudes closer to the ground. But, more changes in climate happens in upper atmosphere because the global warming effects on climate changing more that effects of magnetic field changing. These changes can produce many problems with electrical equipment, GPS signals, and power systems on the surface of the earth, such as disruption on navigation systems, communication systems, telephones, televisions, or on satellites.

Furthermore, changing magnetic field has some effects on human being. The Earth’s magnetic field can effect on the speed of the flowing blood in major arteries and veins. The possible consequence is arrhythmia, which is abnormal rhythm of heart beat. Changing magnetic field has some effects on animals as well. For example, single-cell animals, marine animals, magnetic termites, and some bees are affected by the Earth’s magnetic field. The magnetic termites build their mounds in order to overcome flood in monsoon season by using Earth’s magnetic field. The Earth’s magnetic field also has affection on the immigration of the birds. They navigate by sensing this magnetic field. If the magnetic field is changed or weaken, they are no longer to be able to recognize their way or orient themselves. There are probably other mechanisms for the birds to sense the Earth’s magnetic field, which need to be discovered.
Figure 1.1. Magnetic field

Figure 1.2. Planetary layers
Figure 1.3 The shape of the Earth’s magnetic field

Figure 1.4. How magnetic field of the Earth effects on long immigration
Figures References


Cited References


