Foreword

Paradise Valley Community College is proud to present this 3-volume set of the 23rd Annual Mancini Science Symposium. This symposium was held on May 11, 2017 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 a record number of 105 papers this year. Dr. Mancini retired in 2012. After his retirement, the symposium was renamed the Annual Mancini Science Symposium.

Students enrolled in Astronomy, Biology, 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 3-volume set contains all 105 papers (4 in Astronomy, 25 in Biology, 18 in Chemistry, and 58 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:
Darra Browning, DVM – Biology
Scott Massey, PhD – Chemistry
Julie Olander, MS – Chemistry
Lori Prause, MS – Astronomy
Mike Swingler, MS – Physics

As instructors and faculty advisors for these students, 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 – Physics
Symposium Coordinator
<table>
<thead>
<tr>
<th>Title</th>
<th>Page</th>
</tr>
</thead>
<tbody>
<tr>
<td>The Impact of Gravity on the Human Body</td>
<td>1</td>
</tr>
<tr>
<td>by Elona Aaron</td>
<td></td>
</tr>
<tr>
<td>RoboDoc</td>
<td>14</td>
</tr>
<tr>
<td>by Veken Adam</td>
<td></td>
</tr>
<tr>
<td>Physics in Construction</td>
<td>26</td>
</tr>
<tr>
<td>by Brandon Alcayde</td>
<td></td>
</tr>
<tr>
<td>The Potential Hazards of Talcum Powder and its Uses</td>
<td>35</td>
</tr>
<tr>
<td>by Ariel Alcocer</td>
<td></td>
</tr>
<tr>
<td>Using Chemistry to Explain Love</td>
<td>44</td>
</tr>
<tr>
<td>by Moshe Alishayev</td>
<td></td>
</tr>
<tr>
<td>The Physics Involved in Rotator Cuff Tears</td>
<td>51</td>
</tr>
<tr>
<td>by Dania L. Amaya Zapata</td>
<td></td>
</tr>
<tr>
<td>The Phenomenon of Sonic Booms</td>
<td>62</td>
</tr>
<tr>
<td>by Samantha Ansell</td>
<td></td>
</tr>
<tr>
<td>Ultrasound</td>
<td>71</td>
</tr>
<tr>
<td>by Madeline Arnold</td>
<td></td>
</tr>
<tr>
<td>Cloaking Devices: When Theory Becomes Reality</td>
<td>82</td>
</tr>
<tr>
<td>by Hailee Ball</td>
<td></td>
</tr>
<tr>
<td>How Our Biggest Fears Have Evolved Over Millions of Years</td>
<td>92</td>
</tr>
<tr>
<td>by Jordan Behm</td>
<td></td>
</tr>
<tr>
<td>Stem Cell Research: Mankind’s Next Big Step in Medicine</td>
<td>100</td>
</tr>
<tr>
<td>by Cameron Benvenuto</td>
<td></td>
</tr>
<tr>
<td>To Meat or Not to Meat? That is the Question</td>
<td>110</td>
</tr>
<tr>
<td>by Alexandra Bizzarri</td>
<td></td>
</tr>
<tr>
<td>The Effects of Apple Cider Vinegar on the Human Body</td>
<td>123</td>
</tr>
<tr>
<td>by Lauren M. Borucki</td>
<td></td>
</tr>
<tr>
<td>Nuclear Waste and Safety Issues Involved in Nuclear Waste Power Byproducts</td>
<td>130</td>
</tr>
<tr>
<td>by Abigail Burson</td>
<td></td>
</tr>
</tbody>
</table>
# Table of Contents

## Neutron Stars
by Nick Cioffi

## Railguns: A Beautiful and Devastating Application of Physics
by Johnathon Clark

## Using Carbon Dating to Retrace Human Kind
by Cierra Clement

## Mars: A Brave New World
by Ben Conrad

## Radio
by Austin Corona

## Arsenic: The Silent Killer
by Savannah Cummins

## 3D Printing
by Angela Davis

## The Optimization of Weightlifting with Basic Physics
by Shane Davis

## The Human Heart: The Physics of the Cardiovascular System, Causes of Lost Pressure, and the Maintenance of Homeostatic Balance
by K. Dennis-Mohler

## The Science Behind Laser Hair Removal
by Chandler Dillabo

## The Chemistry within Drinking Water
by Hannah Dixon

## Fission and Fusion
by Vesna Djukic

## The Physics of Nerve Impulses
by Joel Doolin

## Bioelectrics and Living Organisms
by Aubrey Downey
# Table of Contents – Volume I

<table>
<thead>
<tr>
<th>Title</th>
<th>Page</th>
</tr>
</thead>
<tbody>
<tr>
<td>The Physics Behind Gymnastics</td>
<td>284</td>
</tr>
<tr>
<td>by Macie DuBiel</td>
<td></td>
</tr>
<tr>
<td>Iontophoresis Therapy</td>
<td>295</td>
</tr>
<tr>
<td>by Janessa Dunbar</td>
<td></td>
</tr>
<tr>
<td>Physics in Nuclear Medicine</td>
<td>305</td>
</tr>
<tr>
<td>by Nada Elebrashi</td>
<td></td>
</tr>
<tr>
<td>The Physics of Surfing</td>
<td>316</td>
</tr>
<tr>
<td>by Sydney Esposito</td>
<td></td>
</tr>
<tr>
<td>Echolocation of Microchiroptera and Cetacea</td>
<td>327</td>
</tr>
<tr>
<td>by Chelsea Farnsworth</td>
<td></td>
</tr>
<tr>
<td>Aluminum Air Batteries in Times of Emergency</td>
<td>336</td>
</tr>
<tr>
<td>by Jessica Flores</td>
<td></td>
</tr>
<tr>
<td>The Relationship Between Exercise and Diabetes</td>
<td>347</td>
</tr>
<tr>
<td>by Logan Franklin</td>
<td></td>
</tr>
<tr>
<td>The Confocal Microscopy and Wide-field Microscopy</td>
<td>356</td>
</tr>
<tr>
<td>by Shima Golshan</td>
<td></td>
</tr>
<tr>
<td>Genetic Engineering Will Change Forever through CRISPR</td>
<td>369</td>
</tr>
<tr>
<td>by Gage Gommels</td>
<td></td>
</tr>
</tbody>
</table>
The Impact of Gravity on the Human Body

Elona Aaron
PHY112 -11492
Dr. Casey Durandet
April 20, 2017
Abstract

The effects of gravity on the human body have been widely studied throughout academia and professional scientific research. Specifically, the effects of gravity on the human body, organ systems, ageing, circadian rhythm, blood vessels, as well as twin astronauts exemplifies the incredible forces placed upon our bodies while on earth. As our bodies have adapted to gravity, the absence of gravity plays a critical role in specific disease and physiological process. Through this analysis one will understand the vital role of gravity on the human body, as well as the effects of microgravity on disease process.

Introduction

The effects of gravity has been widely studied by NASA due to the overarching implications on human physiology, ageing, and disease processes. Since the first trip to space in April 12, 1961, humans have changed their focus from basic space travel to many other topics (NASA, 2011). Specifically, space researchers have focused their observations around deepening our understanding of the effects of space travel on the human body. However, this experimental research comes with a price due to the effects of microgravity. Astronauts now know that even though the blank space is thrilling and is ours to explore, space is still an unknown and dangerous place. Our human bodies are adapted to various gravitational forces due to our evolution with gravity from the beginning of our existence. It is widely known that gravity affects humans not only physiologically but also pathologically. Many studies performed on this topic, by NASA, have shown that weightlessness in space negatively effects the human body. By studying the effects of gravity on the human body, researchers hope to extend the life of humans, eliminate disease process prevalence, as well as understand the underlying mechanisms of physiology. Specifically, we seek to outline the effects of gravity on the human body, organ systems, ageing, circadian rhythm, blood vessels, as well as twin astronauts. This research paper seeks to illustrate the effects of gravity on the human body and the negative effects of microgravity while remaining in space for extended timeframes (Morey-Holton, 2003).

Effects of Gravity on the Human Body

We all know that when life on earth started, it did so in the presence of gravity. This means that all systems of our body are dependent of the gravity for a proper function. Years ago, before NASA had developed most of their space technologies, researchers were studying the effects of microgravity by doing simulations. The problem with these early simulations were the short duration of time (minutes) compared to the long durations of time (months/years) during space travel. During the first space trips, which lasted for relatively short periods of time, researchers observed very few effects adversely effecting astronaut’s health. What seemed to be the most problematic effect was the bone and muscle mass loss (Morey-Holton, 2003).

Astronauts are researching and experimenting with the effects of microgravity on the human body while travelling to unknown planets. Space travel is the most important research study that NASA is doing. Since the first space flight, there are many physiological effects that have been discovered. We know about many physiological effects of space in our body, such as the effects of blood volume shifts in the body, as well as, how the heart becomes enlarged because of this volume shift subsequently increasing the blood flow in the upper part of the body. This negative feedback loop leads to astronauts drinking much less fluids, yet still ridding the body of the same or even more urine which causes a decrease of body fluids. Subsequently, this decrease in fluids causes a decrease in the production of erythro-protein and then decreases
the RBC count. Although, this decrease is not meaningful enough as our body is such a great creation with a wonderful ability to adapt to a return to Earth’s gravity (Nagaraja MP. 2001)

*Effects of Gravity on Human Organ Systems*

NASA always uses space shuttles to transport astronauts to space as they research the effects that space has on the human body. Astronauts are trained to take care of their health and the health of their crews. Similarly, they are trained to learn specific skills that become very useful when in space without any terrestrial help. It is a reality that going from an environment with gravity to another environment with microgravity will cause changes in an astronaut's body. According to NASA Administrator (2015). “Gravity is defined as an attraction force between two objects” (NASA Administrator 2015). Changes of the structure and function of the cardiovascular system in a microgravity environment are known to cause medical problems, especially during space flights. Microgravity has a much lower gravity compare to the Earth’s gravity. The circulatory system in astronauts is used to working opposed to gravity, but it receives different signals and stimuli when under the effects of microgravity due to the absence of gravitational forces. When in presence of microgravity, a human heart does not need to work as hard to send blood to the upper part of the body as it does when our body works against gravity. Consequently, a then lessened heart action forces the blood volume in the upper body to increase. When in presence of the microgravity, this excess volume in the upper body causes changes to the blood volume distribution and the signaling pathways become altered. The volume distribution and volume effects alter the signaled pathways that then influence blood pressure changes with each beat (Ball JR., & Evans CH. (2001)).

In normal conditions, the heart forces blood toward the body and the brain. However, when astronauts are in space, the heart does not have to work hard against the gravity since there is no gravity to pull the blood volume downward toward to lower portion of our bodies. The body feels this change and starts to work less and be less efficient, thereby lowering the stroke volume. The muscular nature of the heart creates a higher systolic and diastolic blood pressure in the absence of gravity due to the inactivity of the leg muscle groups in microgravity and the decreased demand for blood in the lower extremity. It is anatomically known that during a systole, the ventricle of a fit human heart being pumps blood through the veins and then increases the volume of the blood in the arteries, thereby, increasing the blood pressure in the arteries. Conversely, during diastole the blood goes into the heart which relaxing it and decreasing blood pressure in the arteries. When in space, the body experiences an increase in the systolic and diastolic blood pressure (Canright Shelley (2007), While humans walk on Earth, gravity cause the blood to accumulate mostly in their feet while forcing a higher blood pressure in the lower extremities up to 200 mmHg. This blood pressure is much lower in the brain, only about 60 to 80 mmHg. When people go to space, this critical difference in blood pressure ratio, between head to toe, becomes completely altered. Blood pressure reaches homeostasis when at approximately 100 mmHg. Due to these pressure changes while astronauts are in space they experience physical changes within the body, as well as external appearance. External appearance symptoms often occur as skin becomes puffy as their face fills with fluid while causing their legs to thin out as they lose about 1 liter of fluid in each leg. According to Nagaraja MP. (2001) "If you have less blood," explains Dr. Victor Schneider, research medical officer for NASA headquarters, "then your heart doesn't need to pump as hard. It's going to atrophy."(Nagaraja MP. 2001)
Bones tend to lose lots more Calcium through urination, which weakens the bones making them more prone to breaks. Also, the absence of weight in space due to gravity weakens the bones and the muscles. Regarding bones in microgravity, it is widely known that bone density loss is more extreme, at around 1% each month, with a total loss of 40 to 60%. The effects of microgravity, the absence of gravity, also cause muscle mass loss. Astronauts have noticed that in microgravity muscle atrophy quickens which then shows a reduction in strength. In microgravity, scientist has observed a subsequent confusion to the bodies perception, by muscle wasting measurement, indicating the muscle is not necessary any longer. Important muscles, such as the calves or the spinal column muscular support structure, are a vital the group of muscles maintaining posture. If these muscles are not regularly utilized, they will lose around 20% of their total mass. At a high percentage at approximately 5% a week, vanishing of muscle mass can happen very quick (Nagaraja. 2001).

Other factors influence the physiological response to a microgravity environment, such as weightlessness. Weightlessness causes confusion of the sense of up and down, or special awareness, for astronauts due to the vestibular system becoming unable to discern where the ground and ceiling is located. Space also causes an orientation issue causing confusion of where different body parts are located. In 1969, the astronaut Rusty Schweickart, from Apollo 9, had to cancel his space walks because of feeling sick and fearing choking to death. Therefore, an exercise routine very important for all astronauts while in space, as well as upon their return to earth. Even with 2 hours of exercise a day, adapting to the gravitational field on Earth takes extensive rehabilitation time (Howell E. 2013).

Recent research discoveries regarding microgravity in space illustrate eye orbit pressure changes causing vision issues in the astronauts. In normal conditions on earth, our bodies function in perfect synchrony in the presence of gravity. Our body develops with and is supported by gravity. As previously mentioned, in space our bodies lose muscle mass causing atrophy of the muscles, as well as early ageing due to muscle protein lost. According to Camillo (2013), “In space muscle, atrophy rate is approximately 5% per month, and VO2 max is reduced by 25% after few weeks of stay” (Camillo, 2013).

Other side effects of space traveling occur due to the loss of earth’s atmosphere, magnetosphere protection, and exposure to galactic cosmic rays. Space exposes astronauts to radiation causing skin cancer, central nervous deficiency, and many other disease processes. All these side effects would not be as problematic if the astronauts remained in space under the three to four-month timeframe. Research has suggested that disease processes effected by microgravity occur when astronauts remain in space for longer periods of time. It is well known that as astronauts return home, they suffer from excess thirst as a conseguence of low blood volume because thirst is the bodies way of equalizing blood volume (BBC (2016).

Similarly, muscle loss is regenerated with time as the astronaut’s exercise and recuperate. Astronauts say that it takes them one day on earth to recuperate their muscle loss for each day they spend in space (Nagaraja, 2001). Of note, bone loss was found to be the most problematic to regenerate. Consequently, bone loss is never recuperated entirely. Research suggests that astronauts must spend one year regenerating bone loss for each month spent in space. Astronauts are required to adhere to extreme exercise routines knowing there is a high probability that bone loss will not entirely regenerate. According to Nagaraja (2001), “Dr. Alan Hargens, recently of NASA Ames and now a Professor of Orthopedics at the University of California San Diego Medical School, it is important to keep astronauts in good physical condition. You want the crew
members to function normally when they come back to Earth ... and not have to lie around for long periods of rehabilitation” (Nagaraja, 2001).

A widely-known concept states that exercising is the key to a healthy body and mind. Unfortunately, exercising in space under microgravity is very hard when compared to exercising on earth. On earth, gravity exercises its force while pull us downward causing a resistance force for bones and muscles. This resistance force disappears when exercising in space. In recent years, there are many novel techniques that are being experimented regarding exercising in space. The most promising so far is a negative pressure chamber for the lower extremities developed by Dr. Hargens and his colleagues. This device is a treadmill which uses the suction of a vacuum cleaner to create negative pressure on the body to create a simulated body weight resistance similar to the force of gravity. Research suggests that this device supports the prevention of the weakening cardiovascular functions, muscle loss, and bone loss (Nagaraja, 2001).

Effects of Gravity on Human Ageing

From birth, our bodies require ample oxygen consumption for normal growth. However, when gravity is absent and oxygen is scarce, ageing and cellular hypoxia result. Cellular hypoxia, on the other hand, also cause a muscle mass reduction and muscle tissue loss due to the ageing process, also known as sarcopenia. Unfortunately, long stays in space, in the presence of microgravity, cause larger quantities of muscle proteins to be lost as our bodies produce less CO2 while accelerating the ageing process. According to Camillo (2013), “the Italian physiologist Angelo Mosso proposed in 1904 to supply balloonists (aeronauts) with compressed oxygen containing 8% CO2. If at night we feel oppressed all we should do is move; muscle contraction produces carbonic acid resetting the balance between the gases in the blood” (Camillo, 2013). Dr. Felipe Sierra, the Director of the National Institute on Ageing Division of NASA, said that he has been working on a mechanism, on the molecular levels, which can contribute to a declining function of the T cells. Dr. Sierra’s study will help researchers understand how zero gravity can affect the immune function and provide more information on the immune suppression effecting mostly older people (Figliozzi, 2014).

Human bodies resist diseases by using the immune system as a defense mechanism. When a foreign pathogen infiltrates our body, the pathogen is immediately confronted by the T-cells as the starting mechanism of the immune cells that fight the infection. Inevitably, diseases are developed when T-cells fail to recognize foreign and pathogenic microbes. As a consequence of ageing, our immune system weakens resulting in disease process symptoms. These symptoms are observed in astronaut’s bodies regardless of age and especially during prolonged stay in space. Some of these symptoms are impaired T cell activation and ageing related immune suppression (Figliozzi, 2014).

“T-Cell Activation in Aging is the first study to launch into space that is funded by the Biomedical Research on the International Space Station National Institutes of Health initiative.” (Figliozzi, 2014). By performing this experiment, Dr. Hughes-Fulford tried to find answers regarding how microgravity’s symptoms and consequences compare to the ageing process causing a change in immune cells. Dr. Hughes-Fulford also sought to discover how our immune system dysfunction results from ageing biochemical mechanisms. Dr. Hughes-Fulford also performed another study, called Leukin 2, where the purpose was to identify what causes the disruption of the immune system. This was the first study performed comparing the behavior of T-cell’s in microgravity with samples located in space under false gravity. Leukin 2 showed that
gravity can affect gene expression in humans while effecting T-cells by activating them earlier than they would be in earth’s gravity. Additionally, the Leukin 2 study also found extreme similarity when comparing healthy T-cells in space with elderly human T-cells (Figliozzi, 2014).

As a conclusion, Dr. Hughes-Fulford found the link between microgravity and ageing. According to NASA research, a correlation was suggested between the effects of living in under conditions of microgravity and similar effects of the ageing process in the elderly. Correlated supporting evidence include a decreased capability to fight gravity and lower activity levels leading to a sedentary life in both astronauts in space and elderly on earth. The findings suggest that decreased movement for their muscles and bones degenerate causing muscle atrophy, bone atrophy, and a decrease of their blood volume. Researchers hope that with a discovery regarding the part mechanism capable of sending signals to make muscles and bones stronger, our ability to produce pills, in combination with some exercise, might provide astronauts with a chance to reverse the effects of microgravity in space, as well as the ageing process here on Earth (Figliozzi, 2014).

Researchers are also working hard to understand microRNA, which are small molecules that might affect the cellular behavior and encoding of proteins. Researchers are also focused on samples of microRNA that travelled through space in comparison with samples on earth in order to discover the role that they play in T-cell activation. Dr. Hughes-Fulford said the purpose for this research is to find and test other ways microRNA can help the immune system. He hopes to be able to pinpoint their exact target and subsequently work with drug companies to develop drugs capable of modifying the immune responses supporting immunosuppressed patients (Figliozzi, 2014).

Effects of Gravity on Circadian Rhythm - IV

During prolonged stay in space, the circadian rhythm and sleep pattern are also affected. Astronauts suffer from sleep loss, body fatigue, and even psychological problems. Is noted that astronauts sleep an average six hours each night in space. The circadian rhythm is very important in regulating our body health. On Earth, our circadian rhythm is based on a twenty-four-hour day-night cycle. When astronauts go to space, they need to be in an outstanding physiological state to ensure impeccable high performance of body functions. Our body clock is a very important biological activity controlling many bodily functions, such as temperature, hormone levels, sleep awake cycle, metabolism, and mood (Guo et al., 2014). According to Guo et al. (2014), human’s circadian rhythm is “composed of a central and peripheral clock. The central clock is in the suprachiasmatic nuclei (SCN) of the hypothalamus, which functions as the master pacemaker by synchronizing physiological rhythms in accordance with Earth’s cycling environment. The central clock operates to synchronize the clocks in peripheral tissues” (Guo et al. 2014)

When in space, the circadian rhythm very often is disrupted and sleep deprivation affecting the sleep-awake cycle. Sleep deprivation causes a decrease in daily functions and performance. Consequently, a dysregulation of the circadian rhythm causes many health side effects, such as mood disorders, diabetes, depression, cardiovascular diseases, and cancers (Guo et al. 2014). In space, there are many factors that can affect the changes of the circadian rhythm. Some of the factors differing from an earth environment are being under the effect of microgravity, changes in lighting, working in a confined space, motion sickness in astronauts, and long and heavy work days (Guo et al. 2014). In space, even in a spaceship, the light intensity is so low compared to
earth that it interferes with the human circadian rhythm clock. All these side effects in astronauts affect their neuromuscular and cardiovascular system causing a definite effect on their working performance. Based on a survey regarding an experiment with a space shuttle crew, it was noticed that the most decreases in sleep hours were noticed during the first and the last day of their mission. It is normal for astronauts, when in space, to take longer to fall asleep. Therefore, the most widely used medications when in space are sleep medication (Guo et al. 2014).

Effects of Gravity on Blood Vessels - V

During our everyday life, average humans notice symptoms of a low heart rate. When the heart rate decreases, this causes its muscles to relax decreasing the cardiovascular support needed to support blood circulation around the body and especially the head. When we change position suddenly, for a very short time, the force of gravity imposes its effects on the heart and the blood vessels. Even though there are small moments where we feel the effect of gravity in our blood circulation throughout the body, overall, our cardiovascular system is well adapted to the constant pull of the gravity towards the center of the earth. A defense mechanism that our body uses during the moments when we stand up fast is to constrict leg vessels, thereby stopping blood from collecting only in the extremities. When astronauts travel to space, the gravitational pull towards the center of the earth does not exists (Delp, 2002)

When we are laying down in the supine position, our body feels the gravitational force equally in all parts of the body. We can observe these changes by standing for an extended time with our hands hanging above our heads. Subsequently, we can very clearly see how the veins in our hands become bigger and full of blood. This effect of gravity changes when traveling in space. Due to the microgravity environment, the majority of our blood accumulates in the chest and head area. This will then cause swelling of the face and neck blood vessels. Additionally, the lack of blood circulation in the brain causes a feeling of dizziness. It is essential that our heart, brain, and muscles are supplied with enough oxygenated blood to support normal body functions. For normal body functionality to occur, blood needs to circulate and return to the heart from the lower extremities. As the heart pumps the blood out, blood vessels bring the blood back to the brain (Dunbar, 2007). According to Dunbar (2007), some experiments are done to discover why our body can’t handle long trips to space. They believe that through these experiments, researchers will be able to find out which factors are responsible for stopping our bodies ability to tolerate the presence of gravity after spending long periods of time in space. They think the reason is the large accumulation of blood in the veins, as well as a poor functioning of the arteries helping to direct the blood back to the brain (Dunbar B.).

Effects of Gravity on Twin Astronauts

In March 2015, the twin astronauts Scott and Mark Kelly participated in a one year mission to discover the differences between twins in space and on earth. Mark remained on earth, while his twin brother Scott went to space. The purpose of this mission, the twin paradox study, was to study the effects of space travel and changes to human physiology in microgravity. The twin paradox study compared Scott (experimental variable) in spaces microgravity to his brother Mark (control variable) on earth under the effects of gravity. As mentioned in National Geographic News by Than (2010), Einstein’s theory of relativity, regarding the effects of time bending, was proven to affect time frames and earthbound distances (Than, 2010). Einstein’s theory of relativity explained how a clock ticks slower as the distance between the clock and observer increases. The theory of relativity was used for the twin paradox experiment to study
the ageing process of twin siblings traveling to space and hypothesized ageing would occur slower in space in comparison to earth. The more we feel the gravitational pull, the slower the time goes.

With an understanding of physics, we know that the gravitational pull is stronger the closer we are to Earth’s center mass (Howell 2013). The twin experiment is the dream experiment for many researchers. They will explore the genes in each brother while observing how they are affected by gravity and microgravity. Observed within the normal earthly ageing process, our gene’s telomere shortens. This twin study sought to understand how telomeres are affected in space. It’s also noted that a chemical called methyl attaches to our DNA by subsequently turning then on or off. By taking a sample of Scott’s blood prior to his travel to space, they will examine the quantity of methyl attached to his DNA and compare that sample to his blood in microgravity. Additionally, researchers seek to discover how normal bacteria found in our body will react when remaining for one year in space.

The twin study was one of NASA’s unusual experiments and consisted of the astronaut Scott Kelly remaining in space for a full year. According to NASA, they took genetic measurements of the twins before, during, and after the study to observe their genetic differences. Concluding the twin study, NASA discovered that when Scott returned from space, there was an observed DNA difference between the two brothers. Therefore, remaining in space produced a significant difference between Scott’s DNA methylation when compared to Matt’s DNA methylation. Researchers are still scouring the outstanding and unique data set resulting from the astronaut twin study. According to researchers, the most important finding of this study resulted from performing a comparison between the detailed human genomic mapping of both twins. From biology, we know that the length of telomeres is related to ageing. It was noticed that the telomeres of the twin that spent a year in space were elongated, compared to his brothers (Edwards & Abadle (2017).

Conclusion

This research paper concluded that many negative effects on the human body occur in the presence of microgravity when compared to humans in earth’s gravitational field. By outline the effects of gravity on the human body, organ systems, ageing, circadian rhythm, blood vessels, as well as twin astronauts, the current research illustrated these negative effects. As you can see, humans take gravity for granted on earth when traveling in space and living in microgravity. There are many adverse effects of spending extended timeframes in space. In order to eliminate the negative effects of microgravity, research scientist must continue to explore the causes and effects of gravity and microgravity on the human body.
“Yuri Gagarin: First Man in Space. On that day in 1961, Russian cosmonaut Yuri Gagarin (left, on the way to the launch pad) became the first human in space, making a 108-minute orbital flight in his Vostok 1 spacecraft” (NASA, 2011).

“Gravity attracts objects towards the center of the earth. Gravity: pulls objects on earth towards the center of the planet; holds the earth’s atmosphere in place; holds all the components of the solar system in orbit around the sun; holds all the components of the galaxy together” (BBC, 2014).
This is a simulation of what happens to our bodies in Microgravity. These Astronauts are performing microgravity training in the KC-135 airplane. (NASA Administrator, 2015)
“Identical twins, Scott and Mark Kelly, are the subjects of NASA’s Twins Study. Scott (left) spent a year in space while Mark (right) stayed on Earth as a control subject. Researchers are looking at the effects of space travel on the human body” (Edwards and Abadle, 2017)

(NASA. Rainey 2017)
This infographic highlights a few of the many human health issues and milestones Astronaut Scott Kelly will encounter during his one Year in Space.
Cited References


Abstract

With the utilization of robotics in complicated surgeries, surgeons are able to perform the required tasks with the use of precision offered by robots. The utilization of robotics minimizes surgery trauma, increases dexterity, offers a wider range of motion, and allows improved access to obstructed organs. The da Vinci Robot is a revised, upgraded and an enhanced machine from its predecessors and its original design was made by Leonardo da Vinci. A company in California was able to evolve the design of the da Vinci robotic system. Four models were made by Intuitive Surgical, the standard original model, model S, model Si, and the most recent one is model Xi. Thanks to the modern technological advancements of the da Vinci Surgical System, it is now a popular robotic system that is utilized for many aspects of surgery and allows for procedures to be less invasive than standard procedures. Recorded malfunctions and failures were decreased after the release of the newest da Vinci model, model Xi. Robotic surgery is assisting patients in gaining their health back in places where more experienced surgeons are not available.

I. History and Background Information

The da Vinci Surgical System’s name stems from the robot that was created by Leonardo da Vinci, hence its name. Leonardo da Vinci was a man of many talents, and due to his interest in human anatomy, he was able to invent the first ever mechanical robot towards the end of the 14th century. The robot’s external design matched a knight and the internals were made of sheaves and cables that allowed the robot's limbs to retract and extract when needed. After many centuries, Leonardo’s inventions were discovered, allowing many revisions and updates to its original design and functionality. The advancements branched out to many fields in science, supporting researchers in gaining information about our galaxy, artificial intelligence, and most importantly, to improve human health. A California company, Intuitive Surgical, Inc., has designed a surgical robot that allows surgeons and doctors to operate on invasive surgeries with fewer complications and better care for patients. The dVSS was FDA approved in the year 2000. “The da Vinci has 2,250+ patents and an additional 1,550 pending. The system is available in 6 continents, 64 countries, and 50 states” (Intuitive Surgical 2017).

The surgical system is capable of operating in many different categories such as gynecologic, urologic, thoracic, neck, head, cardiac, colorectal, and general surgeries. Unfortunately, the surgical system does not complete the required tasks on its own. It is controlled by an onsite/offsite surgeon whilst the patient is monitored by a surgical team onsite in case of data communication failure. The da Vinci Surgical System (dVSS) functions as a tele-manipulator, meaning any motion the surgeon performs with their hand, the robot will replicate. The system works as a “slave” and copies its “master’s” movements. In most cases, the dVSS is utilized at onsite surgeries; however, the dVSS is capable of accomplishing full off-site surgeries. For instance, a patient may be located at a hospital in Houston, and the surgeon could be working at a hospital in Seattle. The dVSS’s new model (Xi) consists of four main components, Surgeon Console, Patient-Side Cart, Endowrist Instruments, and the Vision System.
1. Surgeon Console [Figure A]
   a. “Using the da Vinci Surgical System, the surgeon operates seated comfortably at a console while viewing a high definition, 3D image inside the patient’s body” (Intuitive Surgical 2017).
   b. “The surgeon's fingers grasp the master controls below the display with hands and wrists naturally positioned relative to his or her eyes” (Intuitive Surgical 2017).
   c. “The system seamlessly translates the surgeon's hand, wrist and finger movements into precise, real-time movements of surgical instruments” (Intuitive Surgical 2017).

4. Patient-Side Cart [Figure B]
   a. “The patient-side cart is where the patient is positioned during surgery. It includes either three or four robotic arms that carry out the surgeon's commands” (Intuitive Surgical 2017).
   b. “The patient-side cart is where the patient is positioned during surgery. It includes either three or four robotic arms that carry out the surgeon's commands” (Intuitive Surgical 2017).
   c. “The system requires that every surgical maneuver be under the direct control of the surgeon. The da Vinci Surgical System’s safety checks prevent any independent movement of the instruments or robotic arms” (Intuitive Surgical 2017).

3. Endowrist Instruments [Figure C]
   a. “A full range of Endowrist instruments is available to the surgeon while operating” (Intuitive Surgical 2017).
   b. “The instruments are designed with seven degrees of motion - a range of motion even greater than the human wrist” (Intuitive Surgical 2017).
   c. “Each instrument is designed for a specific task, such as clamping, cutting, coagulating, dissecting, suturing and manipulating tissue” (Intuitive Surgical 2017).
   d. “Quick-release levers speed instrument changes during surgery” (Intuitive Surgical 2017).
4. **Vision System [Figure D]**

   a. “The vision system is equipped with a high-definition, 3D endoscope (flexible tube with a camera and light at the tip) and image processing equipment that provides true-to-life images of the patient’s anatomy” (Intuitive Surgical 2017).

   b. “A view of the operating field is available to the entire OR team on a large viewing monitor (vision cart). This widescreen view provides the surgical assistant at the patient’s side with a broad perspective and visualization of the procedure” (Intuitive Surgical 2017).

II. **Da Vinci in Colorectal Surgery**

   Unfortunately, the dVSS has not been implemented by every hospital in the world yet, making most doctors operate on surgeries manually instead of with the power of the dVSS. Traditional surgery is still a common procedure implemented on many operations; however, the use of dVSS allows surgeons to offer better care for their patients with the precision of its Endowrist Instruments. The precision of the dVSS reaches new heights every day. It has been able to “perform totally robotic-assisted colorectal procedures involving the left colon and the rectum. The procedure is technically feasible and yields short-term outcomes that are similar to those of conventional laparoscopic techniques” (Koh et al. 2010). In the future and with sufficient funding, the dVSS should be available in every operating room as a necessity. A recent study for Colorectal Resection that was conducted in 2016 showed that the dVSS required additional operation time than Laparoscopic Surgery and Open Surgery. Nonetheless, the dVSS had lower estimated blood loss and decreased hospitalization time versus Laparoscopic Surgery and Open Surgery. The meta-analysis consisted of a total ten studies that included 2,767 patients. Furthermore, the dVSS “showed a significantly reduced conversion to open surgery, compared with LS” (Huang J et al. 2016).

III. **Da Vinci vs. Da Vinci S in Laparoscopic Prostatectomy**

   The da Vinci’s launch and first operation was in 2003 and offered surgeons a 3D vision of the internals and carried three arms that offered seven degrees of freedom (Da Vinci Robot 2015). Seven degrees of freedom allow the da Vinci’s arm to move up and down, move forward and backward, right and left, arm yaw, arm pitch, arm roll, and grasp objects. The initial model consisted of only one camera located on the first arm of three. The endoscopic camera has 3D capabilities that allow an enhanced image of the operating field. This model is no longer in production or service. The da Vinci S had launched in 2006 was a revision of the initial model and has an improved 3D camera along with a total of four functioning arms instead of three. This was the first model to include the docking technology, which allowed surgeons to perform surgery at a different location. This new docking technology, if proper satellite communication was available, allowed surgeons to operate on patients across the world. “The connection is based on high-speed optical fibers. Exchange of instruments is handled through an intelligent system guaranteeing quick & secure actions. Real-time communications with remote teams are made possible thanks to its integrated web interface” (Da Vinci Robot 2016). It is also possible
to utilize the dVSS in war zones. With proper connection, surgeons placed in “green zones” or even outside of the country are capable of patching in and assist wounded soldiers.

The study was first published in 2011 at the Ohio State University Medical Center. the study was to identify differences and similarities in patient care and operating room efficiency as well as cost efficiency for hospitals. After a trial of 100 operations, the study was concluded with no significant results, “There was no difference in blood loss (P = 0.08), positive margins (P = 0.87), or mean number of lymph nodes removed (10.7 vs 10.6)” (Shah and Abaza 2011). The only difference the study offered was that the da Vinci model S had a lower operation time than the initial model. The “total operative time was 191 min using the standard da Vinci robot (range 132–266) versus 169 min with S robot (range 98–230), representing a mean difference of 22 min (P = 0.002)” (Shah and Abaza 2011).

IV. Da Vinci Si vs. Da Vinci Xi

The da Vinci Si is the 2nd generation surgical system created by Intuitive Surgical and was released in 2009. The Si has had significant improvements done to its vision and camera modules. A 1080p resolution smart camera is introduced into the Si allowing surgeons and medical staff to capture and view the operation in the utmost quality, providing much more powerful visual quality and improved accuracy. The camera has an “integration of an ultra-light and features equipped with integrated commands granting the operator with fast and precise focusing” (Da Vinci Robot 2016). The Si model also includes built-in scanners for convenience and fast operation. It offers an ultrasound and an electrocardiogram to monitor and record the patient’s heartbeats. The Surgeon Console is equipped with multiple monitors for patient vital tracking as well as different points of view of the operation field. The Si is also equipped with a memory configuration module that saves each surgeon’s preferences for operation, much like the ones available in new vehicles. Much like when a new driver is in the process of controlling a vehicle during driving, the instructor is often equipped with a secondary driving wheel, as well as a brake and accelerator pedal to take control in an emergency. With the new revision and advanced technology built into the Si, a secondary Surgeon Console is made possible. The new Surgeon Console is essential in training new medical staff on the dVSS application. The secondary console is not used in a live operation, but only for training purposes.

The da Vinci Xi is the latest 3rd generation surgical system that was launched in 2014, it is the 4th revision of the dVSS, and as usual, amazing new technology is introduced and implemented. The new Xi offers a high-definition 3D camera that has the ability to increase magnification by 10x more than the human eye is capable of capturing (Core77 2017). All new controls offer enhanced movements and seamless operation with high accuracy and precision. The all new Xi is revamped with a new swivel and a complete movable patient-side cart. With this new feature, it allows the present medical staff to gain additional operating room space by placing the patient bed and patient-side cart in different places of the room thanks to the new movable cart. The dVSS Xi predecessors had an all-in-one patient-side cart that was difficult to move and was usually placed in the center of the operating room, often leaving present staff to maneuver around the system for simple tasks. The Xi software is re-introduced with a simpler graphical user interface (GUI) allowing the present medical staff and operating surgeon to perform complex assignments with ease. The “clean and reduced design language elegantly communicates the use of a highly capable system. The new interface also offers guided setup and
surgical site targeting, to optimize configuration for the specific procedure” (Core77 2017). In addition, the Xi offers an enhanced and improved microphone and a speaker system that allows the operating surgeon to communicate with the on-site medical staff with comfort and ease. The Xi’s operating “arms are thinner and the instruments are longer” (Core77 2017). [Figure B vs E] This new design allows the surgeon to advance within the patient’s body with less maneuvering, internally and externally. The arms are also not required to use the Surgeon Console to (pre-docking into the patient) relocate within the patient’s bed. The arms may be manipulated as easily as grasping the intended arm and moving it around. The arms are also capable of magnetically docking into the patient rather than using a twist/snap method.

V. Training

Virtual reality simulators are not new technology. They have been around for many years educating beginners to excel at the given task such as aiding student drivers and pilots to master their driving/flying skills. Likewise, the most efficient training method that is used to train new doctors, as well as current doctors, is the utilization of a virtual reality simulator. The simulator offers a “real-life” experience that replicates a surgical procedure as closely as possible. The first surgical simulation was nothing but a prototype that was initiated at Indiana University back in 2007. Later on, the prototype had shown positive results at training surgeons by accomplishing certain “games” and exercises that assisted the surgeon in improving their skills and further educating them about the dVSS. The Mimic dV Trainer had boosted the company's reputation, allowing them to become a well-known company in robotic simulation. The company is located and operates in Washington, USA. The Mimic dV Trainer (MdVT) has its own software that controls the simulator. It is “the most realistic and life-like simulation available. MSim is at the heart of Mimic’s mission of helping prepare surgeons to deliver better care. It also has allowed the dV-Trainer to extend the benefits of simulation to now include team training (Xperience Team Trainer) and procedure-specific content (Maestro AR)” (Mimic 2017).

The dVSS also has a built in practice software that consists of two main categories, product and skills training. Each category has two phases that provide student surgeons with the utmost knowledge on how to operate the dVSS. Product training, phase one, introduces the surgical system by providing a library of videos of live and practice procedures to show the machine's capabilities. It also forces beginner surgeons to shadow experienced surgeons during live procedures. Phase one educates the student about the different technology that is included in the full system; for instance, how to operate the Vision System within the operating room. The second category, skills training, tests the student’s skills on the surgical system as they advance. Phase one starts off with the surgeon's first live procedure whilst an experienced surgeon supervises the operation and the student’s usage of the system. After the surgical procedure is complete, the entire present medical staff reviews the operation and studies different methods that could have been implemented and cultivate their performance. Phase two of the official training session qualifies the surgeon to enter a network of surgeons with different specialties that hold their own seminars and continue the education of the student.

A study was conducted on ten resident surgeons and twelve med student trainees using the MdVT and the dVSS built in practice guide. The trainees were placed in the Mimic simulator group and dVSS, whereas the residents were in the dVSS group only. The trainees started off
with a series of five exercises on the dVSS, then four training exercises on the MdVT. The resident completed six training exercises without the use of the Mimic simulator. The data collected was based on timing and accuracy. The results had shown that “training with the MdVT provided similar improvement on five exercises performed on the dVSS when compared with training on the dVSS alone. The use of this simulator in resident and student training may help bridge the gap between the safe acquisition of surgical skills and effective performance during live robot-assisted surgery” (Lerner et al. 2010).

VI. Safety

The da Vinci Surgical System is a complex system that has many essential micro components that may experience many complications that can halt proper system performance. The more micro parts that are included, the higher risk of one of those parts requiring servicing. System malfunction is very likely to occur, which can endanger the patient and the surgeon’s career. Fortunately, Intuitive Surgical is always attempting to lower that risk of malfunction of the dVSS, especially in an operation. However, small cases of system malfunction have still occurred. The dVSS is approved by the U.S. Food and Drug Administration, declaring it safe to use. A database study was conducted on the dVSS by the FDA MAUDE (Manufacturer and User Facility Device Experience). The period of the data collection was from the beginning of 2009 until the end of 2010. The purpose of the study was to assist hospitals and surgical robotic companies to identify, resolve, regulate and to revamp overall system performance. The recorded failures were placed into five categories: cautery, shaft, arm tip, cable, and control housing (Friedman et. al. 2012). The results yielded that “a total of 565 instrument failures were documented through 528 reports” (Friedman et. al. 2012). Seven documented failures were related to the control housing. 29 documented malfunctions were related to cables: 76 shaft malfunctions, 174 cautery malfunctions, and a whopping 285 documented failures of the arm tip. The highest amount of malfunctions originated from the arm tip (Friedman et. al. 2012). Now, this recorded data was recorded in the duration of one year, and 565 instrument failures were documented… The study had also used expired components to collect data. Intuitive Surgical has released a newer model than the one used in this study, in addition, failure and malfunction studies of the company's newest model Xi have decreased, which means the Xi model is safer to operate with, or a study period is used to locate new issues.

VII. Conclusion

All in all, problems experienced and malfunctions experienced with the dVSS are under-documented and often not reported properly. A side device should be created and linked to the dVSS for the sole purpose of system diagnosis, error handling, and error logging. In the case of a complete system malfunction, technical support is capable of accessing the error logging device in an attempt to debug the system and prevent future breakdowns. A device of some sort should also be capable of any external changes to its system; for instance, if a non-certified third party is used to repair the system, certified technicians would be aware of the changes and report back with any incorrect changes. Surgeons and OR staff are not educated on the use of the dVSS in the case of an emergency failure, leading to quick panic in the OR and often having to re-schedule the procedure or reverting to manual surgery. Surgeons and manufacturers should work closer together now more than ever. As technological advancements are made, the possibility of
micro components failure may increase. Improved rules and guidelines should be implemented
in the OR containing the dVSS to minimize surprises and keep an easygoing atmosphere for the
staff to avoid patient injury. Allowing a technical support specialist to be on-call or in the OR is
essential to avoid problems. The rest of the OR staff should also have knowledge of the
equipment that is being utilized. Surgeons should also be trained on time management in case of
a sudden error with the dVSS and converting back to manual surgery for time-sensitive
procedures.

Ultimately, the da Vinci robot has certainly come a very long way in hopes of assisting
humanity to reach new advancements, whether it is in health, aerospace and normal day to day
activities such as driving with artificial intelligence. In addition to all the dVSS’s different
models, no doubt newer models will keep surfacing with improved safety features, and greater
technology such as laser projecting instrument arms that can be used to close any incisions made.
With new advancements, new capabilities will also be made on operating on more regions of the
body. Hopefully, new companies will emerge that will also specialize in robotic surgery that will
boost the market and introduce new ideas and prevent a monopoly from occurring. I think
robotic surgery is the next big thing in medicine that can really push our limits as to what we can
achieve to help each other.
Figures

Figure A) Model Xi Surgeon console - This is where the surgeon is capable of manipulating the Patient-Side Cart. [https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/](https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/)

Figure B) Model Xi Patient-Side Cart – The advanced Endowrist Instruments are located at the tip of each arm and docks into the patient. Controlled by the Surgeon Console. [https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/](https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/)

Figure C) Model Xi Endowrist Instruments - Many of the instruments are swappable and all arms can control the instruments. [https://www.intuitivesurgical.com/company/media/images/davinci_instruments_images.php](https://www.intuitivesurgical.com/company/media/images/davinci_instruments_images.php)
Figure D) Model Xi Vision System - Allows the entire OR staff to view the surgical procedure, it replicates the surgeon’s view on this monitor as well. [https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/](https://www.intuitivesurgical.com/company/media/images/da-vinci-xi/)

Figure E) Model Si Patient-Side Module – 2nd Generation of the patient arm and Endowrist instruments, to be compared with the 3rd generation. [https://www.intuitivesurgical.com/company/media/images/davinci_si_images.php](https://www.intuitivesurgical.com/company/media/images/davinci_si_images.php)
References


Physics in Construction
Brandon Alcayde
PHY 112
Dr. Durandet
Abstract

The purpose of this research paper is to inform people of how physics is included in construction. You use physics to build even things like boats or tables. Although many people think physics is only for physicists, there is plenty of equations to know to help you build the most spectacular buildings. In this paper, I’ll be talking about the different types of structures, which equations and aspects of physics are used in building the intriguing buildings. I will show some pictures of these buildings as well. There will be a paragraph about domes, one about bridges, and I am even going to talk about earthquake proof buildings. There is also construction in building a car and even a boat.

Introduction

There is always something new being built around you. Whether it is an office, a bridge, or even a house, the builders need to know physics in order to build it. There is quite a bunch of physics revolving around construction nowadays because people want their building more spectacular and shaped differently than the rest or the bridge has to be tall and extravagant. What my paper is going to tell you about some of the extraordinary buildings that have been built along with the science behind them. Even talk about how you can make a building stay standing after hit by a massive earthquake. These things are truly remarkable.

Physics Behind Bridges

Now have you ever seen a bridge? Like a not some small bridge but one over a river? There are many types of bridges that are all made up of different materials from rope, chain, and steel. The different types of bridges are suspension, beam, truss, arch, and cantilever. A suspension bridge is known to be some of the most elegant bridges that are so beautiful to look at. A beam type of bridge is one of the simplest bridges to construct because they don’t have to be all pretty like a suspension. They are mainly used for highway bridges. An arch bridge is not as effective as a suspension bridge and it does not span to lengths as some of these bridges. A cantilever bridge is one of the most fascinating of them because it works kind of like a seesaw at the park. It works by supporting the weight at one end with a counter weight at the other end. A truss bridge is a bridge that is designed with triangles and it is used to transfer weight from one side to a much wider side. For example, the Golden Gate Bridge. How did they build a bridge of that size across the bay without it ever collapsing? Well in order for the beams to support such a great weight, they have to be really solid throughout and be really placed into the ground. There is a lot of forces acting on these beams. In order for the builder to figure out the forces or masses that the Bridge has to support, they have to calculate how long they need the bridge to be. That way they can use the mass of the road to drive on so they know the types of forces acting upon the beams so they can determine how many beams they need and how wide the diameter of them have to be. An example of a bridge that the structure was instable and collapsed because they did not take in resistance of the wind was the Tacoma Narrows Bridge. This bridge was nicknamed “galloping Gertie” because of how bad it swayed. If you look at figure 1.2 in the figures section, it is a still motion picture of the bridge as it was swaying. In the picture is the Professor at the University of Washington, Professor Farquharson who was the only one on the bridge when it was swaying. Professor Farquharson was on the bridge because he was conducting studies to
find out how to stop the swaying. There were many of people who were eyewitness to this bridge collapsing and they all said a big chunk of concrete was the first thing to fall which means to me that the concrete was not reinforced with rebar. I think they should have put another beam in the middle where it was swaying and had pieces of concrete falling from it to support it and be able to resist a little of the sway. I also think they should have anchored some steel ropes into the water that hooked up the bridge and this would allow some tension and would resist the bridge from swaying as much as it did. This is one of the most popular bridges that collapsed. Now they have all these simulations they can run on a computer to help them build the bridge so that the structure can withstand any force thrown at it. I believe if they had these simulations and some of the other ways we have now to design things, that bridge would have never collapsed. If they would have thought about some scenarios where they could measure the amount of resistance from the wind or been able to figure out how much tension they could use to anchor the bridge down so it wouldn’t sway, they could have been successful. An equation they could have used to help figure out how much tension they needed to anchor the building down would be \( T = ma \) because \( m \) is equal to mass which could be the mass of the bridge and \( a \) is equal to acceleration which they could use in the case of how fast the wind was blowing on that day.

**Environmentally Friendly Buildings**

There are many buildings now a day that are constructed to save you energy and be more environmentally friendly. These theories can even be incorporated in houses as well. But how do they build a house on mountain without it sliding down? The first thing to take into consideration is the foundation, is it solid or is it a base that could easily make the building slip? If the foundation is made up of rock or sand or something similar, they use the process of post tension. Well what they do is put these tension ropes that hook on to one end of the framework and then go into the mountain and when they are ready, they pour the slab of concrete. When the concrete is almost dry, they pull the tension cords together to create post tension. Think of this process as when you go to build a concrete wall and reinforce it with rebar, the similar thing with the post tension technique in concrete. Now there is so much tension in the cords that if they were to be cut, the whole house would blow up. These building have to be managed properly so that when it is being built the workers don’t cut into the rope and release the pressure or tension in the ropes. Now how do make a house environmentally friendly? Well first what you can do is put a rammed earth wall into the house. You can see figure 2.1 in the figures section. This rammed earth wall acts as insulation in the house so when it is cold, the wall would allow cold air to flow behind it but then when it comes out the other side it will be warm. This will allow you to keep you’re a/c bill down because you do not have to use your heater. In the summer time this works by trapping the warm air and making it cool which allows the house to stay cool during the daytime. This method has been used for many centuries but is not very popular these days. Another cheaper way to make your house cooler is insulation in the attic. There are many renewable energy options that are available Another way that some people power their items is by solar power. This allows stored energy or kinetic energy. Another way to use renewable energy is a windmill. Now if you go up to Flagstaff or head out towards California you see many windmills. Well what they do is harvest energy from the wind but some also have little solar panels at the top to harvest a little more energy to last when it is dark. This can help use solar energy and less of electrical energy first. How do they make buildings such as office buildings more environmentally friendly? Well have you ever been up to NAU? They have one of the only
completely solar powered energy buildings. It is completely designed to reuse the energy that it stores. They also have a building called the Applied Research and Development building that is one of their newer buildings. This building is designed to reduce water by almost 60% which helps save kinetic energy that is stored by the solar panels on the roof and the windmill that they have on the side of the building. Now what kind of science does it take to construct a building in that is dome shaped? Well first engineers have to figure out the diameter of the dome. Now to build these domes, it takes a lot of geometry. These building are probably some of the strongest structural wise because they are reinforced so much so it doesn’t cave in. You need to determine the area you need to fill in with brick by using the equation $A = \pi r^2$. Once you are able to figure that out you can start too close of the dome top. There are many dome like building around the world that have been around for centuries. Which is truly amazing how back when they didn’t have cranes or anything they had to get the brick and concrete blocks up to the top of the building. How did they accomplish this though? Well they created a pulley system that allowed them lift the brick or concrete up to a certain height. That weight must have been a real back breaker then but then if the building was too tall and they were not able to lift the heavy concrete up, they devised a way to where it could be rolled to the top. This method was invented by the Egyptians back when they were building the Pyramids of Giza. The ideas they had to be able to move such heavy rock without a crane was truly remarkable. One very remarkable dome is the Florence Cathedral in Italy. Another dome shaped building that was built here in Arizona is the Skydome up at NAU which is one of the biggest dome like building ever built spanning out to a little over 500 feet. Another one is the Pantheon in Rome, Italy which has a huge oculus or eye at the top that is just about 30 feet around. How were they able to build that building like that? Well what they did was build framework going all the way to the top and then as they were constructing the top of the dome, they had to build a structure like support so that way it wouldn’t collapse in as they were building it. It was truly incredible how they built such masterpieces back before all the vehicles and machines to lift the materials up.

A little town in Germany by the name of Wildpoldsried relies on solar farming. With this technique, they have been able to create about 321% more energy than the small town really needs. This has actually been able to let them start shutting down their nuclear power plants. This has allowed them to expand their solar farming throughout. This whole idea has allowed them to collect almost $6 million annually. Their green initiative started back in 1997 and it has only grown throughout the years. They have four buildings in the community with solar panels and they have had seven windmills built in. They have almost 200 houses built that have solar panels installed on them and this allows them to live strictly off of solar energy. They are hoping their idea can spread to other countries and be able to make the world rely on solar energy. The population consists of only 2,600 people but when they can produce over 321% more energy than they need, they actually sell it back.

The Science Behind Earthquake-Proof Buildings

What I have always found interesting is how engineers make buildings earthquake proof. These are starting to be designed for locations that are on fault lines. These buildings have started to be taken in consideration in Japan after they had a huge earthquake. What causes the most deaths in an earthquake is a building collapsing because it is not structurally built to hold during an earthquake. Since you can’t build a completely earthquake proof building, but there is a way to make a building structurally sound so you it can survive a light or medium earthquake.
but there is no way to know if a building can withstand a very strong earthquake. Engineers of the Earthquake Engineer Research Center have designed a house that can resist the sway of an earthquake. What they do is use sort of a shock like system in the foundation that would absorb any tremors. They also reinforce the buildings with steel beams to resist the sway of the building. They can simulate some scenarios by computer simulation and a big 20x20 piece of ground that they use to simulate certain earthquakes. The idea of inertia or Newton’s first law is huge in the idea of earthquakes because whatever is at rest stays at rest and an object in motion stays in motion with the same velocity until acted upon. Another way to keep a building from collapsing during an earthquake is too make sure it can bear a side to side load as well as an up and down load. An up and down load can be anything from snow on a rooftop to a bed on the second story. Some engineers have found ways to make a building more structurally sound so it can withstand a earthquake. One way is to do a cross-bracing structure which can consist of steel or wood columns. These columns are usually built into the frame that look like an X but the steel is used more in a skyscraper of office building and the wood beams are used more in a urban area to help support the houses from swaying side to side.

Another way to make an earthquake proof building is too make the roof as light as possible. You could incorporate ideas with the one of the Engineers of the Earthquake Research Center by building into the foundation and using these pillars as like a shock absorber and then use the technique of the cross bracing and if you don’t want to lose the flexibility of the house, you can add joints and connectors to help brace it better and keep that flexibility. The shock absorbers would allow the building to keep its normal sway but the cross bracing will keep the building from swaying side to side. There are some buildings though that have been built using these methods that are considered earthquake-proof buildings. One building is in Japan and it is the Yokohama Landmark Tower which is 972 feet tall but the idea they used to build it was a shock absorber system along with flexible materials that are connected with joints and connectors. In Los Angeles on the Pacific plate there is a building called the U.S. bank tower which is a whopping 1,018 ft in the air and with the design the engineers used, it can withstand up to a magnitude 8 earthquake. This is one of the most structurally sound buildings in the world and it is on one of the plates that make up the San Andreas Fault line. Many researchers are coming up with new ways to make a building invincible to earthquakes. Researchers who have tested a 26-foot tall building in Japan have been able to make the building withstand a magnitude 7 earthquake. They used the shock absorber technique and also they used cables kind of like the Post-tension theory that is used to hold a slab of concrete onto the foundation on the side of the mountain. The cables were used to hold the house into place but also to bring the house back together after the shockwaves had stopped.

Physics Behind the Design of Cars

Construction doesn’t always have to mean a building. You can use what you know in construction from the physics aspect to build a car or build a boat. When it comes to cars, there is so many ways you can use physics. For example, if you are building a race car and you want it to be fast, then the weight or mass of the vehicle has to be light. It can’t be too light though because if you have too much torque and horsepower the car will just lift off the ground and you will go flying into the wall. A way to keep the torque under control is by putting a mounts in that would allow the engine to be more controlled and to prevent wheel hops. The biggest problem for engineers when designing a car has always been how do we eliminate drag force. Drag force is
the forces that are opposing a vehicle when it is traveling at a steady or accelerating velocity. This is why when you see a drag race or a Nascar race the cars usually can never reach over 200 mph. When you are driving in a car that can exceed 150 mph like for example a Lamborghini Aventador, and you are driving on the Autobahn and you decide to open the throttle and push it to 150 mph or 240 km/h and you try to lift yourself forward to change the station on the radio, you just simply cannot move because all the drag force that is pushing you back. The Drag formula is \( D = \frac{1}{2} C \rho A v^2 \) and this helps designers be able to see the drag force on their vehicle. This resistance on the vehicle can also be a factor in determining the miles per gallon the vehicle gets at top speeds. No matter how strong you are, that force will always keep you back and when you go faster, the more force being pushed against you. When building a car, determining how big the engine is you want in it and what you can build it out of will determine how lightweight it is. When a car is going in a straight line like on a drag strip how do you measure the power of the vehicle? Well you can use the formula \( P = FV \) where \( P \) is equal to power, \( F \) is equal to force and \( V \) is equal to Velocity. Since this is a measure of power you would need to use either joules per second or watts. You can also measure the force the engine is putting out at a certain speed by using the formula \( F = ma \). So say the car has a mass of 4,000lbs and is accelerating at 75mph how much force is the engine putting out. Well if you multiply 4,000 by 75 you get the answer of 300,000lbs of force. This is an insane amount of force. Obviously not all cars are going to be designed like this. In order for the car to be fast it also has to be aerodynamic which allows the air to blow over the car and causes less resistance that opposes the vehicle. Now a vehicle like a pick-up truck or an SUV will be as aerodynamic as a car like a Lamborghini or a Mustang.

**Boat Architecture**

How do you use construction to build a boat? Well just like building a house you must first construct a frame to in order to lay the wood around to make the hull. This will determine what the shape of your boat will be. Just like a house, you need a frame that lies underneath the outside material to hold it in to place. When you start to build a cabin you would put the wood up using nails and screws and these would attach to the frame of the house. Well with a boat you would use the same idea but it is now called Carvel Planking. This is where the wood would butt up against each other around the frame of the boat. Now you can build a boat out of almost anything from wood to steel. Steel you would see more on the bigger boats such as yachts and cruise ships and wood would be used on little boats fishing boats. You can see some lightweight 3-person john boats made out of aluminum for you and your friends when you want to go fishing. Although the aluminum isn’t as tough as the steel it is a good and cheaper alternative. How does a boat stay buoyant? Well to be buoyant means there is a upward force that is exerted on an object is in a fluid. A principle that can be used to talk about boats it Archimedes principle which is the amount of fluid that is displaced relates to the weight. How does a boat float, even the ones made out of metal? In order for the boat to float, it has to be able to displace the water equal to the weight of the boat. A solid boat would have a density that is so much it would sink because the downward force would be more than the upward force exerted. While a hollow boat has a lower downward force than the upward force exerted by the water which allows the boat to stay afloat. The reason a metal boat that isn’t solid throughout can float is because the hull consists of air which allows the density to still stay low and allow the downward force to be less than or equal to that of water.
Conclusion

There are many aspects of building a bridge in the and how to get one to staying up for so long. I have talked about one of the main bridges that was supposed to be the greatest bridge ever built but soon collapsed after being open. I went into how I would have personally built the bridge that would allow it to have stayed standing after that powerful gust of wind hit it. I had also gone into the many different kinds of bridges there are and how each one is built. In the next paragraph I talked about environmentally friendly buildings and houses. I also talked about a small German town that has allowed themselves to create over 300% more energy than they really need to live on a day to day bases. They have been doing this since 1997 and they have only expanded their buildings and resources. What they have been able to accomplish in such a small town is truly amazing and I really hope the world can learn from what they do because it could really help in the long run. I also talked about in the environmentally friendly section on how here in Arizona, specifically up north a Northern Arizona University where they are trying to be more environmentally friendly by building windmills and having solar panels on the buildings. They even have one of the few completely energy sustained buildings. Their Applied Research and Development building was built to be resourceful and help figure out ways to be more environmentally friendly and this has allowed them to reduce their water usage by almost 60% by their irrigation setup.

I then went ahead and talked about earthquake-proof buildings. These buildings are absolutely amazing and if they can be perfected they could save a lot of lives in the effect of an earthquake. The main reason people are injured or even killed in an earthquake is because the house collapses. This is because the building isn’t designed to withstand such a force that an earthquake can produce. There really is no way the magnitude an earthquake would be until it hits at its peak so being able to design a building to withstand it makes it a little hard. I came up with a way that I would construct one of the buildings. I also talked about the different ways scientist have used to test their buildings to see how much of a magnitude they can withstand. I also talked about how you can use physics to build a car. I talked about drag force which is the forces opposing the car that is accelerating down the street. I gave the formula for Drag force which is \( F = \frac{1}{2} C r a v^2 \). I talked about how the mass and acceleration of the vehicle can really effect the force the car is putting out. We can find that out by the formula \( F = ma \) which mass times acceleration which is equal to force. I talked about how building the hull of a boat is similar to building a house and how you need to build a frame before you can start to put together the outside. I also learned this while writing the paper is that when a steel boat is floating on the ocean, it is floating because the hull is air. If the hull would have been a solid structure, it would sink. When a boat floats it must displace the weight of water related to the weight of the boat. This is called Archimedes’ Principle.

There are many news ways that is evolving construction to be able to build buildings that can withstand some natural disasters. I know I learned a lot from just the research I was able to do and it helped me to expand my knowledge on a lot of these topics. I never really knew how much physics was used in construction or even to build a boat. Usually when I was on a jobsite I just went out and worked and didn’t think about the physics aspect as much but looking back at it now, there was a lot. I definitely thought the Earthquake proof buildings was the most interesting part because I didn’t know much on the topic. So to be able to read about it and be able to incorporate that and really think about how it could be used in a community on the San Andreas fault is spectacular. These earthquake proof buildings really can save lives and save a lot of money in having to rebuild after a natural disaster.
Figures

Figure 1.2

Figure 2.1
Reference:


The Potential Hazards of Talcum Powder and Its Uses

Ariel Alcocer

April 20, 2017

General Chemistry II, CHM152

Professor Julie Olander
Abstract:
This paper will be discussing the different uses and potential dangers of a talcum powder. Talcum powder is used in a variety of different products shifting widely from cosmetic purposes to industrial. This is all due to the incredibly versatility that talcum powder provides for manufacturers and consumers alike. However, there are a number of dangers that talc may present such as lung disease, lung cancer, ovarian cancer, inflammation, and even aspiration. Throughout, talcum powder will be discussed as well as a variety of studies and reports either supporting or negating multiple claims regarding the potential health hazards of talcum powder.

The Potential Hazards of Talcum Powder and Its Uses
Throughout history talcum powder has been a common household product. Dating back to the ancient Egyptians uses as a cosmetic to today as a way to keep ourselves dry. Talcum powder has definitely become known as a versatile product. But what about talc makes it so useful? What is the chemical makeup and its properties? Throughout this paper the uses of talcum powder, its chemistry, products containing it, health concerns, and its regulation will be addressed.

What is talcum powder anyway? Where does it come from? When discussing the potential harms of talcum powder, one must first understand the chemical makeup of the substance and where it comes from. Talc is a hydrous magnesium silicate mineral with a chemical composition of silicon, magnesium and oxygen. \[2\text{Mg}_3\text{Si}_2\text{O}_5(\text{OH})_4 + 3\text{CO}_2 \rightarrow \text{Mg}_3\text{Si}_4\text{O}_{10}(\text{OH})_2 + 3\text{MgCO}_3 + 3\text{H}_2\text{O}\.] This is the primary formation of talcum powder. Talc can also be written as \(\text{Mg}_3\text{Si}_4\text{O}_{10}(\text{OH})_2\). (The Molecular structure of talc can be seen in figure one of the figures section provided.) Talc is characterized by water molecules trapped between silicate sheets, which belongs to the silicate subclass phyllosilicate and the clay group montmorillonite/smectite(Muscat JE, Huncharek MS). Cosmetic talcum powder contains greater than 95%–99% pure talc, while other dusting powders are typically composed of talc, cornstarch and other additives(Muscat JE, Huncharek MS). The mineral largely comes from China, South Korea, India, United States, Finland, Brazil, France, and Japan. However, “The United States is self-sufficient for most types of talc used in manufacturing. Estimated 2011 production was 615,000 metric tons”(Talc: The Softest Mineral). When the talc is mined, it is not mined as a pure talcum rock. When the rock is in its natural state after mining, some of it may contain asbestos. Asbestos is commonly known as a carcinogen that primarily attacks the respiratory system upon inhalation. Talc and asbestos are both silicate minerals. “All talcum products used in homes in the United States have been asbestos-free since the 1970s”(Talcum Powder and Cancer). However, the material is still not one hundred percent free of potential asbestos. There are still trace amounts of the asbestos within the powder once it has been deemed as pure talcum powder. As will be discussed later, this is why many individuals are lead to believe that prolonged exposure over time may lead an individual to asbestos exposure.

What cosmetics and personal hygiene products contain talcum powder? When talcum powder is the topic of conversation most people only think of baby powder and various products related to baby powder. However, talcum powder is not limited to just baby powders. Due to to the versatility of talcum powder is such a widely used substance because of its smooth texture, resistance to heat, lubricating abilities, as well as its high absorbance ability. It can be found in
numerous cosmetics, facial creams and masks, and facial powders. In facial powders talc creates a smooth surface as was helping to absorb any excess moisture that may accrue on the surface of the skin throughout the day. Since the particles of the talc powder are an incredibly smooth, and soft material the feel of the cosmetic itself is improved. Preventing a harsh, irritating feel to the skin. The powder provides the same desired smoothness and absorbance in creams as well as shower and body products. Often times, when this powder is being used as a feminine hygiene product, it will be applied to the inside of womens underwear to prevent chaffing throughout the day. When used industrially, talc is added to plastics, ceramics, paints, and even papers. Often times the addition is made because it adds a rigidity to whatever material it will be added to, as well as decreasing the adhesive properties. Also, because talc has a very low hardness, it aids in preventing abrasion of machinery during production. In paints, it help to dry the paint after application as well as preventing the product from dripping off of the piece it is on. For paper materials talcum is used as a filler. Because paper is mostly comprised of fibrous materials there will be small spaces between strands of fiber. The t alc mineral is added to fill these spaces as well as providing a bright white and opaque color. These qualities numerous make talc an ideal additive to so many different products.

Is talcum powder safe? If not, why is it still being sold? The most important question is whether or not talcum powder is safe for consumer use. The main health and safety concerns are that talcum powder may increase a woman's risk to developing ovarian cancer, aspiration caused by respiratory distress, and even lung cancer. Depending on where research is conducted the answers can be both yes, it it dangerous, and no, it is not dangerous. It also depends on whether or not the research was funded by a corporation or if it was an independent study. For example, one such claim that talcum powder and its uses are not a reason for concern comes from information provided to consumers via Johnson and Johnson. The fact that it comes from a company that produces and manufactures baby powder, which is talcum powder, would lead an individual to question the integrity of the research and/or information provided. With that being said, this information may contrary to the truth. The potential for developing ovarian cancer was the main topic of concern on Johnson and Johnson's page. This is due to a case that had recently come into the public eye regarding the powder(The case will be discussed below). According to the site, it claims that “Among the many studies that have confirmed the safety of talcum powder use are two major prospective cohort studies that included more than 130,000 women and were run for more than 14 years”(The Facts on Talcum). Of these studies, data showed no increased risk of ovarian cancer in women who used talcum powder, regardless of the type of use (The Facts on Talcum). On the other hand, according to Consumer Safety claim that research has shown, “an observational study of 18,384 women found that the use of genital powder is tied to “a 20–30% increase in risk of developing epithelial ovarian cancer.”(Ovarian Cancer and Talc). (This information can also be seen in figure two of the figures section.) Consumer Safety also states that the evidence of such dates back to the 1960’s. It is believed that the potential for trace amounts of asbestos is the underlying cause of the increased potential to develop ovarian cancer.

Recently, a case concerning Johnson and Johnson has become a common topic of discussion. The case states that the baby powder created by the company was a cancer causing agent when used near and around the genitalia. The cancer that it is believed to be causing is ovarian cancer. Although it may seem like the concept is far fetched, it was all too real for a woman named Jacqueline Fox. Upon visiting the doctor she was diagnosed with advanced stages of ovarian cancer. “She had chemotherapy to shrink the tumors and surgery to remove her uterus,
ovaries, fallopian tubes, and part of her spleen and colon”(Berfield S, et al). Before Fox passed, she had seen that Johnson and Johnson's was being sued for knowing that there is a connection between long term use of the baby powder and ovarian cancer. It is believed that, over time, talc particles would find their way through vagina, the uterus, then to the fallopian tube, and finally to the ovaries. This contact with the ovaries would then cause inflammation which would then eventually lead to cancerous tissue development. Once again, it is believed that the trace amounts of carcinogenic asbestos are the reason for this. Johnson and Johnson claims that, “After 30 years of studies by medical experts around the world, science, research and clinical evidence continues to support the safety of cosmetic talc. Two widely-accepted, very large studies which followed women over a period of time — the Nurses’ Health Study by the Harvard School of Public Health published in 2009 and the Women’s Health Initiative Observational Cohort by the U.S. National Institutes of Health published in 2014 – found no association between talc and ovarian cancer”(A Message About Talc). However, not all studies agree with the studies conducted by Johnson and Johnson. According to the, “International Agency on Research on Cancer states that using talcum powder around the genital areas increases risk of cancer in humans as studies indicate traces of talc in both the ovaries and pelvic lymph nodes”(Product Liability Claims). This would be enough to support the notion that the talcum powder possesses the capability to travel up through the reproductive system and cause harm. Unfortunately, there is not enough information to truly know and understand whether or not there is a distinct link between the long-term use of the baby powder and ovarian cancer. There have simply not been enough studies to investigative side with one or the other. However, at the time of the lawsuit, a lawyer representing Johnson & Johnson made a statement saying that company had been aware of the association of talc and ovarian cancer but did not think that it was significant enough to add a warning label to their products.

Not only is talcum powder thought increase the risk of ovarian cancer, but it is also the root cause of aspiration and respiratory distress. Unfortunately, the group of individuals that are primarily affected by such are infants. Talcum powder, baby powder, is incorrectly used in the care taking of numerous children and infants. Often times during a diaper change parents will utilize baby powder to powder around where the diaper will be placed on the infant's. This is done to prevent chaffing. However, because talcum powder is so light and easily dispersed, often times the powder will fill the air. Inhalation of the powder will follow suit. Here is where the issue arises. Talcum powder is extremely dangerous upon inhalation. Talcum powder is exceptional at absorbing excess liquid. In a wet environment, such as the lungs, the talcum powder will begin to absorb the liquid. In this case, begin to destroy and dry out the surface of the mucous membranes of the tracheobronchial tree. This will then prevent the cilia of the bronchi to perform the task of removing any impurities from the lungs. Impurities such as the talcum powder(Matina F, et al). When this occurs, many individual will begin to cough, have incredible difficulty breathing, and in extreme cases aspiration and alveolar damage. One such case is as follows. An eighteen month old was admitted to the hospital following inhalation of talc powder. The individual was apyretic some distress to the respiratory system. Upon examination the individual respiratory sounds were relatively normal on both the left and right lung. However, upon running a chest x-ray, it was revealed that a consolidation or mass of powder in central and middle lung zones. The patient was administered antibiotic and steroids over the course of a five day span. “A topical treatment by aerosol with surfactant (Poractant Alfa), budesonide and salbutamol was added. Surfactant was given and it was repeated after 12
hours; after then the aerosol therapy was administered only with budesonide (0,25 mg/dose) and salbutamol (2 mg/dose). Inhalation therapy was carried out with a pneumatic nebulizer and with a spacer chamber, with aiming to reduce oropharyngeal deposition and increasing distal bronchial deposition. A respiratory physiotherapy by clapping was performed twice a day for a week. Chest X-ray showed a clear improvement of the lung condition after five days of therapy, showing only an increased periheral interstitial marking on both sides”(Matina F, et al).

Following all of the procedures the patient fully recovered. The recovery is extensive and very demanding of the patients. Especially to younger children and infants to already possess weaker immune systems in comparison to adults. With everything being considered, it is incredibly clear that talcum powder is without a doubt unsafe to use around infants.

In addition to posing a threat to infants, individual who work as miners and millers where the talc is deposited are at an increased risk of developing lung cancer or even respiratory disease. It is known that talcum powder easily finds its way into the air. In the mines, where individuals are unearthing the mineral, workers are exposed to incredible amounts of talcum dust. Unfortunately, these miners and millers are not only exposed to talc dust but unknown amounts of asbestos as well. As mentioned above, asbestos and talc occur together naturally. Which is why the talcum powder products created for consumers are asbestos free. Because the talc is not free of impurities it is known as talc ore. For the workers however, the talc has not yet been purified. Meaning the potential for exposure to asbestos is incredibly high. Taking a look at the mortality rate among the workers of a mining facility it is clear that an increased risk for lung disease is apparent. For example, a study was done on a group of 1,795 men who had at least worked for a year in a talc mine. Data from this group showcased that there was, “... a significant excess mortality from non-neoplastic respiratory diseases [at a 95%]”(Coggiola M, et al). Non-neoplastic respiratory disease covers a variety of pathologic disorders such as asthma, interstitial lung disease, pulmonary hypertension. A similar research study was done on a group of men who had lived during and after 1950. Of these men, “Employees with high, compared with low, estimated exposure to dust had a rate ratio of 0.5 (CI = 0.2-1.3) for lung cancer and of 11.8 (CI = 3.1-44.9) for pulmonary fibrosis”(Honda Y, et al). Again, results from this study show that although the individual may not have been at a substantial increased risk of developing lung cancer, the vast majority of them would be most at risk of developing a lung disease.

Taking all of this into consideration, why is talcum powder still on the market and considered safe for the consumers? The main reason for this is simply due to the fact that there is not enough substantial research and evidence to support either claim. In order for anything steps to truly be taken to prevent any possible harm caused by the material more studies must be done. Studies on larger scales and with much more rich data collection. Also, the research that is currently accessible has substantial amounts of bias potential. This fact can be seen in the entirety of research and the lawsuit concerning Johnson and Johnson and johnson. So if research and case studies are biased, why isn't the FDA getting involved. Here is why. Talc is considered a cosmetic ingredient. Cosmetic is defined as (1) articles intended to be rubbed, poured, sprinkled, or sprayed on, introduced into, or otherwise applied to the human body or any part thereof for cleansing, beautifying, promoting attractiveness, or altering the appearance, and (2) articles intended for use as a component of any such articles; except that such term shall not include soap according to the Cornell University Law School. Due to this, talc regulation is placed under the FD&C Act. This Act stands for Federal Food, Drug and Cosmetic Act. The Act
is a set of laws that gives the FDA the authority to the U.S. Food and Drug Administration (FDA) to oversee the safety of food, drugs, and cosmetics as appointed by congress. With that being said, not all of the cosmetics that are on the market are necessarily FDA approved but they are FDA regulated to an extent. As stated on the FDA’s website, “The law does not require cosmetic products and ingredients, other than color additives, to have FDA approval before they go on the market…”(Center for Food Safety and Applied Nutrition). However, “FDA follows the procedures required by the Administrative Procedure Act, another federal law, to issue FDA regulations. This typically involves a process known as "notice and comment rulemaking" that allows for public input on a proposed regulation before FDA issues a final regulation. FDA regulations are also federal laws, but they are not part of the FD&C Act”(Commissioner). This means that if the public was to bring to the FDA’s attention that there is a true health concern or threat to the public that is caused by talcum powder that they would have to review it. If that were to happen then the product, whatever it may be, would potentially be taken off of the market.

In conclusion, it is clear to see the numerous uses and benefits of talcum powder in our everyday lives. Without it we many of the products we use on a day to day basis would not work quite as well. If we were to totally omit talcum powder from manufacturing we would need to find a completely different material to add to our products. As mentioned before, talcum powder helps so many of our products and the efficiency in which they work. So far to my knowledge, we have not found any safe material that we could use to do just as much as talc powder does. Let alone perform to the same standards that talc provides. Unless another material was found to replace talc, specifically industrial, get enough support to eliminate it in my opinion.

After researching as much as I did for this paper I decided that it would be interesting to make a personal investigation into my own home. My goal was to find as many products that contained talcum powder as I could to see just how involved in my life the mineral was. The most obvious and easy to find was the baby powder. Like so many other individuals in the world, my mother and father kept this in their bathrooms as a personal hygiene routine. They utilized this as an antiperspirant and as a way to prevent chaffing. I also found the mineral in a number of my cosmetic products. Over the years my cosmetic collection has become pretty large. So it was interesting for me to see just how many of them were made using talc as well as the brands that used them the most. Unsurprisingly, I found it in all of my eyeshadows and my facial powders. I was surprised, however, to find it in one of my lipsticks. This lipstick was meant to be a matte lipstick. Meaning that as the lipstick would dry it would lose its glossy sheen. I believe this effect is created because of the talcum powder in the product. I came to this conclusion because of talcum powders drying ability. Further investigation lead me to find talcum powder to be used in many of my soaps, crayons as a building agent, chewing gum, deodorants and antiperspirants, and even light coatings on some tablets. The coating on the chewing gum and the tablets were to prevent sticking and caking of the products. Again, preventing any sort of moisture buildup.

After taking all of the information that I have found into consideration, here is what I think need to happen. It is clear to see that although talcum powder is incredibly useful in so many different industries because of its numerous capabilities, it is also very capable causing harm to many individual. Some of these individual being infants who cannot help themselves.
And this fault does not necessarily fall on to the parents for the easy shift of responsibility of saying, “the parents should know better.” In the case of diaper changes and talcum, baby powder, inhalation by children the public needs to be better educated. Honestly, until I had started doing my research on this topic I had no idea that there were such things as talcum powder poisoning, respiratory diseases caused by its inhalation, and even ovarian cancer. I believe this is due to companies such as Johnson and Johnson trying to cover up the truth about such instances and cases. The public and consumers deserve to know the truth about what they are using. Is this safe for me to use? What makes it unsafe? Is this safe to use around my children? I think that the purification of talcum powder needs another review to better eliminate all asbestos. Although it is said that all of the carcinogenic mineral has been removed. It simply does not make sense that so many women who had used the powder over the course of many years develop ovarian cancer. There has to be an underlying link to cancer associated with the powder. However, if the powder were to be further rid of the asbestos, then the powder itself would not be prohibited. Additionally, for the safety of the consumer and the safety of the public, the health risks associated with talcum need to be brought to the FDA’s attention. As mentioned before, the public has the power to get the FDA to review the material and its uses. Doing this would also force so many different companies to take a closer look at the products they are producing, how they are producing them, and review their potential dangers and hazards. As a result, children would not be at so great a risk of talcum inhalation, fewer women would be at risk for ovarian cancer, and those individuals working in the mines would be better protected.
Figures

**Figure 1: Talcum molecule**
Here the different elements that makeup the mineral talcum can be seen; Magnesium, Silicon, and Oxygen.


**Figure 2: Informative Poster**
This interesting and informative poster on the potential of talcum powder to cause ovarian cancer it really interesting. It also provides a visual for many individual to pull relive information from. It would also be a great informative flier to have in place like the gynecologist offices.

References


Using Chemistry to Explain Love

Moshe Alishayev

April 20, 2017

General Chemistry II, CHM152

Professor Olander
ABSTRACT:
Love is one of the most complex concepts to understand, not only can it be a concoction of emotions but there is also a large biological factor that aids in the complexity of the situation. Love, or C8H11NO2+C10H12N2O+C43H66N12O12S2, is comprised of dopamine, serotonin and oxytocin. However, other hormones also play a role in love like norepinephrine, pheromones, phenylethylamine, vasopressin, endorphins and more. Lust, attraction and attachment are the three systems/phases of love that aid in the neurochemical process contributing to the feeling love or the desire to mate. This paper will discuss the chemistry behind what is known as “love”.

Love starts at with lust, initiated by physical attraction and flirtation. It is the urge to have sexual desires. In men, the hormone that causes these urges and desires to partake in sexual activity is testosterone, while estrogen is produced in women. Both men and women can lack testosterone and/or estrogen, which can be treated with medications like Androgel (testosterone gel), testosterone injection, estradiol injections/patches and progesterone. These medications increase their sex drive, but they do not fall in love. Sexual desire can be produced with no need or feeling of attachment. In modern day, this can be described as “friends with benefits” where couples relieve their sexual desires without any romantic or deep attachment.

Humans (along with most mammals and insects) produce pheromones, much like an aura, that is released by an organism to attract an individual of the opposite sex. Pheromones are somewhat like a scent that is picked up, either attractive or defensive. It acts outside of the body to influence the behavior of another individual as can be seen in figure A. Wild animals mark their territory with pheromones within their own bodily fluids. Colony insects rely heavily on pheromones not just for mating purposes but also communicating and keeping the colony safe. Scent plays a role in detecting pheromones because one is more attracted to something that smells pleasant and releases a pleasant odor rather than a rancid odor.

Claus Wedekind, a Swiss zoologist, conducted an experiment known as the “Sweaty T-Shirt Experiment” where over 40 men and women were sampled for their MHC gene types, which are the proteins that help T cells determine self from non-self for an immune response. Next, the men in the experiment were asked to wear the same t-shirt to bed for a few nights without any cologne or deodorant. The shirts were then placed in “sniff boxes” with “smelling holes” where the female participants were asked to rate the scent of the shirts based on intensity, pleasantness and sexiness. The female participants were more attracted to the males’ scents with more genetic compatibility to themselves who were the males with different MHC gene types. This is a biological and evolutionary survival mechanism because their potential offspring would be more protected against different viruses.

Although lust and attraction share similar connotations, they differ from one another greatly in part because attraction is associated with focusing our attention to a special person or desire where as lust is the urge for sexual desire. The difference between the two phases stem from a mix of neurotransmitters and hormones in attraction where as lust is a
rage of hormones. In humans, attraction is combined with feelings and exhilaration, constantly thinking about a beloved one and wanting emotional support. Passionate, obsessive love, and infatuation all describes attraction.

Romantic or passionate love is characterized by euphoria when things are going really well and terrible mood swings when they are not. When one falls in love, he or she have many different physical symptoms. The symptoms are as follows; loss of appetite, lack of sleep, no focus or concentration. The reason for these symptoms is due to various types of monoamines, or surging brain chemicals. Various neurotransmitters are responsible for sending certain messages to the brain causing us to feel or think a certain way and these same neurotransmitters are behind why it’s possible to feel love or even fall in love.

Dopamine is a neurotransmitter that is associated with the pleasure system of the brain. It provides feelings of enjoyment and reinforcement to motivate us to do certain activities. Dopamine is released by naturally rewarding experiences such as sex or food. Phenylethylamine is a natural amphetamine like the drug Adderall. It causes the same stimulation effects. The drug causes one to have that on-top-of-the-world feeling that attraction can bring. Phenylethylamine (PEA) is addictive yet the rush or the high does not last forever. The body will only produce so much PEA despite the body’s addiction and craving for higher levels. Both, dopamine and phenylethylamine are easy for the body to become addicted to and continue a certain behavior.

Serotonin is another neurotransmitter. It controls impulses, obsessive behavior, and the sense of “being in control”. Studies show that people with OCD have very low serotonin. So when people with OCD fall in love they are overly attached to their loved ones because their serotonin levels are even lower. Norepinephrine is another neurotransmitter that is mentioned when speaking about attachment. Norepinephrine induces euphoria in the brain and it excites the body by giving it a booster dose of natural adrenaline. This may cause the heart to beat faster and blood pressure to rise and can therefore experience pounding heart or sweaty palms when one is around someone they are attracted to.

That is all just the starter pack for love. The rage of hormones and chemicals in the brain do not necessarily mean that one is in love. Rather, one is “in love” with the way the other person makes them feel and how they feel being with the other person. The “love” is more a selfish kind where one does it simply for the pleasure they continue to get out of it. Romance can quickly fade from this stage if the relationship starts falling apart. However, not all relationships end at attraction.

The purpose of mating is not just for the pleasure but also to survive and reproduce. That being said, it all ties together that there should be a sequence of events starting from your brain subconsciously picking the best mates via sensing pheromones and wanting to constantly be together and mate because it not only feels good but also helps with breeding. Humans, however are special in the fact that monogamy is a value that is highly cared for and that comes with attachment, which is the final step in love. Attachment is loving someone for who they are and having a sense of stability in a long-term relationship.

Oxytocin is known for being the “cuddle chemical” that acts like a neurotransmitter in
the brain. It is well known in female reproduction process that is released during
stimulation of the nipples and during labor. It is generally responsible for a sexual
arousal, facilitates reproduction, and affects maternal behavior; oxytocin, according to
recent studies, increases trust and reduces fear, and even affects generosity level. These
same maternal instincts allow for a closer bond and attachment.

Vasopressin is known as the “monogamy chemical” that is released after sex. Although with
today’s societal norms, one would assume humans should be monogamous but they are not.
Rough three percent of mammals are monogamous, not including humans. In a study done on
monogamous prairie voles and non-monogamous laboratory mice, when the mice were injected
with vasopressin their standard frivolous. Sexual behavior began to change where they had
preference over one partner.

Often after intercourse, it is common to have feeling of closeness due to increase of vasopressin
in men and oxytocin in women, after having achieved orgasm by heightening the male’s sense
of responsibility and giving them protective feeling over their partners. Females however have
an increase in their “maternal” instincts and continue to grow attached to their partner.
Removal of either oxytocin or vasopressin leads to failure in being able to attach with someone,
which can lead to unsteady relationship hopping or infidelity. However, much like in the prairie
vole case, vasopressin injections can help aid in solving the issue. The only problem is it is not a
“cheater’s cure all” and ethically, it cannot be forcibly administered.

Love is complicated, it can come and go. A reason why love is so complicated it because the
high levels of oxytocin and vasopressin interfere with norepinephrine and the dopamine
pathways. One can be attracted to one person, be attached to another person and have sexual
relationships with another person. Monogamy is not promised which can really tear a couple
apart.

The longer a couple is together the more their passion fades. A lot of the times it is very hard for
two different characteristics/personalities to get along. There are a lot of people around the
world that get married, and many of those marriages end up in divorces. When getting married
many do not understand what marriage really is. The divorce rate in the United States is really
high and it is still increasing. The divorce rate is expected to reach 67 percent in the united
states. It happens to be that 80 percent of divorced men and 62 percent of divorced women get
remarried again, and then 54 percent of those men and 61 percent of those women get divorced
again. Love is interrelated but also independent.

The worst thing that has to do with dating, relationships, or marriage is rejection. Getting
rejected by a loved one can destroy someone. Rejection can lead to clinical depression, suicidal
behavior, jealousy and a lot more. In the United States at least 25 percent of homicides involve
spouses, sexual partners, or sexual rivals. Research shows that there are 1 million American
women are followed and harassed by rejected lovers each year.

Studies show that roughly 370,000 men are stalked by former partners and around 1.8 million
wives in the United States are beaten by their husbands. This does not only happen here in the
United States. Unfortunately, these incidents happen all around the world in cultures
worldwide. Such incidents are constantly written in the newspapers, showed on TV news channel and aired on radio stations. Male sexual jealousy is a leading cause in spousal abuse/domestic violence. Both men and women, globally, can experience clinical depression when a relationship fails. Psychologists say that a high percentage of people who commit suicide is because they have been rejected by someone they love.

Love is all about chemistry- the chemistry between two individuals and the brain chemistry within the individuals. Much of it stems from a biological factor allowing the breakdown of love to be noted in phases. The desire to mate is purely an evolutionary trait however, with everything that goes on inside the mind to make “love” possible, it is far more than just selectivity in order to survive. The hormones and neurotransmitters individually have their own purpose but together, make it possible to feel one of the greatest yet most confusing emotions known to man.

When falling in love, one would never think that all this is going on in his or her head and body. Yes, an individual will feel lust, attraction and even attachment but he or she does not know what is behind it all making it work. When writing this paper I constantly looked back at my own relationships in the past and present and am amazed at how this all works. Each of us were given life on this earth and we should all take advantage of the beauty of it.
Figure 1: Diagram of the phases of love and their chemical basis. The hormones involved in lust begin in the body sending messages to the brain because of the pheromones that were sensed. As a reaction, causing the individual to become attracted and have a physical effect on the body like loss of sleep and appetite. Later allowing a release of oxytocin and vasopressin from the brain to fulfill attachment.

References


The Physics Involved in Rotator Cuff Tears

Dania L. Amaya Zapata

November 17, 2016

Physics 112 Course #26104/26171

Professor- Dr. Casey Durandet
Abstract

The rotator cuff is a center point for using the arm for lift, holding, and for doing simple everyday activities. Major injuries may lead to a rotator cuff tear in which may require surgical repair, depending on the severity. The muscles and bones of the rotator cuff work well with each other to exert various range of movement. Newton’s third law of motion, torque, and work force are found in the use of the rotator cuff. They also have a connection in the process of repairing and rehabilitating a rotator cuff tear.

Anatomy

An essential part of the human anatomy is the rotator cuff which consist of four muscles: the subscapularis, the teres minor, the infraspinatus, and the supraspinatus. There are three bones that form the shoulder: the humerus, the scapula, and the clavicle. Together, they form a cuff over the head of the humerus (upper end of the arm) and the scapula. The four muscles form a tendon unit in which helps maintain the arm in the shoulder socket and covers the head of the humerus. This gives the ability to move, lift, and rotate the arm internally and externally. The rotator cuff also helps keep the head of the humerus in the glenoid cavity of the scapula as well as to help avoid impingement of the shoulder (when the humerus is pulled into the subacromial space by the deltoid muscle) by depressing the humerus. The glenohumeral joint is a ball and socket joint in the shoulder between the head of the humerus and the glenoid cavity of the scapula. This joint allows the arm to move in flexion, extension, abduction, and adduction, as well as in a circular rotation. The acromion of the scapula connects to the clavicle. Between the rotator cuff tendons and the acromion is a fluid-filled sac that reduces the rubbing/grinding of the rotator cuff and the acromion, known as the subacromial bursa. The bursa (a lubricating sac) gives the rotator cuff tendons the ability to move freely when using the arm. A joint capsule, an airtight sac with the walls made of ligaments, surrounds the glenohumeral joint to allow the shoulder to move through its abundant range of motion with no restrictions.

The muscles of the rotator cuff are voluntary striated muscles. They are composed of molecular proteins, myosin and actin, in which combine to form filaments. When the muscle contracts, there is an electric force of attraction by the filaments collapsing together about 15-20% of its resting diameter. Because muscles are only able to pull and not push, the muscles on the body are arranged in opposite directions towards each other to give the limbs movement through multiple directions the ability to move back to their opposite directions.

Movement

The infraspinatus and teres minor muscles are responsible for the external rotation of the shoulder, however, the teres minor is mostly operating when the external rotation is abducted 90 degrees, whereas the infraspinatus mostly operates in a neutral position of the arm. The subscapularis muscle supplies about 53% of the total strength of the rotator cuff as it is the largest and strongest of the four muscles. This is the main muscle for the internal rotation of the shoulder. In neutral position it acts as a passive restraint. The supraspinatus muscle is one of the main muscles for the abduction movement of the shoulder and is as strong as the deltoid muscle for this action. When the supraspinatus is dysfunctional, the deltoid is the only muscle raising the shoulder in all directions. (see figure 1a & figure 1b) Figure 1a shows a full extension of the arm in which the deltoid muscle works against the weight of the hand and the arm from the moments the humerus is being pulled down: (multiplying the weight of the arm times gravity times the distance,
the moment results $174\text{Nm}$: $(25\text{kg} \times 9.81)^{0.71}=174\text{Nm}$; $(5.07\times 9.81)^{0.34}= 16.91\text{Nm}$. By adding the two results of moments, $(174\text{Nm} + 16.91\text{Nm})$, the total moment pulling the humerus down is $190.91\text{Nm}$. Figure 1b shows, deltoid $(d)\times \sin10\text{degrees}\times 0.01528d$, determines the moment of the humerus being pulled up. Assuming both equations are at equilibrium, the tension of the deltoid is equal to $12,494.11\text{N}$ (Newtons). When following the same procedure with the elbow bent, the deltoid tension is equal to $4,494.1\text{N}$. The result decreases when the elbow is bent because the weight and the force required to hold the arm are reduced.

The basic laws of physics state that energy is simply transferred from one object to another; it is not created nor destroyed. This includes the transfers of energy from person to person, from person to object, or vise-versa. The greater the energy transferred, the faster and stronger the energy is received. For instance, a baseball pitcher starts their energy with their foot, transfers it to the shoulder and arm, then to the hand, and soon transfers their energy to the released baseball, the catcher then absorbs this energy. The force exerted is delivered by the arm of the pitcher (rotation) and the distance between the catcher and the pitcher is the distance traveled by the ball. The mechanical work for the rotator cuff can be done using the product of the force exerted and the distance traveled: $W=FD$.

A force couple is a process that does not apply a resultant force, but does make use of a resultant movement. A rotation force is acted on two proportionately opposites. In the shoulder, the head of the humerus is proportionately opposite of the muscles of the rotator cuff. To avoid any sort of dislocation, the two forces must be acting upon each other. Another way the rotator cuff serves as a couple is between the four muscles it consists of, including the deltoid muscle. By working together, they secure the unstable joint known as the glenohumeral joint, in its place.

Newton’s third law of motion states that there is an equal and opposite reaction for every action. An example of this law is in weight lifting. When a weight lifter bench presses a high weight, they are pushing the weight away from them as the weight exerts equal and opposite force back to the weight lifter. The quality of the activities matters. The proper use of the muscles, the velocity of the motion, and the range of motion is imperative for professionals and patients to understand to be able to perform for optimal biomechanical position, which improves function and alleviates stress on the tissues.

**Cause of Injury**

A rotator cuff tear may be caused from acute injuries such as falls and hits, or by chronic wear and tear, where the tendon deteriorates. An acute tear may be caused when lifting heavy objects or sharply and quickly falling down. It can also occur due to other injuries near the area, such as a dislocation of the shoulder or breaking part of the clavicle. A degenerative type of tear is most common on the dominant arm and may appear as people age as the tendons wear down. Repetitive stress, lack of blood supply, and bone spurs are factors contributing to degenerative type of tears. Repetitive stress, doing the same movement over and over again, causes stress to the rotator cuff muscles. This factor is most common in baseball players, tennis players, weight lifters, and rowing athletes. A degenerative type of tear caused by low blood supply occurs as people age, impairing the body’s ability to repair damage. A bone spur factor is when there is bone overgrowth that may develop over time causing shoulder impingement when lifting the arms and eventually weakens the tendon and tears. About 20% of the general population are likely to have full thickness rotator cuff tears and the percentage increases for patients older than 50 years of age as they are 50-80% prevalent. People over 40 are prone to get the wear and tear due to aging. Others
such as athletes and people with repetitive lifting and overhead activities are susceptible to overuse tears or caused by traumatic injury such as a fall\(^1\). The higher the velocity, the greater the severity of the injury. In some sports, athletes may conduct repetitive movement at high speeds\(^5\). Damage to the rotator cuff may extend over time if consistently used or if injured again with worsening pain and weakness\(^1\).

**Determining Tears**

Etiologists have found the dysfunction of the shoulder and the pain associated with rotator cuff tears are lead to partial and full thickness tears by changes in the tendon from age-related degeneration, which is the most common, or acute injuries\(^8\). When the cuff is injured such as a tear, the bursa may also get inflamed and painful\(^1\). Tearing at least one of the tendons (most commonly the supraspinatus muscle and tendon), unattaches it to the head of the humerus. Tears begin as a piece of paper and the damage may progress to a complete tear by lifting heavy items or other difficult/progressive task. Two types of tears: when there’s damage to the soft tissue, but it is not completely split apart it is considered a partial tear. When the soft tissue splits completely into two pieces, sometimes may be where it attaches with the head of the humerus, it is considered a full-thickness tear (also known as a complete tear)\(^1\). There are three possible types of tears: crescent shaped tears, longitudinal tears, and massive and contracted tears. Understanding the determine types of tears are essential to determine the best repair technique for the individual, ultimately influencing the outcome after surgery and for their specialized rehabilitation therapy. Crescent shaped tears are typically wide anteroposteriorly and sort lateromedial. Longitudinal tears are long and narrow there are usually shaped as “U” or “C”. Massive and contracted tears are long and wide, typically greater than 2 cm with and length\(^8\).

Due to the low cost, a physical examination is most commonly used to diagnose shoulder injuries such as a rotator cuff tear. More advanced and accurate methods, such as ultrasound and MRIs, are also used to diagnose the injuries when difficult to identify the problem; however, they do tend to have higher costs\(^3\). X-rays and MRI (Magnetic Resonance Imaging) or ultrasounds are imaging tests used to confirm the doctor’s diagnosis. MRIs are the most detailed choice as they show the location, size, and whether the tear is new or old\(^1\). Magnetic Resonance Imaging (MRI) is a great technique used for rotator cuff tear. An MRI detects hydrogen atoms as it interacts with the magnetic nuclei, which is helpful for imaging tissues. Ultrasounds are also used and have frequencies that may extend between 1 – 20 MHz\(^5\). Berkovitch, Haddad, Keren, Soudry, and Rosenberg have introduced a new alternative method for results as accurate as imaging technology, yet simple and cost friendly as physical examinations. By measuring the mechanical force of the rotator cuff muscles, they study showed normal values of torque curves of the supraspinatus muscle: 61.8 +/- 18.3 N*m/kg, p<0.001. These values proved to be significantly different from the values of the rotator cuff tears or other shoulder injuries. These values ranged from 25.9 +/- 8.3, 22.6 +/- 5.6, and 22.9 +/- 5.5 N*m/kg, p>0.05. For the infraspinatus muscle, the normal torque values resulted to be 40.4 +/- 10.4 N*m/kg, p<0.001. The rotator cuff tears of the infraspinatus torque values tested 20.3 +/- 6.2, 26.3 +/- 6.5, and 5.7 +/- 2.9 N*m/kg and generated a force p<0.01. The lowest results of torque curve were found in the subscapularis muscle since the normal values and the values of the tears were about the same, about 29.0 +/- 8.9 and 30.4 +/- 7.0 N*m/kg and a generated force p<0.001\(^3\). The torque has a relationship between the force exerted and the distance of the shoulder and the angle in which it is used. Figure 2 shows how the twisting force can relate to the use of the rotator cuff\(^6\).
Surgical Procedure

In the United States, about 30–75 thousand repairs are executed for rotator cuffs annually\(^2\). Although there have been high rates of surgical failures reported despite the advances in surgical treatments and the conservative treatments, the most accepted or the standard treatment for rotator cuff tears are open and arthroscopic surgical repair. For that reason, designing a plan from a professional which will help improve the area biologically and by strengthening it mechanically, allows the patient to heal at their own pace\(^2\). Repair techniques depend on the severity of the tear. Open or arthroscopic are the main surgical approaches aiming to help restore movement/activities to the patient. Pain is reduced when the tendon heals to the bone. Subacromial depression, or acromioplasty, is commonly executed to reduce pain and protect the repair by increasing the subacromial space, although, some argue whether it is truly necessary or not. An open technique surgical approach is typically when the tear is large and/or complex\(^8\). An arthroscopic technique is the preferred surgical approach, as it involves improved instrumentation and the patient spends less time in the hospital and less pain after surgery, as well as they are able to return to work or any other activities sooner with less complications in the wound. Techniques such as single row, double row with suture bridges, and transosseous repair are grouped together. Metal, PEEK, or plastic anchors are used in the greater tuberosity with sutures attaching the rotator cuff to the bone. The sutures in a single row repair are used to adjust the rotator cuff from the anchor. A double row repair involves anchors and sutures of 2 rows. While the last repair technique, transosseous repair, does not involve the use of anchors. A double row repair is more costly, time consuming, and technically challenging, however, it leads to a tighter repair by increasing the contact of the tendon to the bone\(^8\).

Depending on the size of the tear, augmentation device such as a natural or synthetic scaffolds are used to ease the repair of the soft tissues by sutures and suture anchors, decreasing the chances tearing the tendons again. The larger the tear, the more likely this procedure will be necessary\(^2\). A reverse shoulder replacement reverses the position of the glenoid and the head of the humerus, adding more stability to the shoulder and allowing it to function properly without the rotator cuff. A plastic socket is used to give a bearing to the ball and socket, which is placed above the metal humeral stem. A metal tray and a metal ball are used to replace the glenoid. The metal tray is drilled and secured into the bone and is attached with the metal ball to plastic socket\(^7\). Other factors that also depend on the best surgical approach are the surgeon’s technique and experience as well as the quality of the patient’s bones and tissues\(^8\). Figure 3 shows an example of the surgical repair of infraspinatus rotator cuff tears and displacement measured is displayed through a thick black dot on the bone\(^2\).

Rehabilitation

After the doctor discusses the symptoms and medical history, they check if there is any deformity or tenderness around the area. They also measure the patient’s range of motion and the strength of the arm/shoulder\(^1\). It’s extremely important to maintain rehabilitation with both surgical and nonsurgical treatments as the muscles are weak and limited in movement. Therefore, an exercise or plan done by physical therapy specialist is essential to strengthen and improve the muscles in the area and may take several months to complete\(^1\). When there is immobilization and pressure add there is a better rate of healing, to the tendon. There is greater tension on the superior rotator cuff in 0 and 15 degrees abduction than in 30 and 45 degrees. The quality of the tendon, the surgical approach, the localization and arrangement of the tear, and the cause are factors that
influence the patient’s rehabilitation. Early or delayed range of motion is controversial. Some specialist believe that early passive range of motion for patients under 65 years of age with small or medium tears have no healing disadvantages or more consequences than those with delayed range of motion. On the other hand, other specialists encourage six weeks of immobilization with passive exercises and return to their athletic activities after six months.

To prevent shoulder stiffness, yet allow the tendon to bone healing process is the goal for a optimal rehabilitation. Water has shown/proved to have beneficial properties for relaxation, rehabilitation and training for centuries. It has buoyancy viscosity that not only improves stretching, but also gives the body an appropriate amount of resistive exercise (Figure 4). While making the movements passive and, therefore, easier carry out. Higher temperatures are found to have an effect on collagen, thus increasing the elasticity of soft and tissue and relaxes the muscles. This protective environment between 2–6 weeks after surgery is used to restore glenohumeral range of motion in the beginning phases of rehabilitation for the shoulder. Pools are used to provide this type of therapy, since depending on the direction/type of movement, the buoyancy of the water provides support, assistance, or resistance. For support, exercises involving abduction and adduction of the arms. With an upright standing position, such exercises are parallel to the bottom of the pool and perpendicular to the end of the buoyancy. Resistance exercises are typically done at the bottom of the pool due to its viscosity. As far as assisting exercises, those are done at the surface of the water, facing upward. The buoyancy counteracts with gravity’s force, making the weight of objects to weigh less under water. However if the exercises are done at a fast rate, the viscosity changes the purpose of the exercise. For instance, supportive exercises of movement become resistive exercise when they are executed too fast.

Other therapeutic agents used by PTs are ultrasound, iontophoresis, and transcutaneous electrical nerve stimulation (TENS) as methods to help their patients manage their pain by transmitting low amounts of electric current except the ultrasound provide heat through sound waves. The common phases for rehabilitation of the shoulder. Phase 1: Passive range of motion for week 1 through week 6. Passive exercises are pendulum exercises as it involves only about 15% of voluntary isometric contraction in the muscles of the rotator cuff. An example of the exercise is by holding a chair with the strong arm and hang the affected arm and swing with gravity drawing 20 cm wide circles while bent over at the waist (Figure 5). Phase 2: Active range of motion for week 6 to week 12. This stage helps restore muscle contraction allowing the patient to return to the normal every day activities. Although, resistance exercises are still avoided at this phase. Possible exercises at this phase are internal and external sub-maximal isometric rotations, which consist of bending the arms 90 degrees and holding them under the shoulders. Open chain proprioceptive exercises, scapula-thoracic exercises, and continuing aqua therapy are other possible exercises options the patient may begin, still avoiding resistance and strengthening exercises. Phase 3: Initial strengthening may begin between week 12 to week 16, as it varies for each patient to avoid having too much stress on the repair, subacromial impingement, and pain, the kinematics of the scapulothoracic and glenohumeral joints along with the compliance in soft tissue should be well improved. This phase involves internal and external exercises below the shoulder. Generally isometric exercises and later elastic resistance exercises are given to build endurance for the muscle. The muscles strengthened are the four rotator cuff muscle as well as the muscles surrounding and connecting to it such as the deltoid and trapezius. The arm muscles, biceps and triceps, are also muscles strengthened in this phase. The final phase is advanced strengthening which may begin between week 16 through week 22 after surgery. This phase is the transition to
athletic activities. External rotation movements of abduction at 90 degrees and push-ups from a wall to a chair, later on the floor. If all goes well the patient is strong enough, they may start to do plyometrics of the upper limbs to develop speed and power. Toward the end of this phase, patients are thought to have more awareness of their body’s position, neuromuscular control, and strength. An athletic patient, be cleared to may return to their sport after six months.

**Conclusion**

The rotator cuff consist of four muscles, the supraspinatus, the infraspinatus, the teres minor, and the subscapularis. The four muscles are attached to the clavicle, the scapula and the humerus through tendons and receive some assistance from other muscles around them such as the deltoid for movement. The basic law of physics, which state that energy is neither created nor destroyed, are found when exerting a force with high velocity on rotator cuff. The muscles exert a force towards each other to avoid dislocation. Injuries to the rotator cuff such as tears are may be caused when lifting heavy objects, chronic falls, wear and tears, as well as deteriorating of the tendons as people age. In severe cases, surgical procedures are necessary either by open or arthroscopic technique. Rehabilitation procedures are helpful to avoid stiffness post-surgery and slowly improve the patient’s range of motion. Physical therapists and other rehabilitation specialists may use aquatic therapy as the buoyancy of water provides appropriate resistive exercises appropriate for the body through support, assistance, and resistance. Passive exercises as well as ultrasounds, iontophoresis, and TENS are used to provide heat in the affected shoulder and release pain the low electrical current or sound waves.

Although, rotator cuff tears are known to be complex injuries to help recover from, I have grown to believe technology is evolving quite well to help improve this condition as well as other injuries to the shoulder. Aquatic therapy seems to have the most positive results and could be used more in the future to strengthen the rotator cuff. With the proper care and rehabilitation, I believe the patient can recover from their injury with strong results. Wanting to improve and most importantly time and patience are key to have greater results of recovery from both the patient and the specialist.
Figures:

**Free Body Diagram for calculating Deltoid Force**

![Free Body Diagram](https://www.shoulderdoc.co.uk/education/rotator_cuff_mechanics.pdf)

**Figure 1a:** Available from: [https://www.shoulderdoc.co.uk/education/rotator_cuff_mechanics.pdf](https://www.shoulderdoc.co.uk/education/rotator_cuff_mechanics.pdf)

Description: This image illustrates a full extension of the arm to find the force exerted on when using the deltoid muscle when raising the arm.

**Figure 1b:** Available from: [https://www.shoulderdoc.co.uk/education/rotator_cuff_mechanics.pdf](https://www.shoulderdoc.co.uk/education/rotator_cuff_mechanics.pdf)

Description: This image is an illustration of the bending of the elbow to find the force exerted when using the deltoid muscle.
**Figure 2:** Available from: [http://www.intechopen.com/books/theoretical-biomechanics/induced-acceleration-analysis-of-three-dimensional-multi-joint-movements-and-its-application-to-spor](http://www.intechopen.com/books/theoretical-biomechanics/induced-acceleration-analysis-of-three-dimensional-multi-joint-movements-and-its-application-to-spor)

**Description:** This image illustrates the torque when using the rotator cuff as the center point of flexion and rotation.

**Figure 3:** Available from: [http://engagedscholarship.csuohio.edu/cgi/viewcontent.cgi?article=1018&context=enme_facpub](http://engagedscholarship.csuohio.edu/cgi/viewcontent.cgi?article=1018&context=enme_facpub)

**Description:** Surgical repairs of the infraspinatus.
Figure 4: Available from: http://slideplayer.com/slide/8096978/

Description: the weight of the person and the force are exerted away from each other, providing support, assistance, and/or resistance.

Figure 5: Available from: http://jbjs.org/content/90/10/2171.full

Description: Passive exercises and stretches for strengthening the rotator cuff.
Cited References


The Phenomenon of Sonic Booms
Samantha Ansell
April 18, 2017
Physics 112: 11247
Professor Swingler
Abstract

An extensive review is presented of sonic boom theory, covering the basics of its history when discovered/recreated, people involved in that process and many still researching today, and physics behind the phenomenon. This review is intended to provide the non-specialist with insight into how sonic booms occur in nature when an object passes the sound barrier. When an object, like a military fighter jet, is flying in the air it creates sound waves, such as throwing a rock into a pond creates water waves. These sound waves are projected in front of the aircraft if it continues at a normal and/or regular speed. However, if the aircraft is speed up and catches up to the sound waves it breaks the sound barrier and puts off a noise which emulates a large explosion.

As far back as can be read, seen, or told natural phenomenon’s have occurred in life, in many circumstances. Phenomenon’s can be explained through scientific methods and religious explanations “as a matter of how people want to understand the meanings of natural phenomena as well as where they want to find that understanding” (Bice 2015). Scientific explanations must be verified through observations and facts, approved by many, and presented in a professional manor so that it may be accepted by many in science. Religious explanation is confirmed by faith internally within every individual who believes and does not need fact or reasoning to prove it. Though scientific explanations and religious explanations are based completely different, the existence of one explanation does not discredit the other strictly based on its specific processing methods. In fact, scientific explanations include the coordination of multiple science principles and religious explanations include the coordination of multiple biblical beliefs.

There are many events that happen every day that are categorized as phenomenon’s, such as, the pink lakes of Western Australia, the frozen bubbles affect when Abraham Lake in Alberta, Canada freezes over, sundogs, and so many more, but one major phenomenon in today’s day and age is sonic booms, which is the event at which an object breaks the sound barrier by traveling faster than the speed of sound and makes a sonic boom, loud explosion like noise. “A sonic boom occurs anytime an object is traveling faster than the speed of sound (about 600-760 mph depending on air density)” (Sarkar). The best-known sonic boom phenomenon’s in today’s day and age are events such as the noise of a whip or snapping of a towel, bullets leaving a gun, lightening, which we hear as thunder, and the most commonly known and researched event through military aircrafts, more specifically fighter jets (See Figure 1 for sonic boom from a fighter jet).

Sonic booms have been occurring naturally in life for many millions of years, but no one knew why an object created such a loud noise until research began. The sonic boom was not necessarily discovered, as it is not something that can be found at any specific time, however it was created and explored, first in ancient times and more recently at many research facilities, schools and accredited institutions. Furthermore, no sole individual discovered and/or experimented this phenomenon like many inventors have done in the past with every day products we use today. However, “in 1935 the conversation about challenges of supersonic flight led to the creation of the term ‘sounds barrier’, which seemed to imply a physical wall that could not be overcome” (Redd NT 2012). Objects such as bullets have been breaking the sound barrier for many years before the conversations in 1935, but the questions continued as to whether a
plane, or a man, could break the barrier and withstand the pressure during the flight and phenomenon.

When research was finally ready, researchers, experimenters, and many more professionals began the dive into creating a sonic boom “in 1944 when the US Air Force (USAAF) and the National Advisory Committee for Aeronautics (NACA) agreed on a joint program to investigate the possibility of supersonic flight” (Aerospaceweb.org, 2009). The mission began with the plane, the Bell X-1 or Glamorous Glennis, and the pilot, U.S. Air Force Captain Chuck Yeager. The plane, the Bell X-1 research plane, better known as “Glamorous Glennis,” which the pilot named in tribute to his wife, was a bullet nosed plane modeled after a .50 caliber machine gun bullet, which became the first plane to fly faster than the speed of sound (see Figure 3). This specific plane was built and able to surpass the sound barrier because of its perfect bullet shaped body, rocket engine, the way it was dropped into the sky like a bomb and the mechanics that not only the pilot figured out, but the team too.

The planes shape was simulated after a bullet to allow for excellent air flow around the plane when in flight. Its specific dimensions, weight and propulsion are as follows: Dimensions: length 33.10ft, wingspan 28.08ft, height 10.85ft, and wingspan Area 130.00ft², weight: empty - 4,890 lb. and normal weight - 12,225 lb., and propulsion: powered by one Reaction Motor XLRll-RMB Rocket Motor with a thrust of 6,000 lb. (Aerospaceweb.org 2009). The Bell X-1 used this motor because at the time no jet engine was strong enough, so the Bell X-1 used this four chamber XLRll-RMB rocket engine, which produced 26,500 newtons (6,000 lb.) of static thrust, that they called Black Betsy (See Figure 4) (Smithsonian). Last, but not least, every time but once the “Glamorous Glennis” dropped out of the sky via the bomb bay of a modified Boeing B-29 or B-50 bomber plane. This was done to conserve fuel, but also to be able to run the test multiple times. “The X-1 was lifted to an altitude of 20,000 ft (6,100 m) before being released to ignite its rocket engines. This technique was advantageous since it improved safety in ground operations and also vastly increased the aircraft's performance” (See Figure 5) (Aerospaceweb.org 2009).

As for the pilot, he had just as curious a life and adventure as the Bell X-1, Glamorous Glennis. Chuck Yeager was born February 1923 in West Virginia, where he enlisted in the United States Army Corp to serve in World War II after graduation from high school (Academy of Achievement 2016). Once enlisted, Yeager became a well decorated fighter pilot in the war, and following went on to become a flight instructor and test pilot, which ultimately lead to his assignment of piloting the Bell X-1 in 1947. The first test launch took place on Aug 29, 1947, which reached two hundredths of a Mach number, and by the sixths launch he was reaching .86 Mach with the X-1 plane. Despite the positive progression of the launches, when it came time for the seventh launch Yeager was up in the air when he began to experience turbulence from the shock wave formed by the compression of the air, and at that moment he had to abort the mission, dropping all fuel and shutting off the engines to land safely in the desert (Redd NT 2012). After his seventh launch failed, the eighth launch was set for October 14, 1947 with the intent that Yeager would reach the Mach 1 needed to break the sound barrier. Only one problem presented before the launch. Two days before, Yeager and his wife, Glennis, went horseback riding where Yeager was thrown from the horse breaking two right ribs, which happens to be the side he needed to close the cockpit door. Author Stephen Sherman explains the events to follow in his article on aeropilots.com:
“He couldn't have reported this to the Army doctors; they might have given the flight to someone else. So Yeager taped up his ribs and did his best to keep up appearances. On the day of the flight, it became apparent that, with his injured right side, he wouldn't be able to shut the door of the Bell X-1. In the plane's tiny cockpit; he could only use his (useless) right hand. He confessed his problem to Ridley, the flight engineer. In a stroke of genius, Ridley sawed off a short piece of broomstick handle; using it with his left hand, Yeager was able to get enough leverage to slam the door shut” (Sherman 2011).

With that being said, “Air Force Captain Charles E. Yeager flew this experimental Bell X-1 on the eighth flight faster than the speed of sound” and during this flight “he reached a speed of 1127 kilometers per hour (700 miles per hour), or Mach 1.06 at an altitude of 13,000 meters (43,000 feet)” (Smithsonian). Now, Yeager didn’t stop there, in fact, he continued on to fly with the United States Air Force collecting a long list of prestigious awards, honors and career advancements from enlistee to brigadier general, ran own Pilot School for astronauts in the space program, “he was the first and youngest military pilot to be inducted into the Aviation Hall of Fame (1973)” (Academy of Achievement 2016), had a bestselling nonfiction book, The Right Stuff (1979), and movie made after him, his story and his achievements, and last but not least, he made his last flight as a military consultant and broke the sound barrier once again on the 50th anniversary of his first flight on October 14, 1947, but this time in a F-15 fighter (Academy of Achievement 2016).

Physics Aspect of the Phenomenon

“Have you ever seen a boat move through water? The bow waves (front) and stern waves (back) are similar to the invisible pressure waves created by an object as it moves through the air” (Wonderopolis 2012). When an aircraft passes through the air it also creates pressure waves that push in front of the aircraft and behind. These pressure waves travel at the speed of sound and depending on the speed of the object (plane in this instance) as it fly’s the waves stay in front and back. However, as the plane increases in speed, “the waves are forced together, or compressed, because they cannot get out of the way of each other, which causes them to eventually merge into a single shock wave, which travels at the speed of sound, a critical speed known as Mach 1, and is approximately 1,225 km/h (761 mph)” (See figure 2 for example) (Tek-Th!nk).

Mathematically, a sonic boom is calculated, or caused through, the below equation, which expresses $v_s$ (as the speed of source (the plane) and $v$ as the waves (UNSW).

$$\sin \Theta = \frac{v}{v_s}$$

When $v_s$ is greater than $v$ then you are given a shock wave, which results in ears down below hearing a sonic boom, and the equation below.

$$\sin \Theta = \frac{v_s}{v}$$

At the moment that a plane passes the speed of sound it then breaks the sound barrier, which often creates a white cloud, where the clear air around the plane becomes supersaturated (it’s humidity can exceed 100%). This white cloud can also be explained by “the sudden over and under pressure of the shock wave, which abruptly heat and cool the air respectively, appear to have nucleated water droplets to form a cloud (UNSW).
Another example of the sonic boom phenomenon is the noise of a whip when it is cracked. Professor Goriely of the University of Arizona Department of Mathematics states that "the crack of a whip comes from a loop traveling along the whip, gaining speed until it reaches the speed of sound and creates a sonic boom," (Harris D 2002). Using Figure 6, which is from a research project performed by a computer scientist and paleontologist where they compared a whip to a dinosaur tail, you can see how the whip simulates a dinosaur tail, which they proved could have not only broken the sound barrier, but highly surpassed it in speed. Although the cracking of a whip seems like an unusual motion, it mimics the same motion as other objects that also create a small sonic boom, such as fly-fishing string and a sperms tail. Professor Goriely also states that "although the loop travels at one speed, some parts of the whip, including the tip in the final stages of motion, travel twice as fast. Even though those parts are moving twice as fast, it is the loop itself that generates the sonic boom," (Harris D 2002).
Figure 1:

*Was found free to use with google advanced search.

Figure 2:

Woollaston V. 2014 Dec 1. So secret, its existence is not even acknowledged: Futuristic 'Aurora' spy plane that travels at SIX TIMES the speed of sound is blamed for mysterious booms heard at the weekend <http://www.dailymail.co.uk/sciencetech/article-2855795/So-secret-existence-

Figure 3:


Figure 4:
Smithsonian: *National Air and Space Museum*. Bell X-1 *Glamorous Glennis*.  
<https://airandspace.si.edu/collection-objects/bell-x-1> Accessed 2017 Apr 12.

Figure 5:
References


Ultrasounds
Madeline Arnold
April 18, 2017
Physics 112
Spring 2017
Professor Swingler
Abstract

Ultrasounds are a critical piece of equipment which are used for many different medical purposes. In this case, there seven versions of this machine that have specific purposes. Each of the different machines are similar in appearance and have the same basic parts. The way each machine works and functions differently will be discussed. This piece of equipment uses a basic topic of physics called soundwaves. Finally, potential safety risks, benefits, and what the future holds for this machine will be reviewed.

Introduction

The ultrasound is a very unique and important part of the medical field. It can be used for a wide range of health problems, including prenatal care, cardiology, urology, and it is now being used in the emergency room to provide the doctors with a quick picture so that they don’t have to always wait on results. Although there are many uses for an ultrasound they are each performed in a similar fashion. The main function of this machine is to send soundwaves at high frequency into the area of the body that is being examined, and then receive the wave the body sends back. Depending on what the ultrasound is being used for will dictate which kind of machine will be used and what frequency will be needed.

Basics of the Ultrasound Machine

Each machine consists of seven basic pieces. This includes the transducer probe, central processing unit, transducer pulse controls, the display, the keyboard and cursor, the disk storage device, and the printer. (See figure 1). The transducer probe is one of the main aspects of this machine that uses basic physics. The wand is the part of the machine that sends and receives the sound waves that are produced. In this case, the size of the probe is vital because it is what determines the frequency of the sound waves, which in turn decides how far the waves will travel. In addition, the size is what gives it a visual range and how good the picture will turn out. The central processing unit is defined as the brain of the machine. Its main purpose is to perform the calculations and is the main power supplier. Once it inputs all the data the central processing unit will give an image. Next, the transducer pulse control enables the user to control or measure the frequency, length of the pulse given out by the probe, and the amplitude. The main point of the display screen is to show the readings the central processing unit drives. It used to only show images in black and white, but with new advancements in this technology it can show images in color or 3D. The keyboard and cursor’s main function is to enable the user to enter the information into the central processing unit and add any new information or measurements. Next, is the disk storage device, which is where all the data that is both entered into and relayed out of the machine is stored. The are many types of disks that can be used with this machine including floppy disks, CDs, or DVDs. This is what is kept with the patients file as well as what can be given to the parents after an ultrasound for them to take home. Lastly, the purpose of the printer is to make copies of the final image. The picture that is printed will also stay with the medical file and a copy of it can be given to the patient. As shown, each of the seven types of ultrasounds have a very specific and distinct functions, even though they all contain the same basic parts.
Different kinds of Ultrasounds

There are seven types of ultrasounds that are used in the medical field today. This includes transvaginal scans, standard ultrasound, advanced ultrasound, Doppler ultrasound, 3-D ultrasound, 4-D or dynamic 3-D ultrasound and the fetal echocardiography. The transvaginal scan is used during the very early stages of a woman’s pregnancy. It includes a unique probe that is made to be used inside the vagina. This can allow for an early sonogram image of the fetus. This type of ultrasound can also generate images of the woman’s ovaries and uterus. A standard ultrasound, also called the traditional ultrasound, produces a 2-D image and is usually used to observe the abdomen area of the body. The third type of ultrasound is called the advance ultrasound. With this this type of ultrasound it also creates a 2-D image, but unlike the standard one, it targets a more specific area and needs equipment that is of higher quality. The next type of ultrasound is called the Doppler ultrasound. It is a very special ultrasoun because it has the ability to make images based on miniscule changes in frequency. This occurs when the soundwaves rebound off moving pieces in the body. For example, it can produce an image of a person’s blood as it moves through a vessel. The Doppler effect of this machine will be discussed further in the physics section. Next is the 3-D ultrasound, which has a specific probe that enables a 3-D image to be produced. It is mainly used to create a 3-D image of a fetus in the uterus. The sixth type of ultrasound is called the 4-D or dynamic 3-D ultrasound. This takes the 3-D image produced and adds movement to it. It can be used to examine the face of the fetus and track its movement prior to the actual delivery. The last type of ultrasound is called the echocardiography. This kind of ultrasound uses specific wavelengths to check someone’s heart functions or to evaluate the anatomy of a fetus’ heart. Although each of these ultrasounds have different uses and functions, they each utilize the seven basic pieces that were discussed above.

The Physics of an Ultrasound

The ultrasound is a machine that uses non-ionizing radiation which produces sound waves at a high frequency. These soundwaves at high frequencies are used to create an image of the body that it is scanning. The ultrasound utilizes a mechanical wave called sound. A soundwave can create a disruption in the medium in which it is traveling through, while also transferring the energy created from one area to another. This energy is made from soundwaves bumping into other particles that are present. There are two types of soundwaves that can be produced due to this movement, transverse and longitudinal. The sound wave production is dependent on the directionality that the waves hit the other particles compared to which way the energy is moving. A longitudinal wave occurs when the movement of the particle is parallel to that of the sound wave movement. A transverse wave is when the movement of the particle is perpendicular to the movement of the soundwave. Any type of medium can create a longitudinal wave whereas transverse waves can only appear in solid material. This is the main reason that the ultrasound can be carried through the tissue on the body.

Frequency of the soundwave that is made is done so by the number of cycles that occur in one second. The definition of an ultrasound is a soundwave that cannot be heard by the human ear.

\[ f = \frac{1}{T} \]
F is the frequency of the wave
1 represents one second
T is the time it takes for one cycle to occur

Wavelength, however is shown to be inversely proportional to the frequency given off. It is defined as the length that a soundwave traveled in a single cycle. When a wavelength is short it will give a larger resolution and less penetration of the tissue, a high frequency probe is needed for artificial areas, while low frequency probes are used for deeper areas.

\[ v = f \lambda \]

\( v \) = velocity  
\( f \) = frequency  
\( \lambda \) = wavelength

**Image Production and Resolution:**

The waves of an ultrasound are made from the combination of electrical currents moving across a piezoelectric crystal that is found in the probe. The pulses generated from the ultrasound waves are then transferred through the body’s tissue. The crystal in the probe will take the rebounded echoes emitted from the body and will send out the next pulse. Each pulse usually lasts for 1µs and then it proceeds in intervals of 1 m/s. This means that each crystal receives around 99% of the echoes emitted from the body and only discharges ultrasound waves around 1% of the time. In the medical field, the probe is comprised of multiple crystals in what is called a phased array. When many crystals are placed in the probe, they are excited simultaneously through the electrical pulses. These pulses move back and forth to each end of the probe. When the ultrasound wave moves through the body, it can run into the interfaces that are found within tissues. If this happens, some of the energy continues into the body and some is reflected. When the angle made between the transducer and the interphase is more than sixty degrees then the echoes that are reflected will head straight back to the transducer. Once these echoes are back to the crystal it will convert the waves into pulses which are in turn what creates the 2-D image displayed on the screen. In this case the more echoes that are given off is directly proportional to the brightness of the image that is made. Another aspect that can play a factor in the brightness of the image is the acoustics or density of the air tissue or bone within the human body. The bigger the difference in density the brighter the image. But when the waves hit the interface between bone and tissue almost all the energy will be reflected, which means that no image will appear. This is the main reason that an ultrasound is not used to gain information about bowels, lungs, or bones. In addition, the technician uses a gel on the portion of the body that is being examined because it helps to eliminate air between the probe and the skin. When constructing a 2-D image the depth of the interference of that tissue must be found.

\[ d = \frac{ct}{2} \]

\( d \) = depth  
\( c \) = average speed of an ultrasound through tissue. It is constant at 1540 m/s.  
\( t \) = time it takes for the pulse to travel to the interface.
Attenuation is the process in which the waves passing through the human body slowly lose their energy, which results in the processes of absorption, reflection, diffraction, and refraction.\(^2\) (See Figure 2). In between the tissues in the body is the area in which both reflection and refraction take place.\(^2\) While reflection makes the echoes, refraction will redirect the wave in a new direction. Then there is diffraction which happens when the soundwaves are dispersed. This usually happens when the wave hits a small piece in the body.\(^2\) The main area where energy is lost in this process is during absorption. This occurs when energy is transformed into friction and is then lost in the form of heat. Between the attenuation process and a minute amount of the waves energy that is returned to the transducer, leads to the reason for time gain compensation. This allows the machine to make up for the lost energy, by aiding in enlarging the echoes that move at a slower pace, which in turn creates an even image.\(^2\)

There are two types of resolution used in ultrasound technology. The first is called axial resolution.\(^2\) This type of resolution has the power to tell the difference between objects that sit perpendicular to the soundwave. In this case, the object that is in the way of the wave can only be seen if that object is bigger than the produced wavelengths, meaning that the smaller the object the higher the frequency of the probe needs to be. The only drawback is that the high frequency waves have a higher chance of losing their energy and therefore it can only read objects that are superficial to the surface.\(^2\) The other type of resolution used is called lateral resolution. This type of resolution can show the difference between two objects that are right next each other. Without lateral resolution, the reflected waves will interact with each other which causes the transducer to think that the two objects are one object.\(^2\)

**Artefacts, Enhancement and Shadowing:**

Artefacts, enhancement, and shadowing can both aid and mislead a technician when they are viewing the image produced by the ultrasound. There are three different artefacts, one type of shadowing, and one type of enhancement that can be present in an ultrasound scan. The three types of artefacts include reverberation, ring down, and mirror. A reverberation artefact will happen when there are two reflective surfaces that cause the sounds waves to bounce between, before it heads back to the transducer.\(^2\) This can cause some waves to make it back to the transducer before others, creating some evenly spaced lines on the final scan. As of now it is believed that this happens when the waves come back from object located deeper than the intended target. The next type of artefact is called the ring down artefacts. This happens when the soundwaves hit a small object that in turn resonates at the exact same frequency that the ultrasound had originally sent. It will occur after the first wave is reflected and thus causes it to be seen as another echo, that is believed to have come from an object deeper than the goal area.\(^2\) The last type of artefact is called the mirror artifact. This happens when the beam hits a surface that reflects easily. In this case, the beam will hit a highly reflective surface, bounce off, hit a surface that is close by, and then it returns to the highly reflective surface before heading back to the transducer. When this movement comes up on the scan it looks like it originated from an area opposite from the intended target as well as it looks as though it came from a deeper area of the body.\(^2\) The next item that can be seen on an ultrasound scan is called posterior acoustic shadowing. This occurs when the wave encounters an object that is a very reflective or thin surface. When the bean passes through one of these areas it will not move through and therefore, creates a large black space under the intended object. Due to this, anything that is under that
object will not be seen. The posterior acoustic shadowing is for the most part a good thing because it can indicate areas that are calcified or unnaturally dense. The last item that can come up on the ultrasound scan is called posterior acoustic enhancement. If a wave passes through an area in the body that is less dense or thin, it will cause anything after the intended target to be very light. This is caused due to the fact that a larger amount of the first wave will reach the object that is being observed and then allows for the energy of the echo heading back to the transducer to be increased. Each of these five artefacts can both negatively and positively impact the final image that the ultrasound produces. (See figure 3).

Doppler Effect:

The Doppler effect is what the Doppler ultrasound is based off of. This uses the basic measurement between frequency transmitted and frequency observed. Which is shown as \((f-f^1)\) and known as the Doppler shift frequency. The effect is used when the frequency changes due to the wave being reflected off a moving interface. An example of this would be when blood cells move closer to the transducer forcing the reflected waves to compress and fling towards the transducer, shortening the wavelength and increasing frequency. This can also occur when the blood cells are moving away, but the effect will be opposite. When this occurs it shows how velocity is inversely related to frequency. This equation will also take into account the cosine angle between the vessel and the probe.

\[
\frac{(f-f^1)}{f} = 2\left(\frac{v}{c}\right)\cos\theta
\]

\(f\) = transmitted frequency
\(f^1\) = observed frequency
\(v\) = velocity of the interface
\(c\) = speed of sound

This equation shows us that if the probe is perpendicular to the vessel being observed it will not create the Doppler effect, because if the angle is 90 degrees the cosine would be zero. In this case, the ideal angle used is below 60 degrees. Doppler imaging will help in providing an image of the direction that the object moves. This is usually shown on a color map or a graph showing time versus velocity. Overall, the Doppler effect can give the technician the flow velocity and it is very important when examining the blood flow through the body.

Additional Information

Benefits and Risks:

Some benefits that come from ultrasounds include their ability to check the status of a fetus or check on the blood flow in the body. As of now there are no known short-term medical risks from having an ultrasound performed. This is due to the fact that it does not create any ionizing radiation while being performed. Although if not used in a proper medical facility and used by a medical professional, there could be some unknown biological risks involved. Although, as of now the ultrasound machine has been deemed safe the long-term health risk are still unknown.
Future:

Due to the major advancements in technology in other fields, it can be concluded that the ultrasound itself will advance with it in the near future. Some aspect of the basic machine that will most likely be changed include the size of the machine, the amount of data it can hold, and the size of the probe. In addition, the 3-D ultrasound will probably see more advancements. Another thing that could be updated is the probe itself, meaning more insertable probes will be created to enable the doctors to get a better look at internal organs. A new research study out there that is focused on a new procedure that would enable the technician to see inside the patient as surgery is being performed. There are three new uses for ultrasounds that could be coming in the near future. The first would be using a 3-D ultrasound as a screening tool for breast cancer. It is their hope that this could enable the doctor to make a 3-D image of the breast tissue so that it is easier to find any lumps that are present. Another future possibility for ultrasounds, is using it find any blood clots in someone before they go into heart surgery. This method would utilize the 4-D ultrasound. Lastly, there is a study already underway that is using ultrasounds to see if they have any arterial blockage in their lower extremities.

Conclusion

An ultrasound is the transfer of soundwaves into the body and how they reflect off of different portions of the body. Each of the seven parts of a basic ultrasound was discussed and all seven types of ultrasounds were compared. It was shown that the frequency was measured by the amount of cycles that happen in one second. In addition, the relationship between wavelength was examined in depth. It was concluded that they have an inverse relationship. Then the resolution of the ultrasound scan was researched in great detail. Next, it was revealed how different aspects of the soundwave patterns can impact the final image. Then, the Doppler effect was talked about more in depth and it was shown that this is a unique type of ultrasound. Lastly, the safety, the risks, and the future of the ultrasound were brought up. Overall, it was shown throughout this paper what a major role physics plays in creating that finale scan.
This image shows the different pieces of the ultrasound and how they interact with each other and the patient.

Figure 2

Figure 3.
Causes of attenuation: (a) absorption, (b) reflection, (c) diffraction and (d) refraction all contribute to the overall attenuation of the ultrasound wave seen in (e).

Figure 3
This is an image of the different type of interference that can occur in the ultrasound scan. This includes artefacts, enhancement, and shadowing.

References


Cloaking Devices: When Theory Becomes Reality

By Hailee Ball

20 APR 2017

PHY112

Dr. Durandet
ABSTRACT

It is every child’s dream to own a cloak like the one seen in the “Harry Potter” movies, in order to get away with different situations their parents would not approve of. As great as that sounds, it is not within reach at that capacity for the time being. However, there have been many breakthroughs in different kinds of cloaking devices, which allows different objects to be hidden from the human eye in certain environments. There have been setbacks as well as failures, but in spite of those failures, a few successes have arisen and with those successes, the time to push forward is now. Through the use of conditions such as metamaterials, light rays, thermal detectors, and other various components, cloaking devices are no longer just a theory; it is reality.

In nature, there are different forms of ‘cloaking devices’ or camouflage that in fact do not need technology to exist; there are disguises, coloration, mimicry, and even active camouflage. The disguises in nature consist of animals like the crab spider which resides in south America, and feeds on turtle ants. In order to surpass the colony the spider puts the dead corpse of the last victim on its shell, that way the other ants just think it is another ant and not in fact a spider. An example of coloration in nature is birds, who typically have white bellies, and this makes it hard for hunters, because when they look up trying to find the birds, their light color feathers blend in with the sky. Not only that but the dark feathers on a birds back protects them from predatory birds flying above, because they will blend in more with the trees and ground. Mimicry is something most stick bugs do to survive, each species, depending on the environment in which they reside have varied appearances. Another great example is the scarlet king snake, their coloration mimics the coral snake, so other animals believe them to be venomous when they are not. Active camouflage requires energy and is very cool to witness. Many animals like squid, cuttlefish, and octopods have multiple layers of skin that can react in different ways. One of the layers can actually reflect light back a different way when angled, almost like small mirrors, and that layer is called iridiophores; this is also what chameleons use to change color. Another layer which consists of chromatophores, uses different pigments when needed to change its color to match the surrounding environment. There is even a third layer that can change the skins texture if needed. This is all something that comes from nature and has only been studied, not modified, meaning if this can somehow be translated into use for humans, it would be one of the largest breakthroughs in history presumably.

A breakthrough in science happens almost every day with the rate of discovery the world seems to be at. However, a breakthrough in theoretical science that has long been fantasized about is extremely rare to come across. With modern day technology, cloaking devices have gone from something only seen in movies, to real time successful experiments that will soon have a modern day purpose. The ideal result would be “The invisibility cloak designed with current transformation model would allow one to see the outside world un-anamorphically, while perfectly concealing his/her presence. This model should be fit for cloaking objects with arbitrary characteristics (e.g., size, shape or material properties) under all circumstances” (Zhu et al). Another example for what the end result should be is described as “A perfect cloaking
device can be identified by two specific quantities; First the electromagnetic field is annulled within the cloaked area and second, the wave after passing the device has regained its original shape” (Håkansson, 2007). This equation can be seen below in figure 4 where $s \rightarrow$ is equal to the “binary string of parameters that codes the rod distribution of the SOE device” (Håkansson, 2007). Although our technology is currently not to this point yet, there have still been some large breakthroughs.

Similar to most scientific breakthroughs, the experiments started out on a smaller scale due to the known difficulty of attempting to make something extremely macroscopic invisible. Many experiments also started out in water to attempt the cloaking of aquatic life due to waters higher refractive index, but it “required transforming electromagnetic space around the hidden object in such a way that the rays bending around it have to travel much faster than those passing it by.” (Chen et al, 2013). This is due to the fact that if light rays all around the object were one length, the bending rays around the object would be even longer, which requires those rays to move at a faster speed as to preserve its phase (shown in figure 1). In this particular experiment, six isosceles triangles were placed in the arrangement shown below in figure one, where the center was open. This opening is what the fish had to swim into in order to be fully cloaked. It worked because as the fish went through the opening, it seemed to disappear altogether because it was no longer seen through the camera, although it was known to still be present. This experiment was successful but there were some obvious flaws that would not allow this to be very useful or beneficial in real world applications. The glue that held the triangles together was still clearly seen, and it is not ideal at all to have to be in one particular place to be cloaked. This all being considered, it was soon discovered that this would not be sufficient on a larger, more applicable scale and a new technique would be needed.

It was proved that in a terrestrial environment with air as the medium, using optical glass in the shape of 4 isosceles triangles and 4 right-angle triangles with a refractive index of $n_1=1.78$ (shown in figure 2) would cloak a living, moving, macroscopic being such as a cat. This experiment was the same premise as the first one but instead it was not in water and was on a larger scale. Although a different experiment, it too has the same flaws as the first experiment. Something that was discovered in both the first as well as second experiment was the double-blind issue, which caused not only the observers to be unable to see the object, but also the object would not allow sight of the environment outside the given containment. Although the cloaking of the object was a success, there really is not a great applicable use for it if whatever or whomever you are concealing cannot see their surroundings. These displays of cloaking, although flawed in their own ways, were still a great step towards the end goal.

To try and rid of this double blind issue, as it is not practical in real world applications, external cloaking was presented as a solution, which is where the object itself would in a way be cancelled out. “However, the external cloak can only work in a steady state, since the multiple-scattering between the cloaked object and anti-object has to reach equilibrium to achieve the destructive interference of scattered waves” (Zhu et al). This means that there must be double negative parameters in order for the anti-object cloaking to be efficient. Although this would help with the double blind issue, it does not completely resolve it because someone would only
be able to see the environment around them in a distorted manner due to the distortion of the penetrating waveform.\textsuperscript{3} So, it does technically get rid of the double blind issue for the most part, but seeing distorted images isn’t very helpful especially if this technology would be used for combat situations or something to that caliber where clear vision is simply a must have.

Proceeding this, the idea was modified from double negative parameters, to single negative parameters. The properties of the cloaked object as well as the host material are then independent of the materials with single negative parameters. “…the broad-band cloaking still can hardly be achieved because the operating frequencies of complementary media must overlap each other” (Zhu et al).\textsuperscript{3} Also, adding on to the reason why this is not yet possible is due to the fact that this can only happen with a host medium that is less compressible than the cloaked object itself which isn’t a great combination.

Metamaterials play a large role in creating cloaking devices because “Metamaterials have been shown to revolution the control of electromagnetic waves in the microwave region showing unnatural phenomena such as negative refraction” (Håkansson, 2007).\textsuperscript{5} Negative refraction is when an object doesn’t consist of normal refraction and instead is reversed, as seen below in figure 3, it is an unusual phenomenon. Making an object invisible by guiding the light to go around the object, therefore making it invisible is just as difficult as it sounds and has still only been done on a very small level. Scattering optical elements, or SOE’s “have shown a great ability to control the propagation of electromagnetic waves” (Håkansson, 2007).\textsuperscript{5} These are quite similar to metamaterials in this way, but SOE’s are designed by an inverse design tool. The inverse design tool “is an integration of a two dimensional (2D) multiple scattering theory (MST) electromagnetic solver with a global optimization method, the genetic algorithm (GA)” (Håkansson, 2007).\textsuperscript{5} The MST method is faster than numerical methods and is based on the T-matrix. “The speed and accuracy, makes the MST a very good candidate for SOE design. However, when dealing with non-circular scatterer the method is limited by Graf’s formula” (Håkansson, 2007).\textsuperscript{5} This issue is solved by creating a minimum separation distance to ensure validity, the amount of distance is simply the length between both materials cannot be within the rod scatterer’s surrounding radius. Two rods next to each other may never be overlapping on top of each other in order to obtain this specified distance. On the other hand, the GA is an optimization algorithm that is more powerful, and it uses computational power which is very low cost. “To facilitate the optimization process the position of each scattering rod is determined by a fixed array of lattice sites (LS). The full SOE device is then coded by a string of binary bits where each bit codes the presence or absence of a rod at each LS” (Håkansson, 2007).\textsuperscript{5} This has been proven to work for making very complex devices.

A transmission- line cloak is made up of multiple layered two dimensional networks, and they are arranged in parallel strip transmission lines. The diameter is about 71 mm and the object has 2 mm diameter rods, 21 of them, each with an array period 5 mm. The x-y plane, which happens to be the largest, has an outer dimension of 24.5 mm. “Here we use the measured data obtained with a rectangular waveguide having a passband around 3 GHz and the TL cloak enclosing the metal rod array placed inside the waveguide” (Alitalo, 2009).\textsuperscript{6} The cloak itself is made out of a bronze and beryllium alloy in thin sheets that has been etched into. By these dimensions given, it
is quite obvious that no large object would work for this type of cloaking, however, on a small scale it functions well.

As far as a metal-plate cloak goes, there are multiple metal plates in a cylindrical shape, layered, that have a circular hole that is actually cut inside them, and that is the space where an object would be cloaked. “Due to the fact that the height of these waveguide plates decreases radially as moving from the outer surface of the cloak towards the inside, the electric field which is polarized orthogonally to the metal plates travels inside these waveguides around the cloaked region” (Alitalo, 2009). Due to the fields being concentrated into an incredibly small region, the fields cannot couple into the region that is being cloaked in the inner surface of the cloak itself. This is the result of the waveguides having a very small height. This cloak is different in the way that it operates for electric fields that are z-polarized. The diameter of the actual cloaked cylinder turns out to be 30 mm, however, the inner diameter of the MP cloak is 32 mm, and the diameter of the outer is 70 mm. The cloak plates consist of copper, while the cylinder being cloaked is made from steel. The MP cloak measurement is actually found in the same way the TL cloak is found, through the waveguide.

Coordinate- transformation cloak is another form and it is formed from an anisotropic material layer. Anisotropic is when an object having different values for a single physical property when being measured in different directions. “To enable good operation of the cloak and mitigate the effect of dispersion, we position the resonance frequency \( f_0 \) at a much lower frequency than the operation frequency of the cloak” (Alitalo, 2009). Due to the large inclusion dimensions, this would not be applicable for possible designs. When compared to the last two cloaks, this cloak has a narrower bandwidth, which is to be expected when considering the permeability of the CT cloak is more dispersive. This all is in spite of the permeability resonance frequency to be distant from the design frequency.

Pulse propagation has been studied inside the waveguides that are enclosing the variety of cloaks. Using the frequency-domain transmission data, it is investigated analytically. “First we choose the center frequency of a Gaussian pulse \((f_c, G)\) that we want to transmit through the waveguide and apply a Gaussian filter to create a time-domain signal \(y(t)\): \(y(t) = \sin(2\pi f_c G t) e^{-t^2/2\sigma^2}\), where \(\sigma\) is the standard deviation and the center of the pulse corresponds to time \(t = 0\).” (Alitalo, 2009). A Fourier transform, which is a function coming from a different function being represented by sinusoidal functions, is applied to the time domain pulse and then the complex transmission coefficients and each frequency component are multiplied. Using the inverse Fourier transform, the time domain signal is found from the output of the waveguide section. It is difficult to define, in numerical terms, how the cloaks worked; if it was bad or went well based of the time-domain results. Due to this issue, the correlation coefficients must be calculated between the pulses going through the cloaks as well as the pulse that passes through the empty waveguide. “Small values of \(1/\sigma\) correspond to pulses which are wide in the time-domain and narrow in the frequency domain, whereas large values of \(1/\sigma\) correspond to pulses which are narrow in the time-domain and wide in the frequency domain.” (Alitalo, 2009). For narrow pulses in the frequency domain, the cloaks work well. However, as the bandwidth of the pulse increases, the operation for the CT cloak deteriorates at a faster rate than the other cloaks.
Considering the CT cloak has the narrower bandwidth, as well as its stronger dispersion, this is to be expected when compared to the other cloaks.

In spite of most experiments to date being only partially successful or applicable in the real world, there is one form of cloaking that is as of now without flaw and ready for real world use. It is thermal cloaking in the use of defense for military members on night missions, where normally they might not be seen but if the enemy could sense their heat signatures they would be in trouble. With the research done at the National University of Singapore, not only did they successfully figure out how to camouflage thermal heat signatures, but they did it on a large scale applicable for a human to use, and even at a very low cost. “We have managed to control the thermal illusions’ shapes, material properties, distributions, and locations using bulk ‘natural materials without sophisticated fabrication” (NUS, 2014).4 This is a big breakthrough in metamaterials considering in other forms of cloaking devices, it is one of the main issues, which is why the team at NUS states “This drastically overcomes practical and challenging limitations of metamaterials which are not found in nature and hence would require complicated and complex design to imbue them with special properties” (NUS, 2014).4 They are already putting it out to the military and it should be seen being used relatively soon, once provided on a large scale in numbers.

Now that many of the theories and current successful as well as unsuccessful experiments have been explained, it is important to know why exactly this topic is so important. In modern everyday use it may not mean much, as it may never be to the point of “harry potters cloak” where a person can just throw it on and suddenly they become invisible to run around and get away with whatever they want. It is actually a great thing this is not the case because if that type of technology surfaced to the public in an easy to use way, there would be extreme chaos worldwide. People would use it to commit numerous crimes and acts that would normally land them in jail, so really the current halt in the progression of immediate, easily accessible ways of cloaking is a ‘blessing in disguise’. Instead, all that is currently available is pre thought out and precise uses of this technology and we are aware of the limits it possesses due to many different factors.

Although there may not be many great uses for the general public and society as a whole, there is a limitless use to military applications, like one previously stated. On the other hand, once this technology becomes extremely successful and beneficial, other countries will soon possess the same technology and want to use it against us. Expectantly by then our country will have found a way to counteract those measurements, if they are not already found through experimental failures and errors currently. Once this becomes successful, it can be used to cloak our soldiers fighting for us so that the enemy cannot see them and extremely increasing the survival rates when in a combat situation. If it comes down to a great breakthrough that allows it to be easily made and used but does not work perfectly, it would still be a benefit if it even slightly renders the enemy’s’ vision of the soldiers less useful. If this ever gets to the point where we can use it on an extremely large scale, every fighter jet and helicopter that flies over war zones can be covered and be protected from enemy sight. The enemy may still be aware of their approximate location through radar but it would be very difficult the hit the moving jet without the use of a
heat seeking missile. Another great example, yet on an even larger scale is our fleet of ships in the navy, we could disguise ships as small as a frigate to ships as large as an aircraft carrier. This would be especially helpful in water combat scenarios considering ships can turn off witch would disable the enemies attempt at location through sonar and radar. They would be completely blind as to where our ships are. There is also such thing as acoustic camouflage, and “Since the 1940’s many countries have experimented with sound absorbing coatings to reduce sonar reflection on submarines” (Tack, 2007). Another great example is gun silencers, although you can still hear them, it is significantly less loud which is extremely beneficial for stealth operations. Technology is only getting more advanced so maybe in a few more years both of these will be perfected.

In another article it was found that a product called “Adaptiv is a camouflage system that detects and projects an infrared signal, not visible light; therefore, it has sensors that detect the heat signature of the environment surrounding the target to be camouflaged. Hexagonal panels mounted to sides of a tank or personnel carrier generate heat to both match the background and to disguise the heat signature of the target” (Springer, 2017). This technology will be able to save countless lives and help in many other aspects as well. The fact that we have taken ideas from nature and brought them to life is groundbreaking and it should continue to be done.

Although this technology seems amazing and life changing, in the end it may not be a good life changing if it gets into the wrong hands. In science there are always limits being pushed, and at some point it will break and the goal will either be achieved or lost. At this point it is best to push through that point so if it does in fact achieve the end goal, its capabilities will belong to us first, so that we can better defend ourselves before it becomes more common. Just like everything else in life, there is some debate to whether this type of technology is alright to use in a combat scenario. It has been discussed that “Article 36 of Additional Protocol I (AP I) to the Geneva Conventions, entitled “New Weapons,” requires that “[i]n the study, development, acquisition or adoption of a new weapon, means or method of warfare,” countries must determine whether its employment would, in any or all circumstances, be prohibited by other provisions of AP I or other binding international law” (Shani, 2015). The mans and methods are not technically defined in the article, however it is known that ‘means’ is the type of weapon itself, and ‘method’ is known to mean how that weapon will be used, in what manner. Personally, I think that it comes down to the fact that if we do not achieve this, someone else will and we would regret it down the line. Our enemies do not care about the laws and the rules, and that is not to say that we shouldn’t either. The fact that we have laws and rules makes us more justified to be the first with this technology, because once it is founded, there may be a reasonable way to ensure other countries do not have it. Having laws does not mean we should halt our progression and wait for some other country to obtain the technology first leaving us defenseless and playing catchup. We must push forward, because if not, we may not even have the opportunity to defend ourselves later on. Technology can be like life itself in some ways; outstanding, yet destructive, but constantly moving forward with no regard to wants or needs.
FIGURES

Figure 1 displays the aquatic example where the fish would be blind to the observer once inside the hole. It also shows that the bending of the light rays around the object are longer than the surrounding rays, proving that those rays must travel at a faster speed.

Figure 2 shows the basic cloaking in a terrestrial environment with air as its medium.
Figure 3 shows negative refraction.

Figure 4 is the equation for the perfect invisibility cloak.

\[ f(s) = \frac{a + b + c}{a\alpha + b\beta + c (|1 - \gamma|)} , \]
REFERENCES

1. Chen, Hongsheng. Et al. (June 2013). Natural Light Cloaking for Aquatic and Terrestrial Creatures. [Cited 01 APR 2017]. Available from file:///C:/Users/20hai/Desktop/Spring%202017/Physics%202012%20LAB/source%201.pdf


3. Zhu, F, X. Et al. Harry Potter’s Cloak. [Internet] Department of Physics, Nanjing University. [Cited 01 APR 2017]. Available from file:///C:/Users/20hai/Desktop/Spring%202017/Physics%202012%20LAB/source%203.pdf

4. (March 2014) Effective thermal camouflage and invisibility device for soldiers created. [Internet]. National University of Singapore. [Cited 04 APR 2017]. Available from https://www.sciencedaily.com/releases/2014/03/140311100322.htm


How our biggest fears have evolved over millions of years
Jordan Behm
23 March 2017
Bio181
Dr. Browning
Abstract
Sharks have been around for millions of years and it is known in the science community that this species has evolved drastically. Sharks have been roaming the earth for over 450 million years. In this extensive time period, there are many reasons why sharks evolved as they did. The life of a shark millions of years ago was a lot different than life for a shark in the modern world. This goes for all creatures on earth, because they evolve and adapt to the constant changing world. There are several traits in sharks right now that were not present millions of years ago. These new but ancient creatures still roam the earth and continue to adapt to the ever changing world. Humans have only been in contact with sharks for a tiny part of their existence but have had a huge impact on the way they act and conduct daily life.

Content
The world has been around for a few billion years and a creature that has been around a vast amount of Earth’s time is the Shark. Sharks have been creatures that can date back to over 450 billion years, and have survived through all of Earth’s disasters and still live on today. This is a great feat for anything to stay in existence for that amount of time. Hopefully, with awareness and responsible, sharks can continue to live on this Earth for years to come. There are hundreds if not thousands of species of shark from when they first swam Earth’s oceans. The study of sharks is a very interesting and extraordinary topic. To be able to put together images of what sharks used to look like and make comparisons to how sharks look now is just amazing. In this vast amount of time sharks have spent on this Earth there have been many things that have changed, through evolution and adaptation. This paper will address how sharks have evolved, what has changed through evolution and the history of sharks.

“Sharks come in a huge range of sizes; from the Spined Pygmy Shark, which fits into the palm of your hand, to the Whale Shark which is reported to reach up to 20 meters in length. Surprisingly, only 20% of sharks are larger than 1.8 meters” (SharkTrust). With sharks always changing there are a few things to look at to distinguish between sharks. To start off, sharks might seem like they have bones but indeed they do not. Their bodies are made of cartilage. This allows them to swim faster and have a lighter body, which serves many purposes in the wild. Many fish in the ocean have scales to cover their skin but sharks have skin covered with millions of tiny teeth called dermal denticles. These tiny teeth are pointed backwards to give the shark a more aerodynamic body and allow them to swim faster. The teeth of a shark are literal bigger versions of their dermal denticles, and there are rows of teeth that gradually move forward as the front row of teeth are replaced around every 2 weeks. The Hollywood picture of a dorsal fin in the water is what most people think of when they talk about sharks. They have fins different from most fish. “Bony fish have fins that fold down when they are not being used. Sharks have five pairs of rigid fins, which can't fold down” (SharkTrust). Sharks on average have 5-6 gills on each side of their body and the way sharks breathe and stay alive underwater is by absorbing oxygen in the water up through their gills, so for sharks to have clean and working gills is a life or death situation. Inside the sharks body, it has two types of muscles. Red muscles that allows them to swim long distances at a slow rate and white muscles that are used for when the shark
needs short bursts of energy and speed. Now knowing what makes a shark a shark, seeing them in person will be a little easier to notice.

What we know about sharks in current day is nothing compared to what they were like before they have evolved into the modern day shark. To understand what has evolved on a shark, knowing the past history of what a shark was is very useful. Sharks have not always been massive creatures like today. Sharks used to be no bigger than 4 feet long and had many different body shapes. The stethacanthus is an ancient genus of shark that had a flat dorsal fin. The frilled shark (Chlamydoselachus anguineus) is a modern dinosaur and has a body more like an eel than a shark. Over the years sharks have evolved such that all these different body shapes have started to combine and become a lot more similar. The importance of sharks have changed a lot through evolution. Looking back, 450 million years ago, sharks were mostly small and were similar to modern day fish. But now that they have grown as a species they are at the top of many food chains. Any species at the top of the food chain plays a big role in the ecosystem. Sharks now keep fish populations down and make sure select species don’t become overpopulated and in the end helps out the living fish because there is now less competition. With the growing role of sharks it is key to continue to keep track of how they adapt with such important roles in an ecosystem.

Everything known to man has gone through the process of evolution even if it is the smallest change. Evolution and adaptation of species is very critical to life. “Sharks appear to have evolved from fish between 400 and 450 million years ago, during the Early Devonian period. The earliest known sharks were fairly small and measured just 12 to 16 inches long” (Reference). As know today many species of sharks are a lot bigger than 16 inches, some sharks have been 20x the size of these original species. Over evolution these small sharks have branched out and grown as a species in many different ways. Before evolving into the modern species of sharks, these prehistoric sea creatures evolved from other fish of similar nature. In some cases sharks have evolved in size to better adapt to the giant oceans. Sharks have also developed new body styles such as a hammerhead shark compared to the regular shape of a shark. In prehistoric times sharks have many fins around there body but now sharks normally have 3-4 fins. Around 400 million years ago the mouth of a shark was on the front of its head but in modern day the mouth is more towards the bottom of the body, this helps sharks while eating and hunting. The ability sharks have to adapt to their environment is the sole reason they still exist today.

Evolution is something that all species go through over time to better adapt to situations and increase their chance of survival. An example of evolution would be if someone moves to a foreign country and does not speak the native language. Over time they will adapt to the new environment and eventually learn how to speak the new language and communicate. They way things adapt is different for every species some can do this much faster than other species.

The oldest shark fossil to be found was over 400 million years old; this means sharks have been around during the 5 massive extinction periods that have destroyed prehistoric life. Surviving all these periods is very impressive and shows how well adapted a shark is to its environment. When many species were destroyed sharks had the ability to flourish from the lack of competition around. This is something that helped them recover from such disastrous
situations. Knowing that the oceans used to be filled with more life and used to be a lot more untouched than they are shows how sharks struggled their way to the top of the food chain from surviving and thriving.

Before evolution millions of years ago, sharks were not always at the top of the food chain. There were prehistoric sea creatures that were known to be much bigger than sharks. This shows how sharks have adapted to their environment and were not at the top of the food chain. Some examples of sharks before evolution include the Stethacanthus, this was “a two-foot long shark that lived in warm, shallow seas” (SharkSavers). These sharks were very different looking than any modern shark. A flat topped dorsal fin was used by the shark for protection and mating.

With sharks being so old the oceans they originally lived in had to be different than the oceans swam in today. These very old oceans were filled with a lot more life than oceans today for a few reasons, one being humans were not around to kill millions of creatures a year. Also there were a lot of mammals and fish that lived in the ocean because these species had not been around long enough to evolve and move to land or they thought the ocean was the better place. The times sharks lived in these old waters they were not at the top of the food chain and this means sharks has to struggle everyday to live with these other sea creatures.

Sharks now are a lot more dominant than they were before, seeing something put up a fight with a great white is very rare. Most sharks are carnivores and do not seem to have very much competition for food because of their size and ability to hunt. Not all sharks now are at the top of the food chain but for the most part there are not many other creatures that will put up a good fight with a shark. Sharks eat a wide variety, from small sea creatures to much bigger sea animals like seals.

Sharks have many different traits and some of them have changed over the course of their existence, and some traits have remained the same. Size is something that has changed in sharks and always will continue to change. The Megaledon for instance is the biggest shark known to man, and the hybodus which is nearly the size of a human. Both of these sharks are prehistoric and there are many different modern day sharks that carry similar characteristics and behavior.

Humans have had a huge impact on modern sharks, some negative and some positive. Starting with the negative, humans are known to hunt many species of shark for personal gain. This in no way helps sharks and is affecting the way they live. Humans invade on their homes and in many reports with basking sharks boats have hit and injured these sharks while they were feeding. If humans continue this deadly nature, sharks are going to eventually adapt or go extinct, which they have done many times in the past. Aside from the hunting of sharks, humans have done something’s that have positively affected sharks. This includes groups that help rehabilitate injured sharks and allow them to have a second chance at life in the wild or in sanctuaries. Humans and sharks can coexist if both species respect where they can and cannot go. This is going to be very hard because neither of them knows their limits but it is the only way to peacefully coexist. Before humans existed sharks had to deal with many other predators of the sea but not a lot of these predators have gone extinct and sharks new bigger fear are humans. “In the past, Chinese Emperors favored the soup as a dish that honored guests because it was thought to have medicinal benefits and represented a victory against powerful sharks” (Fairclough). Shark fins are a delicate amenity in the Chinese culture which means they are very hard to get
but worth a lot of money. Shark finning is very popular for people to supply the need for shark fin soup.

Discovering the very first sign of sharks is nearly impossible just on the fact of sharks are made up of different material than normal fossils. Sharks are made up of mostly cartride and this is something that over time deteriorates and is not useable. With many fossils found before sharks they are made of bone so it is a lot easier for people to discover what the creature was because of the intact fossil. Cartilage is good for the shark because it is lighter than bone so the shark is more mobile than say a fish that has real bones. Humans also have cartilage in places like the nose and the ear and this is why when looking at a human skeleton there is nothing in the place of where the nose was and why there is a hole where the ear is normally located.

Prehistoric sharks and modern sharks are not really the same so knowing the difference is very important when talking about how they evolved. The stethacantus is a prehistoric shark that is very oddly shaped with a flat fin on the top of its body instead of a normal streamline fin, This shows that for some reason the shark needed a fin like this for a certain reason and when it didn’t need that flat fin anymore it when back to normal. The megalodon is also another extinct shark that is very interesting the look at. This is one of the biggest species of shark and to give perspective of what really roamed earth’s oceans “ In a Megalodon with a total length of 52 feet (15.9 metres), the first dorsal fin would be over 5.5 feet (1.7 meters) tall, the pectoral fins would each be over 10 feet (3.1 meters) long, and the tail would be over 12 feet (3.8 meters) tall. The girth (maximum diameter) of such a shark would be about 32 feet (9.7 meters)” (elasmo-research). Compared to modern sharks like the Great white which that is ¼ the size of a megalodon they are very small. Even though great whites are small compared to some they are not an ocean creature to mess around with. Some having 7-8 rows of teeth and the swim speeds of 35 mph they are still very deadly. The Hammerhead shark is another type of shark that is oddly shaped, with its face flattening out and looking just like a hammer. These fish have changed over time and for some reason evolved into these hammerheads for a certain purpose to make their lives easier. Knowing the differences of modern sharks and ancients is very important because if you are looking at evolution you must know where the species came from and where it is today.

One big question about sharks is why do they attack people when we are not really directly bothering them. “Sharks attack and kill 10 humans per year, on average. Humans, in contrast, annually kill about 20 to 30 million sharks, according to the Florida Museum of Natural History’s Department of Ichthyology” (Mislinski). This is a terrible ratio and it shows who the real killer is. This shows why sharks might be scared of humans in the first place. Some researchers think that once sharks start tasting and attacking humans they are starting to see humans as food and not visitors. It is also expected that shark attacks on humans were from a misidentification, meaning the shark thought the human was something different. The connection of shark evolution and shark attacks is that now that humans are in contract with sharks a lot more sharks are finally getting used to it and some are okay with it and ignore the human and few actually attack the human for unknown reasons, this shows how sharks as a species reacts to new life in their environment.
Conclusion

In conclusion, shark evolution is something humans have studied for years, not much is known about prehistoric sharks but as everyday passes the code of what sharks used to be is getting closer and closer to becoming cracked open and available for our knowledge. Sharks have changed in many ways from when they roamed the earth millions of years ago, from body shape, to the size of sharks. Sharks that used to exist have helped evolve modern sharks into something we can study on a daily basis. I have gained knowledge on how sharks adapted to their ever-changing environments and what they have done to do this. It is not exactly known why sharks are the way they are to this day but with continued research it will soon all come together.

The topic of shark evolution may seem to be focused only on sharks but it is much more. I can use what I learned from sharks to compare when I am looking at other species because they all have similarities of where they came from. Before this paper I had no idea when sharks were discovered and I really didn’t know much about the ancient species. Learning this has been very cool and has taught me a lot more than just about sharks.

In the future I feel like technology will become more advanced. It may even be possible to discover fossils older than 450 million years. Could it be that we may be able to truly find out how sharks evolved? I also think that sharks will not go extinct for a long time because of how well they have adapted to a deadly area before, living 450 million years and I still think they will be around for much longer as long as their environments in the ocean stay clean and suitable like they always have been. If people’s knowledge went up about sharks and what they do for the world I think that these mass killings of sharks will go down dramatically and sharks will not have to worry so much on humans making them go extinct.
Figures

Large Hybodus compared with a 1.6 meter tall person.

Stethacanthus Fossil

Hammer Head Shark
Citations


Stem Cell Research: Mankind's Next Big Step in Medicine

Cameron Benvenuto

March 23rd 2017

BIO181

Dr. Browning
Throughout history, humans have made incredible advancements in medicine. Just less than one hundred years ago, people would die of strep throat or the flu; sicknesses that are considered trifle and easily treatable in this day and age with antibiotics. The next biggest step for mankind is to find an antidote for diseases that we have yet been able to cure. The medium for this advancement is stem cell research. Stem cells are essentially cells found in the human body that are not designated to a specific function, and can be used to replace other cells in the body. By doing further research on these cells, we have the potential to cure diseases such as heart disease, diabetes, cancer, Parkinson's disease, and the list goes on.

This study will discuss what stem cells are, their functions, what we know of them, and what the future holds for their role in medicine.

**What are Stem Cells?**

Stem cells by definition are undifferentiated cells that maintain the ability to become other specialized cells as well as the ability to renew indefinitely. The human body is consisted of various cells that have specific functions, including: brain cells, muscle cells, bone cells, etc. Stem cells are different however; because they do not carry out any specific function, other than merely evolve into other cells. There are two main types of stem cells: embryonic stem cells and somatic or adult stem cells. Embryonic stem cells are stem cells that form during pregnancy, specifically during the embryonic stage of development. These types of stem cells have the ability to become practically any functioning cell in the human body. Scientists will often extract these cells in a controlled culture and because of their ability to become any cell type, scientists are also able to stimulate and control them to change into a specific type of cell; this process is called directed differentiation.

Somatic stem cells are stem cells that are found underneath the tissue of many functioning organs that have completely developed. Somatic stem cells are not nearly as flexible in their ability to differentiate or specialize as embryonic stem cells and most somatic stem cells will only evolve into the cells of the organ to which they are located by. For instance, somatic stem cells residing in the bone marrow, also known as hematopoietic stem cells, can give rise to all types of blood cells, but no other type of cell (see Figure 1). However, further research suggests that somatic stem cells may still have the potential to differentiate into a wider range of cells beyond their designated tissue.

The ability of a cell to differentiate is called potency; there are five levels to stem cell potency. The widest range of cell differentiation a stem cell can be is totipotent. Totipotent cells are stem cells that can become any type of specialized cell. The only examples of totipotent stem cells are the zygote, the conceived cell consisting of both the male and female reproductive cells (sperm and egg), and the cells produced from the first few rounds of cell division after conception. The next level of cell differentiation consists of pluripotent stem cells. Pluripotent stem cells are stem cells that can become practically any specialized cell, except for the placenta (embryonic sac) or the embryonic membrane. Pluripotent cells include the embryonic cells developed after several rounds of cell division. Multipotent stem cells are stem cells that can only form into a handful of cells, all included in the same family. Examples of multipotent stem cells include cells that can give rise to only a specific group of cells such as different blood cells; they also form blood lineage. Oligopotent stem cells are the next level down, and are very restricted on how they can specialize. They are only able to differentiate into a select amount of cells, such as myeloid or lymphoid cells. Finally,
unipotent stem cells are stem cells that can only become one type of cell and only form into the cells of the organ in which they are associated with. Many would go as far to say that these are not considered stem cells since they are destined to become only one type of cell, and cannot differentiate past the organ that it is associated with. However, since they still possess the ability to self-renew indefinitely, they are still considered stem cells. Furthermore, current knowledge of stem cells states that embryonic stem cells are considered pluripotent while somatic or adult stem cells can be either multipotent, oligopotent, or unipotent.

With this knowledge of stem cells, we can understand that as cells to divide more and become older, they tend to lose their potency. In humans, cell division begins at conception; when the sperm meets the egg forming a zygote. The zygote at this point is totipotent for it has potential to self-renew and give rise to any type of cell. It then begins to divide, producing cells that all contain two genes: CDX2 and OCT 3/4. However, after about 10 rounds of cell division (approximately 4 days after conception), CDX2 is found in all of the cells produced, while OCT 3/4 remains strictly in the inner cell mass (et al Wu). This collection of cells is now called a blastocyst, containing the trophectroderm (outer most cells containing only CDX2) and the inner cell mass (inner cells containing both CDX2 and OCT3/4) (see Figure 2).

The trophectroderm will go on to form the placenta (embryonic sac) as well as the embryonic membrane while the inner cell mass will eventually form the embryo itself. Because the inner cell mass contains both genes, it remains pluripotent and has the potential to become any cell in the body with the exception of the placenta and embryonic membrane. Although these cells have incredible potential to differentiate at this stage in development, it does not last very long. As the embryo begins to develop and cells continue to divide, those cells that did carry out any function, begin to specialize and become other cells, such as red blood cells, neurons, skin cells, etc. When the human has developed past the embryonic state, it no longer contains pluripotent stem cells that have the ability to give rise to any type of cell. Instead, newly developed stem cells now only possess the ability to differentiate into to cells in the region of the body in which they reside. Although there is some current research that suggests that somatic stem cells may have potential to specialize in other tissues, it has been widely accepted that embryonic stem cells are entirely more potent than somatic stem cells.

**Experimentation**

While the potential that stem cells have for the future of medicine is unlimited, the current research and practical use of them is quite limited. Therefore, it is imperative that scientists continue to experiment with stem so that their use may one day be applied to the medical field. Scientists are constantly growing, manipulating, and stimulating stem cells so they can reach a further understanding of their properties. Different tests are used on embryonic stem cells vs somatic stem cells to determine their characteristics.

Embryonic stem cells are contracted from clinics and are frozen to be sent to laboratories for experimentation. Note: embryos are not used from aborted fetuses; only stem cells from a vitro fertilization clinic are tested. The embryonic stem cells are placed in a culture containing a nutrient broth serving as a medium for the cells to live. The culture is often lined with mouse embryonic skin cells to prevent the cells from differentiating. This of
course runs the risk of the transferring of diseases to the human stem cells, which is why scientists have found other ways to prevent the stem cells from differentiating without using the mouse feeder. It is difficult to keep these cells alive, but those that do live will replicate and divide, producing millions of stem cells. The cells that live and replicate for months on end without specializing into other cells are concluded to be fully functioning, pluripotent embryonic stem cells (see Figure 3). These embryonic stem cells by themselves will not differentiate unless they are clumped together to form embryoid bodies. If these embryoid bodies are formed, they can go on to form any cell in the human body. Scientists however, work to stimulate the stem cells so that they can specialize in a certain type cell, not just the cell it spontaneously produces. They do this by either modifying the surface of the culture, chemically modify the nutrient pool in which the stem cells live, insert specific genes in the stem cells, or surround the cells with the type of cells they are trying to produce. Thus far, scientists have developed basic procedures for modifying stem cells to differentiate, but specialized fully functioning cells derived from embryonic stem cells have yet to be produced through human stimulation (see Figure 4).

Somatic stem cells are much more difficult to test and culture, due to the fact that once removed from the human body, they tend not self-renew nearly as often, making the process of creating a large population of stem cells problematic. Therefore, scientists are looking for ways that allow somatic stem cells to keep their ability to self-renew, even once they are placed in a culture.

But just because somatic stem cells are more difficult to produce experimentally does not mean that scientists have not made use of them. Shinya Yamanaka, from Kyoto University in Japan, won a Nobel Prize in 2006 for being the first to reprogram somatic stem cells to an embryonic state, a process by which results in the production of induced pluripotent stem cells (iPSCs). iPSCs are essentially somatic stem cells that regained their potency through human stimulation, after which they mimic all of the functions of embryonic stem cells. They reprogram these cells to express specific genes expressed by embryonic stem cells, so that the cell can function as if it were an embryonic stem cell. By doing this, scientists can “de-specialize” specialized cells, and re-differentiate them into a completely new cell. In addition, iPSCs have the potential to be quite beneficial to the diseases treatment for it would not require the extraction of embryonic cells, and would allow the patient to use their own cells rather than waiting on a donor and running the risk of disease or the cells being rejected by the immune system. Megan Scudellari, a journalist for Nature International Weekly Journal of Science, stated that the Yamanaka’s breakthrough in stem cell research, “promised to be a boon for regenerative medicine: researchers might take a person's skin, blood or other cells, reprogram them into iPS [induced pluripotent stem] cells, and then use those to grow liver cells, neurons or whatever was needed to treat a disease. This personalized therapy would get around the risk of immune rejection, and sidestep the ethical concerns of using cells derived from embryos” (et al Scudellari). However, scientists currently use viruses as a medium for reprogramming somatic stem cells, which has been known to cause cancers in other animals on which iPSCs were tested; hence why scientists are currently looking for non-viral alternatives to reprogramming adult stem cells.

Consistent experimentation on stem cells is essential to creating medicines and drugs. Obtaining a better understanding of stem cell characteristics and discovering new methods to
modify and manipulate stem cells in a manner suitable for human treatment is what scientists are currently working on to further progress the treatment and cure of diseases.

**Current Disease Treatments**

As stated previously, the amount of potential stem cell research has to treat diseases is unlimited. However, the application of stem cells to treating disease is nothing new; in fact, it has been present in the medical field for over forty years. Realistically, the only stem cells used as treatment are hematopoietic stem cells: stem cells that reside in the bone marrow and form all types of blood cells. But since 1968, these hematopoietic stem cells have been used to serve as bone marrow transplants, and in recent years, to treat lymphoma and leukemia. Stem cells have also been used to treat several inherited blood diseases and even diabetes and renal cancer, although these have only been applied to a select number of patients.

When a cancer patient goes through chemotherapy, radiation tends to kill many of the hematopoietic stem cells in the bone marrow, requiring them to be replaced. New stem cells are transplanted into the patient via veins, a process known as engraftment. A stem cell transplant can either be autologous in which the stem cells are harvested from the patient, or allogenic in which they are harvested from a donor.

In an autologous transplant, the patient’s cells are removed from either the blood or bone marrow the body prior to chemotherapy and frozen. Once the radiation therapy is over, the patient’s stem cells are injected back into the patient. This procedure, depending on the condition of the patient, may be conducted a second time, known as an autologous tandem transplant, or double autologous transplant. In a tandem transplant, “the patient gets 2 courses of high-dose chemo, each followed by a transplant of their own stem cells” (et al Shen). However, stem cells are not harvested from the body before each round of chemotherapy, but rather all removed at once before the first round, leaving the patient with a significantly low amount of stem cells. The second transplant is not conducted until after the patient is recovered from the initial transplant.

In an allogenic transplant, the patient is given stem cells from a donor whose stem cells are most genetically like the patient’s; they can be either related or unrelated to the patient. Allogenic transplants are most commonly done with the stem cells of a related family member, most preferably a sibling since their cells are most similar in age development. The procedure itself is just like that of an autologous transplant: stem cells are injected through the vein after chemotherapy; and just like autologous transplants, allogenic transplants can be tandem as well. In rare cases, stem cells in blood from an umbilical cord of a newborn child are extracted and given to the patient. This procedure is uncommon however, since the blood of the umbilical cord does not contain enough stem cells for a transplant, and is most only used on small children, to which it is still a rarity.

Autologous transplants can be beneficial to a cancer patient for it allows them to use their own cells, and eliminate the risk of the transferring of diseases from a donor as well as the risk of the stem cells attacking the body. Additionally, there is no waiting for a donor, and the process can go by much more quickly. On the contrary, an autologous graft has can potentially fail by which the stem cells do not go back into the blood or bone marrow,
although it is not very likely. A greater disadvantage however is the fact that when harvesting stem cells from where the cancer is located, cancerous cells are most likely to be collected along with the stem cells. When the culture is injected back in the body, the immune system may not be able to fight it off, considering it could not do so before. To fix this, doctors will attempt to purge the culture, or kill off as many cancer cells from culture as possible. When exercising this procedure however, many stem cells collected from the body are often killed in the process; so when they are injected back into the body, the patient may possibly obtain an insufficient amount of stem cells. Having a significantly low amount of stem cells is quite dangerous since with less stem cells present, it takes longer for white blood cells to be produced, leading to a weaker immune system and risk of severe bleeding; this is also one reason many doctors are against tandem transplants.

Allogenic transplants for cancer patients can be quite beneficial to killing cancer cells because the stem cells from the donor form its own white blood cells, creating its own immune system. Therefore, the patient receives chemotherapy and new white bloods, both of which can fight off cancer cells. Another advantage is that since the donor is cancer-free, the patient can keep receiving stem cells from the donor, and does not have to worry about deprivation of white blood cells since purging is not required. Nonetheless, there are still a number of downfalls to allogenic transplants. One disadvantage is the risk of disease the donor may carry that the patient’s body may not be able to fight off due to the immunosuppressive medications they are given. Although the consequences are dire, the risk is not too great, considering the patient is check for disease prior of transplantation. However, graft-versus-host disease is very likely. This is a condition in which the patient’s body (the host) is attacked by the donor’s stem cells (the graft). Furthermore, the donor’s stem cells kill off not only the cancer cells, but the patient’s healthy cells too. In addition, there is always a high risk that the patient’s body is not suitable for the donor’s stem cells to live in, resulting in the death of the donor’s stem cells before it reaches the blood or bone marrow.

Both autologous and allogenic transplants have their pros and cons, and the more beneficial of the two is conditional; it all depends on the state of the patient. Some patients even go through a hetero-tandem transplant in which they receive their own stem cells and then a donor’s and vice versa; this is often used on patients suffering from lymphoma and multiple myeloma. Patients can also receive a mini-transplant, a type of allogenic transplant in which they receive a significantly lower amount of chemo/ radiation. This allows for some cancer cells to die, and use the donor’s cells to fight off the rest of them. With a mini-transplant, both the donor’s and the patient’s stem cells are used to fight off the cancer cells, to which the donor’s cells will eventually take over.

Controversy

Even with the amazing breakthroughs in the medical field with stem cells research and therapy, not everyone is on board with them. This is due to the controversy that comes with the use of stem cells, regarding: the use of embryos, reproductive cloning, and human-animal chimeras.

The use of embryos is most likely the largest controversy out of the three. The use of embryos is in conflict with not only people’s personal and religious beliefs, but with the law
as well. This is because extracting embryonic stem cells requires the removal of the inner cell mass of the blastocyst of an embryo, which is against the law in several states. In addition, many people do not believe extracting cells from a blastocyst is morally correct either. Therefore, even if embryonic stem cell research was legal nationally, it may still have trouble receiving the research money as it does today. However, as stated previously, these blastocysts are donated through vitro fertilization, in which couples donate them to stem cell research voluntarily, mostly due to sterility. Moreover, the blastocyst used for embryonic stem cell research is not that of which was once ever in a woman’s body. They are merely donated by couples who cannot reproduce; clinics most likely proceed to terminate these blastocysts if not donated to embryonic stem cell research.

In addition to the use of embryos, reproductive cloning is also a major controversy when it comes to stem cell research. Although stem cell research and cloning are independent of each other, they do share a common factor: nuclear transfer. Nuclear transfer is the process by which blastocysts are created from genetic cloning of somatic stem cells. This process is used to further allogenic stem cell therapy, in which a matching donor’s somatic stem cells are “cloned” by turning a somatic cell into an identical blastocyst by removing the nucleus. A report from the National Academics stated that this practice “is dangerous and may fail” (et al Mandal).

Lastly, there is the controversy of human-animal chimeras. A chimera by definition is a single organism composed of two different zygotes; hence human-animal chimeras are organisms containing both human and animal cells. The first major experiment using human-animal chimeras was in the 1980s when scientists were able to grow an ear on the back of a rat, using human stem cell. This led to much controversy, for the idea of injecting animals with human cells is considered morally incorrect to a great deal of people. However, chimeras are essential to stem cell research because in order to test certain therapies and medications derived from stem cells, they must be ensured safe on animals before any testing is done on humans. However, according to the National Academic’s guideline, both injections of human stem cells in other primates as well as the breeding of chimeras are constrained.

Each one of these controversies is nevertheless important for stem cell research and the advancement of stem cell based therapies. In order to overcome these boundaries, it is imperative that people are correctly informed about the motives of stem cell research, so that it can further be progressed.

**Conclusion**

In conclusion, stem cell research is a key area of science we as a species need to invest in in order to make progress in the medical field, most specifically the treatment of terminal diseases. With stem cells merely being undifferentiated self-renewing cells that all humans obtain, we can conclude that they have the incredible potential to renew harmful or damaged tissue by specializing in the cells that have destructed. We have seen that through much experimentation, scientists are working diligently to understand the functions stem cells carry out and how they can be manipulated to be used to create medications, therapies for diseases that have no cure and very little treatment. We have also seen how stem cells are already being used to help patients with several different diseases and the processes that go
into them. I have come to the idea that stem cell research is essential for fighting life-threatening diseases, considering that the majority of diseases require the renewal or replacement of abnormal cells, stem cell therapies being one of the only possible mediums for this achievement. All in all, stem cell research has provided tremendous breakthroughs in science, and we must overcome the challenges that prevent its progression.
Figures

Figure 1: Hematopoietic stem cell differentiation

Figure 2: Parts of blastocyst

Figure 3: How Human Embryonic Stem Cells Are Derived

Figure 4: Embryonic cell differentiation by human stimulation
References


Hematopoietic Stem Cell Transplants. Encyclopedia of Cancer


http://www.stemcellresearch.org/

http://www.closerlookatstemcells.org/


http://www.nationalstemcellfoundation.org/
To Meat or Not to Meat? That is the Question

Alexandra Bizzarri

Biology 181
Dr. Browning
March 23 2017
Abstract:

The nutritional impact on the human body from a meatless diet can be detrimental if one does not have the needed information to correctly substitute the benefits meat will bring when consumed adequately. There are many classifications of vegetarianism in the world today, however, if one decides to strictly eliminate meat from their diet, then they would be considered a lacto-ovo vegetarian. The decision of removing meat from one’s diet entails both pros and cons. The cons of not eating meat with no effort to retain these sources elsewhere include an insufficient protein intake, a nitrogen imbalance, a vitamin B12 deficiency, cognitive impairment, depleted energy, bone loss, and more. On the other hand, individuals who do not consume meat will lower the probability of food borne illnesses, are at less risk of a heart disease or stroke, and will also avoid unnecessary hormone intake. Meatless diets will not work unless the individual finds their balance of digestible protein, B12 vitamins, and nitrogen intake from other sources. Today, vegetarian diets are easy to maintain in a healthy well-balanced manner given the advanced knowledge and resources available. However, what is difficult to do when one becomes a vegetarian is maintaining appropriate biological functions through the supplemented sources he/she chooses to replace. This is due to the amount of information one needs to know about nutrition in order to correctly substitute what they choose to take away from their body. The human body will make up for what may be lacking in various ways. Some ways are healthy and are suppose to happen, while other ways can be detrimental to one’s health. In brief, the body will begin to excrete particular biochemicals away from certain areas that may or may not be appropriate. Other concerns to consider is how B12 vitamins are not present in plant foods and how nitrogen cannot be directly used from the environment. Therefore, one will need to take into serious consideration on what should be done if one decides to meat or not to meat.

Content:

A meatless diet is defined as eliminating all foods that contain animal flesh. The broad term used in society for this diet is vegetarianism. With the specificity of this diet, an individual can be one of various combinations of vegetarianism. These include but are not limited to: vegan (no animal flesh, or by products of animals), lacto vegetarian (only consumes dairy products), ovo vegetarian (consume egg products) and lacto-ovo vegetarian. A lacto-ovo vegetarian does not consume meat products but does consume eggs and dairy, making it the most common type of vegetarianism.

One might ask why people choose to be vegetarian in the first place? A person can choose to be vegetarian for many different reasons: religion, ethics and food safety being the most common. Some religions prohibit or restrict the consumption of meat or a particular type of meat all together or for ceremonial purposes. For example, certain sectors within Hinduism, Buddhism, and Seventh-Day Adventists forbid the consumption of meat all together. Outside of religious reasons, one could decide to become vegetarian based on their own philosophy of right and wrong. These individuals believe it is morally or ethically wrong to eat animals. Some individuals are simply animal lovers who are not comfortable with eating them; while others feel most modern day animal industry practices are inhumane. If this was the case, vegetarians who still choose to consume dairy will go to a family farm where they believe the animals are treated humanely.
Food safety can be considered another factor to eliminate meat from one's diet. One of the well-known epidemics that struck the nation was mad cow disease. Mad cow disease affects the brain and spinal cord of cows causing holes to form in their brains. These holes come from prions, which are abnormally shaped proteins. Prions influence the good proteins to form into mutated shapes. If a human ingests meat from an infected cow it can cause neurological problems to that human. In the 1980’s, hundreds of people were killed by this catastrophic event. During this time, meat industries wanted to make their cows grow as big as possible in order to keep up with demand of product. They often produce feed filled with grinded up meat from other animals, including leftover meat and bones from previous cows. Scientists believe cattle became infected from eating “feed made with the brains and the spinal cords of other infected cattle” (Thompson et al. 2005). Fortunately due to meat regulations situations like this are highly unlikely. However, diseases created from animal products do still occur in our modern day society.

Aside from the risk of purchasing infected meat, there is also a possibility of heart disease from high cholesterol contents in various cuts of red meat. This is due to the saturated fat content in animal products. Having a diet composed with a high content of saturated fat is known to increase blood cholesterol levels, creating a potential risk for heart disease. Scientifically, vegetarians have shown to have a reduced risk of heart disease. There is also a risk of bone loss due to animal foods containing more methionine and cysteine. These are the sulfur amino acids. When you metabolize sulfur amino acids it makes the blood more acidic. In order to buffer these high acid levels calcium will be pulled from the bones. The body will also find a way to get what it needs. Too much or too little protein intake can cause the excretion of calcium. Scientists have found however, that “adequate intakes of animal and soy protein have been shown to protect bone in middle-aged and older women” (Thompson and Manore 2010). As one can see there are many reasons as to why people decide to become vegetarian. Though there are benefits to one who chooses to become vegetarian, including peace of mind and food safety, there can also be major risks if one does not have the proper educational background to meet all nutritional needs.

In order to comprehend why meat plays such an important role, one must know the value an animal based diet has towards their nutrition and how it affects the body. Biologically speaking, eliminating meat from the diet can affect protein status and deficiencies in vitamins and minerals. If you asked the average adult what the purpose of consuming meat products is for the body their response would most likely contain the word protein in it. Proteins are macronutrients necessary to fulfill everyday activities in the human body. Macronutrients are substances needed in large quantities to sustain life. In order to have a balanced diet one needs to have large amount of macronutrients in their diet. These include fats/lipids, carbohydrates, and proteins. Macronutrients are essential to life because they are needed to create enough energy to continue the functions within your body. The function of a protein depends on the protein’s structure. Each protein molecule consists of chains of amino acids that twist and fold into functional domains. Amino acids are only found in nitrogen, which is why it is so important. Meat is valuable compared to other protein sources because it is the only source that contains nitrogen and all the essential amino acids. This is important because your body needs nitrogen in order to form the proteins that make up your major tissues, as well as your blood and DNA.

Nitrogen is one of the four major elements in the human body and is essential for humans to function because it is abundantly needed in the tissues of the body. These tissues include the skeletal, smooth, and cardiac muscles. To translate, skeletal muscles can be visualized as the layer of muscles shown if our entire epidermis (skin) was peeled back. The smooth muscles can
be found in the intestines, assisting in moving food through the digestive system. Smooth muscles are also found among other places. Lastly, cardiac muscles are found in your heart.

These tissues are what allow our heart to contract and relax as it pumps blood through the circulatory system. The information stated above is relevant because nitrogen (along with carbon, hydrogen, and oxygen) is essential for the production of proteins and amino acids (Ruby 2012), allowing our tissues to function properly. Not only do your various tissues contain nitrogen, your metabolic processes depend upon “enzymes, all of which consist of various kinds of proteins” (Lyons 2008) that use nitrogen and other amino acids for anabolism and catabolism. The nitrogen in amino acids consumed through meat or other protein sources allow our bodies to synthesize, or create, other human proteins for various functions. The nucleic acid DNA, which makes up your genes, and RNA, which is involved in protein synthesis, also contain nitrogen”(Lee 2015). Cells, the very thing that humans are made up of, require nitrogen for normal growth, cell replacement, and tissue repair”(Lee 2015). Nitrogen is also a component of producing energy when amino acids are broken down.

Unfortunately, humans cannot directly use the abundant sources of nitrogen found “primarily in rocks and soils... and the atmosphere”(Forestell 2012). Instead, we have to depend on “microbes and green plants to convert it into form our bodies can use”(Thompson and Manore 2010). Because variable types of meat contain a valuable source of nitrogen, it is important for individuals who do and do not consume meat to make sure they have a proper nitrogen balance as well. There are three types of nitrogen balances one can have, the first being the most desired. The balances include, positive, negative, and equilibrium. A nitrogen balance is essential to maintain for various reasons. For example, if one were trying to build muscle, they would want a positive nitrogen balance because it is the “best situation for muscle growth”(Ruby 2012) to occur. A positive nitrogen balance also means that your body is in an anabolic state. What this means is that “the protein (nitrogen) intake is greater than the nitrogen output”(Ruby 2012) causing a faster workout recovery. A negative nitrogen balance is the most detrimental state for a fitness enthusiast to be in. It is the opposite of a positive nitrogen balance, where the loss is greater than the intake. One's body is considered to be in a catabolic state if they have a negative nitrogen balance. This is because a negative balance “destroys muscles”(Ruby 2012) by pulling out the nitrogen from the muscles and vital organs. This can cause consequential damage to the body as well. How one attains a negative nitrogen balance may result from “consuming an insufficient amount of high-biological value proteins, poor-quality proteins (e.g., lunch meats, fatty meats, and vegetables), or protein sources lacking an optimal balance of the essential amino acids”(Ruby 2012). Lastly, an equilibrium nitrogen balance is one achieved by most individuals who are not considered to have an overly active lifestyle. This occurs when the nitrogen intake and loss is equal. In terms of muscle growth, equilibrium will not appreciate muscle gain. As one can see obtaining nitrogen in a healthy manner is a necessary to a healthy lifestyle. In summary, humans and other animals consume nitrogen in order to “synthesize their own necessary biochemical’s”(Piccoli 2016).

Though the body uses proteins as a source of energy, proteins also contribute to a number of other duties within the body. One of the main bullet points in the cell theory is that all cells come from preexisting cells, and proteins, are what allow cell growth, repair, and maintenance. This is because proteins are constantly being broken down and recycled into new proteins. Our cells are constantly turning over, meaning out with the old and in with the new, and using the preexisting cell. In addition, damaged cells must be repaired in order to maintain a good health. This explains why “constant turnover of proteins from our diet is essential for such
functions” (Campbell et al. 2009). Proteins also act as enzymes. Enzymes can be defined as “a catalyst that act to bind substances together or break them apart, and transform one substance into another” (Campbell et al. 2009). Enzymes help lower activation energy and allow substances to react and do its job, whether it be to make adenosine triphosphate or fuel chemical reactions occurring in the body.

Hormones are another job in which a protein holds. To define, “hormones are substances that act as chemical messengers in the body” (Seeker 2015). These guys essentially tell our body how to respond to changes in the body’s environment. Some hormones made from amino acids include insulin and glucagon. These “play a role in regulating blood glucose levels and controlling the rate at which glucose is used for fuel” (Seeker 2015). Other roles that proteins have within the body are, maintaining fluid and electrolyte balance, transporting nutrients and other substances, controlling our PH level in our blood, and supporting a strong immune system.

As one can see, proteins are essential to all things remaining neutral and constant, but what happens to the proteins we eat? Digestibility affects protein quality. This is important because digestibility is “how efficiently our bodies can digest and absorb a protein” (Thompson and Manroe 2010). The number of amino acids a protein holds will affect its quality. Higher quality proteins contain more essential amino acids. The reason why meat is a favored source of protein is due to animal foods having a high digestibility, allowing us to absorb almost all of their proteins. In turn, this will bring a greater benefit and efficiency to one's body. However, just like everything else in the world, it is all in moderation. Consuming too much or too little will result in health related issues. Refer to figures three, four, and five for more information on adequate food consumption.

According to the American dietetic association, sedentary adults are recommended a protein intake of 0.8g/kg body weight/day. With basic math and a few simple conversions anyone can determine how much protein they need to consume per day. For example if a person weighs 150 pounds and has a sedentary lifestyle how many grams of protein per day would they need? The first step would be to convert the pounds to kilograms by dividing the pounds by 2.2. By doing this you get 150/2.2 = 68.18kg. The next step would be to multiply the new weight in kilograms by 0.8. When we multiply these two numbers we get about 54.54 (68.18 x 0.8 = 54.54) This person would need roughly 54 g of protein per day to sustain a healthy lifestyle. For more information on adults with various lifestyles and protein intakes, reference to figure one.

Now, if one were to decide not to eat meat, their protein would need to be supplemented through various options. An active vegetarian would need to consume 1.3 to 1.5 g/kg body weight/days. The three particular health conditions that reside with animal sources are heart disease, bone loss, and kidney disease. High protein diets composed of predominantly animal meat are associated with higher blood cholesterol levels. This is due to the saturated fat in animal products. Vegetarians have been shown to have a greatly reduced risk of heart disease. However, there is also another option for meat lovers, that is, eating lean meat. This type of meat has less saturated fat and therefore lower cholesterol levels. Since animal foods contain more of the sulfur amino acids including methionine and cysteine, metabolizing these amino acids will make the blood more acidic. In order to buffer these amino acids calcium is pulled from the bone and into the blood.

Preventing vitamin deficiencies are equally as important as maintaining protein levels when eliminating meats from the diet (refer to figure two). Vegetarians are at high risk for vitamin B12 deficiencies specifically. The B vitamins are essential to the human body because they are able to produce the energy necessary to keep us alive. In order to comprehend why B
vitamins are essential to life, one must know what a coenzyme does. Essentially, “coenzymes are molecules that combine with an enzyme to activate it and help it do its job” (Thompson and Manroe). This is important to know because the B vitamins act as a coenzyme, helping the body access the energy in the food we eat. More specifically, B vitamins are water soluble, “allowing the vitamin to get absorbed through the intestinal wall and go directly into our bloodstream” (Saladin 2012). From there, they travel to the needed cells. Because our kidneys act as a filter, our body cannot store large amounts of water soluble vitamins. This means that humans have to consume an adequate amount of B vitamins on a daily or weekly basis.

Individuals who are not vegetarian happen to have a valuable source of vitamin B12. This vitamin is found in meats, fish, poultry, and dairy products. Vitamin B12, also known as cobalamin, is an energy generator as well as a blood booster. Simply put, cobalamin helps convert food into energy as well as optimizes blood health, respectively. In addition to assisting with the formation of blood, it is also required for a healthy functioning nervous system. The nervous system is one of the most important systems our body has. Its importance resides with the fact that it is the commander in chief for our voluntary and involuntary movements. If one has an insufficient amount of vitamin B12, then numerous cognitive and physical complications can result fairly quickly. This happens because cobalamin specifically acts to “maintain the sheath that coats nerve fibers” (Campbell et al. 2009). If the sheath is damaged or absent, this can cause the nerves to fire at inappropriate times. When nerves are not working in unison, as they should, detrimental physical and cognitive effects will take place. Another risk that increases if one does not have adequate levels of vitamin B12 is heart disease or stroke. This surfaces because the body needs the appropriate amount of cobalamin acting as an “enzyme cofactor” (Thompson and Manroe 2010) to break down homocysteine. Homocysteine is an amino acid that requires adequate levels of folate, vitamin B6, and vitamin B12 for it metabolism. High levels of homocysteine in the blood are associated with an increased risk for vascular diseases such as cardiovascular disease” (Thompson and Manroe 2010). As one can see, meat carries an essential vitamin that humans cannot live without. Cobalamin needs to be consumed regularly in order to “avoid diminished energy and exercise tolerance, fatigue, shortness of breath, vascular diseases, and neurological deficiencies” (Gammon et al. 2012). One who chooses a meatless diet will not meet the appropriate level of vitamin B12 intake unless it is consumed through byproducts, fortified cereals, vitamin B12 supplements, or vitamin B12 injections. This varies depending on how strict one’s vegetarian diet is made.

Though vegetarians are at a higher risk of vitamin B12 deficiencies, they happen to avoid an unnecessary hormone intake from not consuming meat. The public has been exposed to the environment of the homes in which our livestock live and grow. To many, seeing the meat industry in raw form is unfavorable because of the way our cattle is handled and fed. It is also the reason as to why some people decide to become vegetarians. Individuals are opposed to the practices of the modern day meat industry because they inject hormones and antibiotics into their cattle. Farmers do this in order to yield more meat per cow, causing a greater profit. Instead of expanding and improving the land and feed to yield more meat, many farmers turn to drugs for the benefit of their wallets. Though the amounts are small, the presence alone has created controversy on whether or not they are harmful. The fact is human antibiotics are being administered to healthy cattle. The main purpose antibiotics hold is to treat infections or diseases caused by bacteria. Therefore, injecting animals that are not sick “with antibiotics can kill off weaker bacteria. Over time, this creates the perfect environment for antibiotic-resistant bacteria to grow and thrive” (Seeker 2015). If the infected meat is consumed without being cooked
properly, then said person could get sick with a bacteria that an antibiotic cannot treat. Hormones in some meat industries are close or behave in a similar manner to anabolic steroids. This turns off people because it is a chemical that does not occur in the body naturally. Because the meat industry is injecting livestock and in turn not preserving natural animal meat, individuals who consume meat need to be aware of the possible risks.

Although vegetarians mainly focus on eliminating all meat products from their diet, some vegetarians will eat animal by product in moderation. In developed countries, one can obtain an adequate amount of protein efficiently from being a vegetarian as long as they have the knowledge and the means to do so. Being a vegetarian is not as easy as it seems because there are a lot of substitutes one needs to make in order to maintain an adequate intake of certain nutrients. A vegetarian who simply decides to cut out meat from their eating habits has to obtain their source of protein from other foods or resources. Abstaining from meat is not necessarily a bad decision if one knows how to supplement or make up for what they are losing from not consuming meat. Individuals who consume a meatless diet need to get their protein, nitrogen, and B12 vitamins from other sources. These resources come from food supplements, vitamin supplements, and balanced meals. There are many great and obvious sources of protein, including meat, dairy products, legumes, soy products, whole grains, and nuts. Though protein from meat is digestible, allowing your body to absorb all the amino acids, there are some legumes that rival meat. This includes soy because it is a complete protein. This means that it has all of the essential amino acids. Soybeans have nearly twice as much protein as any other legume. Some people choose to stay away from soy because of its fat content being extremely high compared to other beans. However, most of this fat is unsaturated, and soy has no cholesterol. Other legumes include kidney beans, pinto beans, black beans, garbanzo beans (also known as chickpeas), lentils, green peas, black-eyed peas, and lima beans. In addition to being excellent sources of protein, legumes are also high in fiber, iron, calcium, and many of the B Vitamins (Thompson and Manroe 2010). However, legumes do not contain vitamin B12, which is another important component found in meat. Legumes other than soy are also deficient in methionine, an essential amino acid. With that being said, legumes are often served with something else, including whole grains or as salad toppings. “Eating legumes regularly may help reduce the risk of heart disease by lowering blood cholesterol levels. Diets high in legumes and soy products are also associated with lower rates of some cancers.

**Conclusion:**

Based on the evidence stated above, I conclude that both sides are correct. It truly comes down to personal opinion; however, scientifically eating meat is a more efficient source of protein because of its digestibility. Rather than taking the extra steps to consume alternative substitutes, eating meat is beneficial if retrieved, cooked, and consumed properly. Individuals who do eat meat need to know more about where their meat is coming from and what that entails. Taking the time to research particular industries that show up in your local market is going to give the individuals a better understanding of what they are consuming. I also recommend limiting the consumption of red meat while consuming more lean meat. The reason being is because lean meat has a lower saturated fat content and therefore lower cholesterol levels. This will lower the risk of heart disease, bone loss, and kidney disease for people who are meat eaters. On the other hand, being vegan is also beneficial if done with precision. Vegetarians commit more time and energy in order to ensure they are consuming a proper, healthy diet. Just
like a simple algebraic equation, what is done to one side has to be done to the other. Meaning, if vegetarians rearrange or rewrite the equation while still holding the original equation to be true, then there is nothing wrong with their decided path; the outcome will remain the same. To meat or not to meat is a preference of opinion. However, it is important to note that too much or too little of such nutrients can and will become detrimental to one's bodily functions.
Figures:

Recommended Protein Intakes: Figure one

<table>
<thead>
<tr>
<th>Groups</th>
<th>Recommended Protein Intake (g/kg body weight/day)*</th>
</tr>
</thead>
<tbody>
<tr>
<td>Most Adults +</td>
<td>0.8</td>
</tr>
<tr>
<td>Non-vegetarian endurance athletes</td>
<td>1.2 to 1.4</td>
</tr>
<tr>
<td>Non-vegetarian strength athletes</td>
<td>1.6 to 1.7</td>
</tr>
<tr>
<td>Vegetarian endurance athletes</td>
<td>1.3 to 1.5</td>
</tr>
<tr>
<td>Vegetarian strength athletes ++</td>
<td>1.7 to 1.8</td>
</tr>
</tbody>
</table>

* To convert body weight to kilograms, divide weight in pounds by 2.2

Figure one depicts the recommended protein intakes for various lifestyles of meat eaters and vegetarians. This is important to ensure the correct amount of protein consumption per individual.

Nutrients of Concern in a Vegan Diet: Figure two

<table>
<thead>
<tr>
<th>Nutrient</th>
<th>Functions</th>
<th>Non-Meat/Non-Dairy Food Sources</th>
</tr>
</thead>
<tbody>
<tr>
<td>Vitamin B12</td>
<td>Assists with DNA synthesis; protection and growth of nerve fibers</td>
<td>Vitamin B12- fortified cereals, yeast, soy products, and other meat analogs; vitamin B12 supplements</td>
</tr>
<tr>
<td>Vitamin D</td>
<td>Promotes bone growth</td>
<td>Vitamin D- fortified cereals, margarines, and soy products; adequate exposure to sunlight; supplementation may be necessary for those who do not get adequate exposure to sunlight</td>
</tr>
<tr>
<td>Riboflavin (Vitamin B2)</td>
<td>Promotes release of energy; supports normal vision and skin health</td>
<td>Whole and enriched grains, green leafy vegetables, mushrooms, beans, nuts, and seeds</td>
</tr>
<tr>
<td>Iron</td>
<td>Assists with oxygen transport; involved in making amino acids and hormones</td>
<td>Whole-grain products, prune juice, dried fruits beans, nuts, seeds, leafy vegetables such as spinach</td>
</tr>
<tr>
<td>Calcium</td>
<td>Maintains bone health; assists with muscle contraction, blood pressure, and nerve transmission</td>
<td>Fortified soy milk and tofu, almonds, dry beans, leafy vegetables, calcium-fortified juices, fortified breakfast cereals</td>
</tr>
<tr>
<td>Zinc</td>
<td>Assists with DNA and RNA synthesis, immune function, and growth</td>
<td>Whole-grain products, wheat germ, beans, nuts, and seeds</td>
</tr>
</tbody>
</table>


Figure two shows the nutrients that could be lost from a meatless diet. In addition, substitutions are included for recommendations.
The Vegetarian Food Guide Pyramid: Figure three

This pyramid is a reference guide for vegetarians to maintain a healthy diet. It depicts which food choice should be consumed weekly, daily, and at every meal.

Protein Content of Commonly Consumed Foods: Figure four

<table>
<thead>
<tr>
<th>Food</th>
<th>Serving Size</th>
<th>Protein (g)</th>
<th></th>
<th>Food</th>
<th>Serving Size</th>
<th>Protein (g)</th>
</tr>
</thead>
<tbody>
<tr>
<td>Beef:</td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Ground, lean, baked (15% fat)</td>
<td>3.0 oz</td>
<td>22</td>
<td></td>
<td>Refried</td>
<td>0.5 cup</td>
<td>7</td>
</tr>
<tr>
<td>Canned beef, brisket, cooked</td>
<td>3.0 oz</td>
<td>15</td>
<td></td>
<td>Kidney, red</td>
<td>0.5 cup</td>
<td>7.7</td>
</tr>
<tr>
<td>Prime rib, broiled (1/2-in. trim)</td>
<td>3.0 oz</td>
<td>17</td>
<td></td>
<td>Black</td>
<td>0.5 cup</td>
<td>7</td>
</tr>
<tr>
<td>Top sirloin, broiled (1/8-in. trim)</td>
<td>3.0 oz</td>
<td>23</td>
<td></td>
<td>Pork and beans, canned</td>
<td>0.5 cup</td>
<td>6.5</td>
</tr>
<tr>
<td>Poultry:</td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Chicken breast, broiled, no skin</td>
<td>3.0 oz</td>
<td>28</td>
<td></td>
<td>Peanuts, dry roasted</td>
<td>1 oz</td>
<td>6.7</td>
</tr>
<tr>
<td>Chicken thigh, roasted, no skin</td>
<td>3.0 oz</td>
<td>23</td>
<td></td>
<td>Peanut butter, creamy</td>
<td>2 tbsp</td>
<td>8</td>
</tr>
<tr>
<td>Chicken drumstick, broiled, with skin</td>
<td>3.0 oz</td>
<td>24</td>
<td></td>
<td>Almonds, blanched</td>
<td>1 oz</td>
<td>6</td>
</tr>
<tr>
<td>Turkey breast, roasted, Louis Rich</td>
<td>3.0 oz</td>
<td>14</td>
<td></td>
<td>Sunflower seeds,</td>
<td>1 oz</td>
<td>5.5</td>
</tr>
<tr>
<td>Turkey, dark meat, roasted, no skin</td>
<td>3.0 oz</td>
<td>26</td>
<td></td>
<td>Pecan halves</td>
<td>1 oz</td>
<td>2.6</td>
</tr>
<tr>
<td>Seafood:</td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Cod, cooked</td>
<td>3.0 oz</td>
<td>10</td>
<td></td>
<td>Barley, cooked</td>
<td>1 cup</td>
<td>3.6</td>
</tr>
<tr>
<td>Salmon, Chinook, baked</td>
<td>3.0 oz</td>
<td>22</td>
<td></td>
<td>Oatmeal, quick instant</td>
<td>1 cup</td>
<td>5.4</td>
</tr>
<tr>
<td>Shrimp, steamed</td>
<td>3.0 oz</td>
<td>18</td>
<td></td>
<td>Cheerios</td>
<td>1 cup</td>
<td>3</td>
</tr>
<tr>
<td>Oysters, steamed</td>
<td>3.0 oz</td>
<td>16</td>
<td></td>
<td>Cereals, Grains, and Brands</td>
<td></td>
<td></td>
</tr>
<tr>
<td>Tuna, in water, drained</td>
<td>3.0 oz</td>
<td>22</td>
<td></td>
<td>Barley, cooked</td>
<td>1 cup</td>
<td>3.4</td>
</tr>
<tr>
<td>Pork:</td>
<td></td>
<td></td>
<td></td>
<td>Rye bread</td>
<td>1 slice</td>
<td>2.7</td>
</tr>
<tr>
<td>Pork loin chop,broiled</td>
<td>3.0 oz</td>
<td>25</td>
<td></td>
<td>Vegetables</td>
<td>1 slice</td>
<td>2.7</td>
</tr>
<tr>
<td>Beef ribs, roasted, lean</td>
<td>3.0 oz</td>
<td>19</td>
<td></td>
<td>Bagel, 3 1/2-in. diameter</td>
<td>1 each</td>
<td>7</td>
</tr>
<tr>
<td>Ham, roasted, lean</td>
<td>3.0 oz</td>
<td>20</td>
<td></td>
<td>Carrots, raw (7.5 x 1 1/8-in.)</td>
<td>1 each</td>
<td>0.7</td>
</tr>
<tr>
<td>Dairy:</td>
<td></td>
<td></td>
<td></td>
<td>Asparagus, boiled</td>
<td>6 spears</td>
<td>2</td>
</tr>
<tr>
<td>Whole milk (3.3% fat)</td>
<td>8 fl. oz.</td>
<td>7.9</td>
<td></td>
<td>Green beans, cooked</td>
<td>1 cup</td>
<td>2.4</td>
</tr>
<tr>
<td>1% milk</td>
<td>8 fl. oz.</td>
<td>8.5</td>
<td></td>
<td>Broccoli, raw, chopped</td>
<td>1 cup</td>
<td>2.6</td>
</tr>
<tr>
<td>Skim milk</td>
<td>8 fl. oz.</td>
<td>8.8</td>
<td></td>
<td>Collards, cooked from frozen</td>
<td>1 cup</td>
<td>5</td>
</tr>
<tr>
<td>Low-fat, plain yogurt</td>
<td>8 fl. oz.</td>
<td>13</td>
<td></td>
<td>Spinach, raw</td>
<td>1 cup</td>
<td>0.9</td>
</tr>
<tr>
<td>American cheese, processed</td>
<td>1 oz</td>
<td>0</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Swiss cheese</td>
<td>1 oz</td>
<td>7.6</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Cottage cheese, low-fat (2%)</td>
<td>1 cup</td>
<td>28</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Soy Products:</td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Tofu</td>
<td>3.3 oz</td>
<td>7</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Tempeh, cooked</td>
<td>3.3 oz</td>
<td>18</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
<tr>
<td>Soy milk beverage</td>
<td>1 cup</td>
<td>11</td>
<td></td>
<td></td>
<td></td>
<td></td>
</tr>
</tbody>
</table>


Having an adequate amount of protein is imperative to a healthy diet. This table shows commonly consumed foods and their protein content.
## Food Groups and Recommended Serving Sizes for Vegetarians: Figure five

<table>
<thead>
<tr>
<th>Food Group</th>
<th>Number of Servings</th>
<th>Foods and Serving Sizes</th>
<th>Calcium-Rich Food Sources</th>
</tr>
</thead>
<tbody>
<tr>
<td>Grains</td>
<td>6</td>
<td>Bread-- 1 slice&lt;br&gt;Cooked grain or cereal-- 1/2 cup&lt;br&gt;Ready-to-eat cereal-- 1 oz</td>
<td>1 oz calcium-fortified breakfast cereal</td>
</tr>
<tr>
<td>Legumes, nuts, and other protein-rich foods</td>
<td>5</td>
<td>cooked beans, peas, or lentils-- 1/2 cup&lt;br&gt;Tofu or tempeh-- 1/2 cup&lt;br&gt;Nut or seed butter-- 2 tbsp.&lt;br&gt;Nuts-- 1/4 cup&lt;br&gt;Meat substitute-- 1 oz&lt;br&gt;Egg-- 1 each</td>
<td>Cow's milk or yogurt or fortified soy milk-- 1/2 cup&lt;br&gt;Cheese-- 3/4 oz&lt;br&gt;Tempeh or calcium-set tofu-- 1/2 cup&lt;br&gt;Almonds-- 1/4 cup&lt;br&gt;Almond butter or sesame tahini-- 2 tbsp.&lt;br&gt;Cooked soybeans-- 1/2 cup&lt;br&gt;Soy nuts-- 1/4 cup</td>
</tr>
<tr>
<td>Vegetables</td>
<td>4</td>
<td>Cooked vegetables-- 1/2 cup&lt;br&gt;Raw vegetables-- 1 cup&lt;br&gt;Vegetable juice-- 1/2 cup</td>
<td>Bok choy, broccoli, collards, Chinese cabbage, kale, mustard greens, or okra-- 1 cup cooked or 2 cups raw&lt;br&gt;Fortified tomato juice-- 1/2 cup</td>
</tr>
<tr>
<td>Fruits</td>
<td>2</td>
<td>Medium fruit-- 1 each&lt;br&gt;Cut up or cooked fruit-- 1/2 cup&lt;br&gt;Fruit juice-- 1/2 cup&lt;br&gt;Dried fruit-- 1/4 cup</td>
<td>Fortified fruit juice-- 1/2 cup&lt;br&gt;Figs-- 5 each</td>
</tr>
<tr>
<td>Fats</td>
<td>2</td>
<td>Oils, mayonnaise, or soft margarine-- 1 tsp.</td>
<td>None</td>
</tr>
</tbody>
</table>


Rather than a broad pyramid, figure five allows the reader to source diet consumption in further detail.
References


The Effects of Apple Cider Vinegar on the Human Body
Lauren M Borucki
April 20, 2017
BIO181
Dr. Darra Browning
Abstract
Apple Cider Vinegar has caught the eyes of many and has grown increasingly popular, due to its remarkable health advantages, in the last few years. Many are aware of the weight loss advantages that ACV provides, but there are many other advantages as well: balancing blood sugar levels, improving cholesterol levels, aiding heartburn, promoting nutrient absorption, preventing gout, etc. There is something about this product that reacts very well with our human bodies, however little research has been clinically conducted. There’s an endless amount of vitamins, minerals, and antioxidants that can lead to these medical miracles, but it is hard to prove what this product does scientifically within our bodies. It has been shown in case studies that diabetics have been able to keep a steady flow of insulin production. Another case study was conducted to show the remarkable weight loss advantages. There is a recommendation for everyone to simply take a tablespoon of ACV a day for a minimum of a month, and watch the remarkable changes which will arise from your body.

Content
In the 1700’s farmers began fermenting apples into cider, after experimentation of further fermentation these farmers found that they could make vinegar as well. This Apple Cider Vinegar that they had created was commonly used for household cleaning, and for medicinal purposes. ACV has come a long way since then, just by reading the bottle that it comes in you can learn about it’s many nutritional benefits (see figure 1). Patricia and Paul Bragg, the co-founders of the most successful ACV selling company, began the BRAGG company in just 1912. Their product includes 2 very simple ingredients: organic apple cider vinegar and water in which they dilute the vinegar to maintain a 5% acidity level. “Miracle health system” it says on the bottle, followed by “internal and external health tonic” and “control weight & banish obesity” They recommend taking three teaspoons of it a day for noticeable changes. These changes could include feeling more energized, having a reduction of acne, healthier skin nails and toes, having a lighter appetite, and much more. Studies and experiments for balancing blood sugar levels with the help of ACV began later on in the 1900’s along with studies on the improvement of cholesterol levels and the reduction of heartburn.

Why does apple cider vinegar react with our bodies the way that it does? What makes it so healthy for our bodies? It has been passed down from generation to generation, from normal household cleaning agent, to health enhancer, to medical miracle initiator. It’s been around for centuries, and the popularity of this product continues to grow. Once the all natural, organic product enters one's body, it is put to work. “An apple a day, keeps the doctor away.” Parents tell their children this, and it is true! Apples contain a high amount of vitamins and minerals and have many nutritional benefits to them (see figure 2). “More and more research has linked consumption of apples and apple products to the reduction or prevention of: Alzheimer’s disease, asthma, breast cancer, colon cancer, heart disease, type II diabetes, liver cancer, [and] digestive cancers” causing apples to be of the healthiest fruits one may indulge in. Apples are fat, sodium,
and cholesterol free, and they contain a high amount of fiber; which is where a large portion of the benefits originate from.

As previously stated, all natural organic apple cider vinegar only contains two ingredients: apples and water. Making all of the nutritional benefits of apples pass down to the vinegar while also producing more benefits. The major difference between apples and apple cider vinegar is apples do not contain “the mother”. “The mother” consists of healthy bacteria, strands of proteins, and enzymes, giving ACV a lot of its major health advantages.

Likely side effects from apple cider vinegar may include: increase in nail strength and hardness, clearer skin, softer & smoother hair, a glow in skin color, and whiter teeth.

The vitamins, minerals and bioflavonoids found in apple cider vinegar are actually antioxidants that have been shown to prevent cancer and also “protect the body against damage caused by exposure to chemical toxins.” There is a fiber called pectin in ACV which absorbs water, cholesterol, fat, and other toxins from the stomach and remove them from the body. The improvement of cholesterol levels, and blood pressure will lead the human heart to a much healthier lifestyle. Insoluble fibers found in ACV help to regulate the digestive system and clean out your stomach and colon. Apple cider vinegar contains a high amount of potassium, “Potassium works together with sodium in the control of the body's water balance, conduction of nerve impulses, contraction of muscles, maintenance of a normal heart rhythm, and it is essential for the storage of carbohydrate and its breakdown for energy.” Finally, the last major nutrient of ACV is the magnesium, which promote teeth and bone health. The magnesium makes ACV good for any arthritis patient or anyone suffer from gout. Jean Van Kleek summed up his experience with gout and the effect ACV had on it with “Have been experiencing the worst and longest gout attack I've ever had on my foot and ankle. It's been two very painful months and several meds that made me feel ill until now. Started drinking ACV three days ago and the swelling is almost gone. Still some pain, but nothing like before.” All of these remarkable benefits are located in just one single substance. Two tablespoons daily does not seem like it would have much of an effect, and while most prescription drugs take weeks to reduce symptoms and often come with side effects, apple cider vinegar begins working right away and the many benefits come quick with no unnecessary harm to one's body.

Apple Cider Vinegar has been in many case studies for its appetite reduction ability. All of the advertising for ACV and it’s weight loss performance is mainly due to the fact the consumption of it leads to a decrease in hunger. “Ingestion of vinegar significantly reduced quantitative and subjective measures of appetite” (Darzi). Darzi conducted an experiment with normal vinegar and apple cider vinegar and found that if taking just simply teaspoons the normal vinegar had little to no effect while the ACV on the other hand gave a much greater effect. The reason ACV does that is “largely due to poor tolerability following ingestion invoking feelings of nausea” (Darzi). Rather than feeling hungry you begin subconsciously feeling nauseous making you uninterested in food.
Diabetics suffer through many side effects with their pancreas not working (see figure 3). One author noted that delayed gastric emptying was a common side effect. “30-50% of diabetes patients have delayed gastric emptying and this is believed to be, at least partially, due to vagal denervation caused by autonomic neuropathy” (Hlebowicz). Damaged stomach muscles or vagal nerves leads to the difficulty of breaking down food, once this effect begins it leads to a diabetics inability to control their glucose levels. When their stomach is not breaking down their food in the correct manner this is what leads diabetics to have random and re-occuring high spikes and low drops in their blood sugar levels, this makes it hard for them to know how much insulin to be pumping into their bodies accurately. “Delayed gastric emptying may cause poor glycaemic control, especially in those with preprandial antidiabetic treatment leading to/causing postprandial hypoglycaemia and gastrointestinal symptoms such as postprandial nausea, vomiting, bloating and early satiety”. Hlebowicz and others conducted a research experiment to see how much apple cider vinegar can effect someone suffer from this problem. This experiment was considered due to vinegars history of improving blood glucose levels. “It is common for patients with diabetes in Sweden to drink vinegar daily because of its positive effect on blood” the authors state. Their study consisted of 10 people (five men and 5 women) all of which were about 40+ years of age. They had each of the subjects eat the same meal, and check their levels a the same time after the meal with the addition of apple cider vinegar. By the end of their study they had inconclusive results, little effect from the apple cider vinegar. There were some flaws in their controls which lead to the possibility of incorrect results, they will revise and reconduct. The authors of this study were aware of the idea that potassium leads to the conduction of nerve impulses, and that diabetics suffer from nerve damage; this could have been a strong relationship between the two however the science behind it is not producing any evidence. Also taking into consideration that ACV aids in the regulation of the digestive system, the likelihood that it also works in a diabetics body seems high, yet could not be backed up with evidence in this study.

Another study, however, was conducted with vinegar infused bread in which the subjects ate, and this study lead to results that proved the increase of vinegar intake had a direct correlation with the reduction of postprandial blood glucose and insulin levels. “... It was shown that white wheat bread served with white vinegar, reduced the postprandial blood glucose and insulin levels, and not only increased but also prolonged satiety in healthy subjects.” If this exact study was done with the substitution of apple cider vinegar instead of white vinegar, it could likely be the same results.

The evidence is in the nutrients and the positive effects on the body from many different dynamics. It fails me that so little research has been conducted on such a remarkable all natural product. The way this product reacts with our human bodies is not just a coincidence, the exact effect it has, if studied, could not only lead to prevention cases but to cures. I believe apple cider vinegar has a very bright future ahead of it, but only as long as researchers continue to look into it. If it has been proven to prevent cancer, then why could I not find a single case study that was
interested in that connection? It is difficult to understand the lack of motivation and research going into this study, but I believe it is growing stronger in the scientific community and future case studies are right around the corner.


Nuclear Waste and Safety Issues Involved in Nuclear Waste Power Byproducts
Abigail Burson
April 20, 2017
General Physics II – PHY112
Dr. Casey Durandet
Abstract

The primary focus of this paper is to explain and discuss nuclear waste and issues related with nuclear waste storage. Nuclear waste is extremely dangerous and can remain in the atmosphere for thousands of years. The toxicity content is so lethal that exposure from even a few meters away can lead to death. A large amount of controversy has risen due to the use of nuclear elements as fuel. Nuclear waste is the product of nuclear fuel reactions that exit a reactor in the process of producing power. The history of nuclear elements and the development of applying nuclear elements for use to humans is summarized. There are six main categories of nuclear waste: uranium mill tailing, transuranic (TRU) waste, low-level waste, intermediate-level waste, high-level waste, and naturally occurring radioactive material (NORM). Details concerning the different types of radioactive decay and emission types will be provided. Additionally this paper will cover the application and derivation of radioactive kinetics. This paper will conclude with the most efficient methods of disposal and storage that we currently know of hazardous nuclear waste storage.

Introduction

Nuclear waste, often called radioactive waste or spent fuel, is hazardously radioactive and can remain so for thousands of years. When the waste leaves the reactor, it is extremely toxic and if anyone is standing a few meters away a deadly radioactive dose is received, leading to death within a couple days [18]. Nuclear waste is a side product from fuel processing plants, research facilities, and even hospitals. This spent fuel can also be accumulated while disassembling nuclear reactors. The two general classifications of waste are high-level and low-level waste. High-level waste refers to spent fuel that is left over from reactors that produce electricity. Low-level waste comes from the commercial applications of radioactive materials, as well as from medical, academic, and industrial uses [3]. Nuclear waste is an ongoing issue that needs to be properly controlled so lethal exposure is not a problem.

What is Nuclear Waste?

Nuclear waste is, in essence, the product of nuclear fuel reactions that exit a reactor in the process of producing power. While in the reactor, nuclear fuels undergo a change in the chemical makeup of the material. This nuclear material enters the reactors as uranium and oxygen atoms in the form of Uranium Oxide (UO2) [18]. The reactor system is then loaded with neutrons where the uranium atoms absorb these neutrons and become unstable. The instability of these particles causes the atoms to undergo a splitting process called fission resulting in two atoms called fission products. These isotopes are lighter and consist of elements like cesium-137 and strontium-90 [18]. This fission creates the energy used to produce electricity but consequentially results in a great deal of heat and radiation. Alternatively the uranium atoms in this nuclear fuel cycle absorbs the neutron but does not split, instead become heavier transuranic elements. These are isotopes such as uranium-239 produces much less radiation but can take thousands of years to decay into nontoxic materials [3].
History of Nuclear Waste

Uranium was first discovered by a German scientist named Martin Klaproth who named the element after Uranus the planet in 1789 [11]. From the 1900’s to the 19402’s scientists including Ernest Rutherford and Niels Bohr furthered our understanding of the atom electrons, and radioactive elements and their spontaneous tendency to emit beta or alpha particles creating different isotopes from one element [11]. In 1938 Otto Hahn and Fritz Strassmann verified atomic fission happening from Uranium. After this Lise Meitner and Otto Frisch calculated that the energy release from the fission of uranium to be approximately 200 million electron volts from 1938 to 1939 [11]. These discoveries lead to the understanding of isotopes such as U-235 and U-238 and nuclear fissions ability to sustain its own chain reaction and create massive amounts of energy [18]. In 1942 this discovery was put into action by Fermi at the University of Chicago [11]. Further research lead to the development of the first atomic bombs, one was made with uranium and the other with plutonium [11]. In August of 1945 these two bombs were dropped in the Japanese cities of Hiroshima and Nagasaki where as many as 240,000 people were killed due to these bombings [11].

Throughout the 1960’s the United States had almost 32,000 nuclear weapons [11]. These accumulation of these nuclear weapons lead to a build of dangerous radioactive waste and attempts by other countries to keep pace with the generation of these weapons through the Cold War. In 1997 the Three Mile Island reactor underwent a malfunction where a limited amount of nuclear waste was emitted [11]. Following this was the devastating Chernobyl accident of 1986 where approximately 4000 innocent bystanders were exposed to the toxic radioactivity and diagnosed with cancer [11]. In more recent years, four reactors of Fukushima Daiichi lost power due to a tsunami resulting in meltdowns of the plant, emission of radioactivity into the air, and hydrogen explosions [11].

Although many casualties have occurred and a great deal of land and natural resources destroyed, not all nuclear inventions have been damaging. For instance, another use of nuclear energy came from Admiral Rickover who successfully created the first nuclear charged submarine. This submarine design was used as a blueprint for making reactors that produce electricity powered from nuclear activity. Also, in 2013 an environmental and climate specialist James Hansan published an article demonstrating how using nuclear energy as an alternate supply of power has saved an estimated 1.8 million lives that would have been killed due to air pollution issues [11].

Types of Nuclear Waste

As previously stated, there are two broad classifications of nuclear waste known as high and low level waste. These classifications of nuclear waste are further divided into six categories: uranium mill tailing, transuranic (TRU) waste, low-level waste, intermediate-level waste, high-level waste, and naturally occurring radioactive materials (NORM). Uranium mill tailings or uranium waste is predominantly composed of the sand-like waste material leftover from a typical uranium mill. Ore residue is left over that contains radioactive decay compounds that come from the U-238 chain and heavy metals [21]. These uranium tailings, proximately located by the mills, are arranged in mounds that are called tailing piles. These tailing piles are most commonly found in the southwestern part of the United States. The most important component of uranium tailings is a highly radioactive element called radium, which can stay radioactive for
thousands of years and decays to produce radon. Gamma radiation is also produced from uranium tailings. The production of radioactive decay can pose a hazardous health issue by contaminating surface and groundwater that can be used for human consumption [22].

Transuranic (TRU) waste, which is only found in the United States, is composed of man-made radioactive elements that have a higher atomic number than uranium (92). The word “transuranic” is used because all the elements that come from this type of waste appear after uranium on the periodic table [19]. Some of the elements that contaminate TRU waste are neptunium, plutonium, and americium, with the most predominant element being plutonium. The two primary sources of TRU waste are from using plutonium to create nuclear weapons and recycling spent fuel [20]. Most of the TRU waste does not release elevated levels of radiation, but it can be a problem when small particles are ingested. The ingestion of small particles can pose a risk to internal organs and lung tissue.

Low-level waste is categorized by four subcategories known as classes A, B, C, and Greater-Than-Class-C (GTCC). The hazardous level ranges from class A, about 95% of all waste, being the least hazardous to GTCC being the most hazardous [5]. Low-level nuclear waste all together makes up about 90% of all radioactive waste. Since this type of waste is predominantly composed of the least hazardous class A material, it is not as dangerous to handle as the other types of nuclear waste. One factor that distinguishes low-level waste from other types is that it contains no alpha emitters, leading to alpha radiation [8]. The low-level waste comes from commercial applications of radioactive materials in academic, medical, and industrial use such as contaminated wiping rags, mops, protective shoe covers, medical tubes, and injection needles. Licensees usually do the disposal of low-level waste on-site with removal consisting of two options. One type of removal being disposing in a regular trash can after the waste has been decayed away and the other type being accumulation of large amounts being shipped away to a low-level waste disposal site in containers that are funded by the Department of Transportation [9].

Another type of nuclear waste, called intermediate-level waste, contains higher amounts of radioactivity than low-level waste. Special shielding is required when handling the waste, which is generated during nuclear power plant operations. Ion exchange resins typically compose intermediate-level waste, which are used to clean the water that disperses throughout the reactor. After a reactor is deactivated, several parts of the reactor are categorized as the intermediate-level waste [7]. Other components that make up intermediate-level waste are chemical sludges and metal fuel cladding. Disposal of this type of waste that are in small quantities or any non-solids may be properly removed once the solidification process is complete in bitumen or concrete [14].

High-level waste is what is left over from the burning of uranium fuel in a reactor. It is highly radioactive so shielding is required and the temperature of the waste gets extremely hot and cooling of the waste is also a necessary precaution. There are short and long-lived components of high-level waste that vary in levels of radioactivity. Short-lived components take a shorter amount of time to decrease the levels of radionuclides to a point that is known as non-hazardous for the environment and the people that inhabit it. High-level waste consists of over 95% of the total radioactivity produced in the generation of electricity. Considering that the only way high-level waste can become non-hazardous is through the process of decay, it is highly alarming that the this type of
waste represents almost all of the nuclear waste that is produced. The process of decay can take up to thousands of years and can remain harmful if they are long-lived. There are two different forms of high-level waste: spent fuel, also known as used fuel, that is accepted for removal and the waste materials that are left over after the used fuel is processed again. It is a primary focus in reference to nuclear power so management and storage of this type of waste is highly important [14].

Naturally occurring radioactive materials (NORM) contains one or more radioactive isotopes called radionuclides that originate in nature. NORM results from activities such as oil and gas production, burning coal, and use of fertilizers. The two types of naturally occurring radioactive waste are known as diffuse and discrete [16]. Diffuse NORM consists of a high volume of waste and a very low concentration of radioactivity. Some examples of diffuse NORM include fly ash and phosphogypsum. Fly ash is produced when coal is burned from fuel and phosphogypsum is what is made from the process of turning phosphate ore into phosphoric acid for fertilizer. Discrete NORM is the opposite, with a high concentration of radioactivity in a small volume of waste. Two examples of discrete NORM are radium, which is used in medical procedures, and aircraft instrument panels [13].

Radioactive Decay and Half Life

As stated previously, radioactive elements such as uranium and plutonium undergo decay until they are no longer hazardous. When dealing with decay of nuclear compounds, the term radionuclide is often used. A radionuclide refers to a radioactive element with a certain amount of protons and neutrons [6]. These isotopes are unstable and give off radiation as they try to regain their stability through radioactive decay. The three main types of decay include alpha, beta, and gamma decay. There are other, less common, ways that these unstable elements can decay such as positron decay and electron capture [6].

In alpha decay an atom ejects a helium “alpha particle” which consists of two protons and two neutrons [4]. This is the largest particle that is released in any radioactive decay. If an atom undergoes alpha decay the net charge is changed to 2+ due to the loss of the two protons in the alpha particle resulting in an entirely new element [4]. Beta decay occurs when a neutron is converted into a proton by the release of one electron [6]. This is due to an excess of neutrons in the nucleus and results in an additional 1+ charge [4]. Gamma decay differs than that of alpha and beta decay as it does not release particle but instead releases gamma rays of electromagnetic radiation [4]. These rays are smaller and can travel farther and faster than other types of decay making them some of the most damaging radiation to humans. The behaviors of these types of emissions can be seen in figure Figure 1. Additionally, the specific materials that can cease these types of radiation are demonstrated in Figure 2.

When referring to radioactivity of nuclear substances half life is an important approach when attempting to understand the kinetics of radioactive decays. Half life is defined by the Linda Hall Library Exhibition as “The time it takes for a given amount of the substance to become reduced by half as a consequence of decay, and therefore, the emission of radiation.” [17]. Half life equations are critical to understanding and predicting the behaviors of nuclear elements for things such as carbon dating, radiotherapy, and cancer treatments. This measure of decay rate is based on probability and does not depend on physical entities such as pressure, chemical state, or temperature.
The half life of elements consists of a huge range in times from seconds to over millions of years depending on the element and its level of instability. For example, the half life of silver-94 is .42 seconds while uranium-235 can have a half life of 713,000,000 years [17].

Decay Rate Kinetics
Since radioactive decay does not rely on physical surrounding factors but on the concentration of the reactant it is a first order reaction [23]. The equation for radioactive decay can comprised based upon the amount of substance, time, and the rate law of first order reactions. The relationship between decay rate and number of atoms can be expressed as:

\[ A = \lambda N \]

Where;
\[ A = \text{Total activity} \]
\[ N = \text{Total number of particles} \]
\[ \lambda = \text{Decay constant} \]

The rate law of first order reactions in differential form can be expressed as [23]:
\[ \frac{dN}{dt} = -\lambda N \]

This can be combined to give
\[ \ln N = -\lambda t C \]

where \( C \) is a constant of integration.

Taking the exponent of both results in
\[ N = e^C e^{-\lambda t} \]
where \( N = e^C = N_0 \) at \( t = 0 \)

Giving the standard form of the decay equation [10]
\[ N = N_0 e^{-\lambda t} \]
Additionally the half life can be calculated and expressed as [10]:
\[ T_{1/2} = \ln(2) / \lambda \]
Or \( T_{1/2} \approx 0.693 / \lambda \).
A graphical representation of half life decay kinetics can be seen in figure 3.

Alternative Disposal and Storage Solutions
As nations continue to generate nuclear waste the amount of high-level waste is accumulating in nuclear power plants across the United States. This spent fuel is temporarily being held at 60 nuclear power plants and is in need of a permanent disposal solution for the wellbeing of people [15]. According to the Nuclear Energy Institute Nuclear power plants generate approximately 2,000 - 2,300 metric tons of waste per year, comprising of 76,430 metric tons of waste in the last four decades [15]. Nuclear fuel now comprise around two thirds of the United States electricity that does not come from a carbon source [2]. With the growing pollution epidemic of our environment nuclear energy is an efficient way to create substantial and reliable energy without polluting the environment. With all factors considered, the use of nuclear compounds as a fuel source should be continued in order to fight the environmental crisis we are in but serious and permanent solutions to storing nuclear by products need to be executed.

Although there are several options that should be considered when implicating how nuclear waste should be stored until it is no longer radioactive, one of the most secure storage is deep geological disposal. This long term storage practice involves
storing the waste at about 500 meters below the earth's surface in formations of the Earth [12]. These formations are stable, strong, and undisturbed consisting of materials including granite, salts, and clay of the Earth [12]. This solution is targeted at creating an environment where the waste can be stored out of human contact and undisturbed for many years or until the substances are no longer emit radiation. This solution to storage has been around for more than 40 years so the technology and designs are advanced enough that building these holding cells is currently reasonable [12]. For even more safety, the geological structure is reinforced with corrosion resistant, impermeable clay, and concrete formations. Due to all these precautionary components the waste should be able to be contained for extremely long periods of time without harmful effects to humans or nature [12]. These deep geological disposals may allow for humankind to continue to advance and grow without creating even more complications for future generations. Combined efforts such as education of the risks, benefits, and technologies and involvement of both the government and its people should allow for the best solution on how to deal with this nuclear waste.

Another possible solution to for extremely toxic radioactive waste could be done through transmutation. Transmutation is done in accelerator driven systems (ADS) and through a process called spallation [2]. “Spallation is the process where nucleons are ejected from a heavy nucleus being hit by a high energy particle.” [2]. The process of transmutation includes long living radioactive elements to be transformed into particles that take less time to degrade. In turn this will greatly reduce the time it takes for radioactive waste to stabilize and become safe [2]. This practice, if successful, will greatly reduce the storage time of radioactive waste. Exposing neutrons in a neutron reactor to radiation can cause these isotopes to experience nuclear fission. This results in the degradation of the original radioactive isotope and produces a fission product [2]. Potentially this process could be performed on nuclear waste and then stored through the process of deep geological disposal. With these combined efforts, the nuclear waste problem could solve our nuclear waste storage problem and in effect help reduce environmental pollution.

Conclusion

Nuclear waste is the material that is left over from the process of producing power in nuclear fuel reactions. The waste is predominantly composed of uranium and oxygen atoms in the form of uranium oxide (UO2). Of the six classifications of nuclear waste, the most dangerous nuclear waste to be aware of is high-level waste. This is because it composes over 95% of the total radioactivity that is generated and is considered highly dangerous. Nuclear waste needs to be controlled properly before exposure becomes a bigger issue.

Nuclear energy may seem like an extreme and frightening way to generate energy. I believe that with education, patience, and further research that nuclear energy can be one of the safest and most productive ways to end the global environmental crisis. Understanding nuclear elements, types of nuclear waste and using the techniques discussed in this paper such as transmutation and deep geological disposal is necessary for success. Nuclear compounds can be used as a great portion of our energy supply for our future while safely sustaining and protecting the environment.
Figures

Figure 1: Demonstration of how radioactive decay behaves
http://www.kentchemistry.com/links/Nuclear/AlphaBetaGamma.htm

Figure 2: Representation of different radioactive decays and their ability to penetrate through different objects

Figure 3: Graph of radioactive decay
http://hyperphysics.phy-astr.gsu.edu/hbase/Nuclear/halfli.html
Figure 4: Relationship of equations concerning radioactive decay kinetics
http://hyperphysics.phy-astr.gsu.edu/hbase/Nuclear/halfli2.html

\[ A = A_0 \left(\frac{-t}{T_{1/2}}\right)^{0.693t} \]

\[ = A_0 e^{-t/T_{1/2}} \]

\[ \frac{T_{1/2}}{\lambda} = \frac{\ln 2}{\lambda} \approx 0.693 \]

Radioactive half-life

Radioactive decay constant

Mean lifetime
References


https://www.tceq.texas.gov/permitting/radmat/uranium/norm.html


https://whatisnuclear.com/articles/waste.html

http://www.hanford.gov/page.cfm/TRU


https://chem.libretexts.org/Core/Physical_and_Theoretical_Chemistry/Nuclear_Chemistry/Radioactivity/Radioactive_Decay_Rates
Neutron Stars

Nick Cioffi
Professor Lori Prause
AST 112
April 12, 2017
Neutron Stars

The universe is an interesting and massive setting because it contains all matter that has been created and destroyed since the beginning of time. When someone says that planet Earth is enormous, just compare the planet to the whole cosmos; Earth and the whole solar system is completely insignificant to the rest of the universe. The cosmos contains billions of galaxies and in each of these galaxies are billions of stars. Stars come in all various sizes and colors; they are the fuel source of light in the universe. Every night, people look up into the dark sky to see the twinkling of light from stars that have been traveling for thousands of light-years to reach Earth. They contain mostly hydrogen and can be categorized into two main groups, low mass and high mass stars. Low mass stars live longer lives, and when one dies, it leaves behind a carbon core called a white dwarf. High mass stars live shorter lives however, and when one dies, it leaves behind a black hole or another celestial object called a neutron star. Neutron stars are interesting phenomena because they are relatively rare and are not as well known by society as other celestial objects. They are one of the great discoveries that lie in the endless cosmos.

A neutron star is a leftover star core in space that forms in a unique way. When a high mass star starts to die, it will run out of its hydrogen fuel and release the remains in an outer layer. Meanwhile, its core starts fusing helium and the process is repeated until it reaches the element iron. The star finally collapses in an extraordinary event called a type II supernova, where the star explodes and releases its outer layers leaving behind the small and dense core that continues to collapse under its own gravity (Redd par. 2). It is remarkable to think that anything could survive such a powerful and violent blast like a supernova. During this event, gravity presses the leftover core so tightly that protons and electrons neutralize until the core becomes a
neutron star (Redd par. 2). The core is now mostly made of neutrons and is massively dense. One theory of a neutron star’s life span is that they last for billions of years while they slowly cool and fade away until they are completely undetectable (Dunbar par. 4). Neutron stars are extremely hard to detect in general, so there is no known explanation of what happens to them after they become undetectable. A neutron star has a more fascinating life compared to a normal star.

Astronomers consider neutron stars to be one of many recent discoveries in the universe. Fritz Zwicky and Walter Baade first proposed the existence of neutron stars in 1933 after studying the Chandrasekhar Limit of dead stars (Egdall par. 5). This news was ahead of its time because neutrons were still a new discovery, so the fact that they existed in space was surprising to many astronomers. In the 1960’s, a student at Cambridge University named Jocelyn Bell was testing a new radio telescope to study the radio waves of distant objects until she came across something strange. During the study, she found many radio signals from celestial objects with periods of approximately one second. She originally thought these unusual signals could be from an extraterrestrial civilization, so she labeled them LGM that stands for little green men. Eventually she and her advisor, Antony Hewish, discovered with additional evidence that these were neutron stars that were transmitting repeating radio signals called pulsars. Hewish later won the Nobel Prize in 1974 for the discovery of pulsars (Palms par. 7). The discovery of these celestial objects is unusual because they were detected unexpectedly, but observing pulsars on radio telescopes is the method that astronomers still use today to detect neutron stars. Neutron stars have a unique background in astronomy.
Neutron stars also contain distinctive properties that separate these objects from others. A neutron star contains a strong magnetic field and is extremely dense for a dead star. When the original high mass star erupts, it ejects most of its magnetic field, but some gets trapped in the core along with plasma. The field and plasma intensify through neutronization that causes the core to become denser (Odenwald par. 1). In fact, a single teaspoon on a neutron star would weigh about a billion tons. Even though neutrons are what give the celestial object its name, they also contain some protons and electrons, but they only make up about one percent or less of the whole body. However it is these protons and electrons that give neutron stars a slight charge and help maintain its magnetic field (Odenwald par. 2). Also due to their size, neutron stars are extremely dim objects and are too faint to be displayed on the Hertzsprung- Russell Diagram; this is why astronomers use radio telescopes to detect pulsars (Palms par. 6). Neutron stars are also extremely heavy despite their small size. These objects are usually about 1.4 solar masses, but can weigh as much as three solar masses (Redd par. 1). Another interesting property is a neutron star’s gravity. The gravity on one of these objects is about two billion times stronger than on Earth and can even bend radiation in a process called gravitational lensing (Redd par. 3). Lastly, neutron stars are able to rotate at insanely fast speeds. These objects can rotate as quickly as 43,000 times per minute while gradually slowing down over time (Redd par. 5). A type II supernova forms each of these distinctive properties of a neutron star, which explains how massively powerful a supernova explosion can become in the universe. Neutron stars have some of the most interesting properties compared to other celestial objects in the universe.

Finally, neutron stars and black holes have certain similarities, but each has their own separate qualities as well. First of all, neutron stars and black holes may form similarly, but there
is a main difference between them that sets the two formations apart. When a neutron star forms, the core can weigh up to just three solar masses, and goes through neutronization, but a black hole can weigh on average 2,000 solar masses, and the entire core completely collapses and is no longer visible (Mosher par. 4). Another difference between the two objects is light source. Neutron stars do give off some light, but it is extremely dim. Black holes contain no light source at all, and if light from a star reaches one, the light cannot escape from it and vanishes (Mosher par. 5). Lastly, the objects are different in size because neutron stars are much smaller than the average black hole. Black holes come in all different sizes, but they can collide together to form a super massive black hole (Mosher par. 5). If two neutron stars collide together, they would form a black hole because the new object would be greater than three solar masses (Mosher par. 5). Although both objects originate from a type II supernova explosion, the objects are two different celestial bodies that lie in the universe.

Neutron stars are one of the most interesting phenomena in the cosmos. They have a unique formation, interesting history and background in astronomy, and specific properties that separate them from other objects including black holes. For an object that is light years away from the Earth, astronomers have done an excellent job studying neutron stars by discovering a tremendous amount of information on their qualities. However, the universe is a vast location filled with an endless amount of questions on neutron stars and other celestial objects that are waiting to be answered. Neutron stars are interestingly unusual, and astronomers will hopefully conduct further research and discover more significant information to explain them to the rest of the world.
Works Cited


Railguns, a Beautiful and Devastating Application of Physics

Johnathon Clark

4/17/2017

Physics 112 General Physics II 11492

Dr. Casey Durandet: B.S., M.S., Ph.D. (High Energy Particle Physics)
Abstract

An Electromagnetic Railgun is the realization of what used to be science fiction, now becoming fact, with the goal of replacing traditional ship mounted chemical propellant heavy weapons with the purely electrical ones, as well as creating a significantly cheaper method of sending items into space. Railguns fire a non-chemically propelled projectile housed inside a moving conductive ‘armature’ in contact between two stationary metal conductive rails through which a large current is applied. This current travels through the ‘north’ rail into the armature, and exiting the ‘south’ rail to complete the circuit; the current moving through the rails produces a magnetic field that applies a Lorentz force to the armature proportional to the amount of current applied, and the length of the rails themselves. The greater the current the greater the force created and by extension the greater speed of the armature housing the intended projectile. The United States Navy has been working with BAE Systems since 2005 to produce a 32 megajoule ship mounted railgun, however there are difficulties overcoming friction, heat, and the power required for each shot. Beyond that of a weapon there are also other theoretical uses for railgun technology such as launching things in to space.

Body

The Electromagnetic Railgun, or EM Railgun, has been the dream of militaries, science fiction nerds, and probably Elon Musk alike for the better part of a century, ever since Andre Louis Octave Fauchon-Villeplee designed the first electric cannon in 1922. The idea of using magnetism and electricity to accelerate a projectile from zero to Mach 7 in milliseconds is equal parts fascinating and ludicrous, however this idea is grounded in solid physics and only limited by how much current we can discharge and how rapidly that current can be made to travel though the railgun. The railgun itself is a simple idea that beautifully makes use of Lorentz Force. Two conducting bars act as the positive and negative ends of an electromagnet, these are the ‘rails’ of the gun. A free moving armature is positioned between the two rails and completes the circuit between them, when a current is applied to the rails the armature is accelerated forward along the rails due to Lorentz Force, which will be explained in greater detail later in this paper, it is important to note that no chemical propellants such as gunpowder or rocket fuel are used in the firing of a Railgun.

The first railgun was designed Andre Louis Octave Fauchon-Villeplee, and patented in the United States in 1922 as an “ELECTRIC APPARATUS FOR PHOPELLING PROJECTILES”, however the one kilometer per second muzzle velocity was not all that jaw dropping for the times and his invention remained as a footnote in history. The first real application of a railgun was created and tested during World War II by the German scientist Joachim Hänsler, who was handed the task by the Office of German Arms. Hänsler created the ‘LM-2’ railgun, which was capable of firing a “aluminum cylinder weighing 10 grams to speeds up to 1,080 m /s”. This design never saw combat due to Germany losing World War II, and the design fell into allied hands who were fascinated at the idea, yet disliked its incredible power requirements; approximately equal to the amount of electricity used to power half of Chicago. The complete history of railguns is for the most part not very well documented, largely due to them not even approaching being useful till recently because of their large power requirements that were infeasible for the time, making them not very interesting to many people; as well as exact current designs are either closely guarded corporate secrets or classified military technology. The current front runners creating railguns for military use are being designed in
parallel by BAE Systems, who worked directly with the United States Office of Naval Research, and General Atomics working independently; each producing their own designs ranging from 3 megajoules to 33 megajoules with the goal of producing a 64 megajoule railgun$^3, 8$.

As previously stated, the main benefit of a Railgun over conventional firing solutions is that the railgun does not use chemical propellants of any kind; making them overall cheaper to fire than using other means such as a guided missile. These missiles could run several million dollars each, as well as present a safety risk while they are stored on a ship waiting to be used. This lack of chemical propellant lowers the risk of having a stockpile of ammunition on board of a ship, as well as makes the whole weapon over all more cost effective per use$^3$. This distinction is important to remember as their efficacy is only in relation to how it can be used instead of conventional weapons.

Currently the primary use of a railgun is purely experimental, the United States Office of Naval Research has been working with BAE Systems since 2005 to produce a ship mounted railgun$^6$. Currently at the end of Phase II of the project, BAE Systems has been refining the 32-megajoule muzzle energy railgun for multi-shot use$^6$. The Office of Naval Research states that “A future weapon system at this energy level would be capable of launching a 100+ nautical mile projectile”$^6$. The listed information for the BAE Systems Railgun will be used for all calculations later in this paper.

The primary physics behind a railguns operation are deceivingly simple, however the application and use present several challenges. Railguns rely on using a very large Lorentz Force to propel the armature along the rails, very large being a slight understatement. The magnetic Force acting on a particle can be given by the equation

\[ F = q \times v \times B \times \sin \theta \]

where \( F \) is the resulting force on the particle, \( q \) is the charge of the particle, \( v \) its velocity, \( B \) the magnitude of the magnetic field acting on the particle, and \( \sin \theta \) being the angle of the particle in relation to the magnetic field. This force equation yields the magnitude of the force itself, however since we are dealing with the motion of something within that field we also want its direction. The direction can be determined simply with the right-hand rule, or can be specifically derived from the angel of the magnetic field lines in relation to the current. Because the magnetic field lines generated by the rails radiate out in a circular path of increasing radii from the rails, and the current flowing through the rails and into the armature meets the magnetic field at a 90° angle, the resulting force vector is propelled out perpendicular to the current flowing through the armature; this is shown graphically in figure 1.

To move from our simplified Magnetic Force equation to the Lorentz Equation we want, we have to consider the amount of current traveling though the rails themselves. Applying the definition of current equaling the charge times time, and that time in this instance equals distance over velocity and after some algebra we get the equation we want of$^2$:

\[ F = I \times \ell \times B \]

Here \( I \) is the current flowing through the rails and the armature itself, \( \ell \) is the length the current is traveling from one rail to the other through the armature, and \( B \) remains the magnitude of the magnetic field. \( \sin \theta \) is not present due to the angle of the current traveling through the armature in relation to the magnetic field being 90°, the sin of 90° being equal to 1. See figure 1.
for a graphic representation of this applied to our railgun. The $\times$ denotes that this is a vector calculation since we want both the magnitude of the force, and the resulting direction of the calculation; the combination of both being the whole purpose of the railgun. Theoretically there is no real limit to how quickly you could make the projectile travel, other than exceeding the speed of light; the more current you apply to the rails, the more force you generate. This use of the Lorentz Force equation greatly simplifies the vast other number of factors going into the operation of a railgun: The shape of the rails themselves affecting the magnetic field shape, friction between the armature and the rails themselves, the resistance of the rails and armature, EM radiation generated, heat production and dissipation, and most importantly the amount of current and how quickly it can be discharged from the capacitor bank and into the rails$^2$.

The complete path of the current flowing through the railgun as well as the capacitor bank itself can also be described as a RLC Circuit, as shown by John P. Hartke at the Monterey California Naval Post Graduate School$^10$. This is depicted individually in figures 6 and 7, and the full circuit shown in figure 8. Figure 6 depicts “The ideal railgun circuit”, showing the path of the current through the circuit to the rails, and back, while figure 7 is a basic design for a simple capacitor bank. Figure 8 is the complete circuit diagram for a 100-kilojoule railgun used by Allan S. Feliciano in his Naval Post Graduate thesis.

Using basic equations of energy and motion, we can approximate the muzzle velocity, and the time to target of the BAE System’s 32 megajoule railgun. This railgun is designed to fire a 23 pound, or 10.4326 kilograms, projectile with a launch energy of 32 megajoules. Given that Kinetic Energy is equal to one half the mass times its velocity squared, we can calculate that the projectile’s velocity at launch is 2476.82 meters per second. That velocity is just a little over Mack 7, or seven times the speed of sound. That greatly exceeds that of a Tomahawk missile’s subsonic speed in flight of 244 meters per second$^{18}$. The time in flight can vary depending on how far the projectile travels, the wind speeds, barometric pressure, and several other ballistic factors, however for the purpose of this paper let’s assume no wind resistance or other factors complicating our calculation and also assume the same 10.4326-kilogram projectile is being fired the previously mentioned desired range by the United States’ Navy of 101 nautical miles. 101 nautical miles is equal to 187052 meters, since the velocity of the projectile is equal to 2476.82 meters per second we can calculate that the projectile will arrive at its target in roughly 76 seconds, our friendly Tomahawk missile would do the same trip in 767 seconds, or about thirteen minutes. BAE Systems lists that its Electromagnetic Railgun can fire ten rounds per minute$^{11}$; therefore, in the same amount of time that granddaddy Tomahawk takes to make one trip we could fire roughly 127 of those projectiles at more than ten times the speed.

Currently the greatest difficulty in the practical use of a railgun is the incredible power required to make it more useful than its chemically powered competitors. While the 32 megajoule BAE Systems railgun is more cost effective per shot, a Tomahawk missile is self-contained, it does not require a massive load of power to ‘spark off’ a shot. For the BAE Systems railgun generating 32 megajoules of energy needs at least 15-30 Megawatts of power to operate$^2$. Considering a single Megawatt can power approximately 164 American households, the small city of 2460-4920 homes that could use the same amount of power puts this power demand into perspective$^9$; using the upper power bound of 30 megajoules and taking into account that the average family supposedly has 2.4 children, for a household size of 4.4, that is roughly the same power consumption the entire city of Nogales Arizona$^{20}$. To preform the Tomahawk missile shaming volley from before of 127 rounds would require approximately 3810 total Megawatts of
power, to put that into perspective using the same process as before; the volley would use roughly the same power consumption of every home in the United States of America combined. Granted both of those figures are very rough estimates, the power need does not scale one-to-one as well as all that power is all not used in the same instant, so these perspectives should be taken with a dump truck filled with salt, however it just shows how much power is needed to accomplish this amazing feat of physics. Figure 2 gives a wonderful visual example of the power of this weapon, you can see the projectile milliseconds after it is launched. The silvery cone shape leading the projectile is the resulting shock wave as it breaks the sound barrier, the fireball surrounding it is the result of the intense heat and pressure from the launch; not any chemical propellant.

To mount this weapon onto a ship would require that ship to produce enormous amounts of power, beyond what it needs just to operate itself. The United States Navy is designing an entirely new class of ships just for the purpose to house railguns, the Zumwalt Class Destroyer. The United States Navy plans to have these weapons deployed on its ships in about ten to fifteen years. This massive discharge of power also presents its own issue of a capacitor bank that can discharge that much power, as rapidly as possible. In addition, railguns are self-destructive. Each shot degrades the rails greatly because of the intense friction, heat, and pressure created with each shot, and a weapon that you can only use five times is not feasible for anything other than a novel science demonstration.

Now that we have thoroughly shown the power and cost of an individual railgun, lets shift focus to the pieces of the railgun itself. As stated before a railgun is made of three main components: the rails, armature, and capacitor bank.

The rails of the railgun present a few interesting challenges, first they need to be very conductive of electricity to minimize the loss of current as well as the time it takes for current to flow through them, however they also need to be very heat resistant. Because no substance is a perfect conductor, some current is always lost as heat. Normally this isn’t the biggest issue for a household appliance, however with the massive amounts of current being applied to the rails in the range of millions of amps, and the friction of the armature sliding along the rails can generate incredible amounts of heat. This heat in addition to the massive amount of energy being used can create a massive ‘fireball’ of plasma as well as pieces of the armature and the rails break off and become gasses, then plasma. Figure 2 shows this effect, the apparent flame trail is not the result of any chemical explosion per say, but the plasma wave as well as other gases igniting from the intense energy being expelled. Due to this massive energy expenditure and the vastly large forces in play for each firing of the railgun, rail degradation is currently one of the largest challenges for railguns. At a minimum, the United States Navy wants to be able to use the weapon one hundred times before they need to replace or repair the rails. Early designs could only withstand a single high powered firing before needing to replace the rails, which puts a damper on its usefulness as a weapon.

Acting as the projectile housing for most railguns, the armature is arguably the second most important component of the railgun, as it is the piece that is doing the acceleration. The armature itself presents the same issues as the rails. The armature needs to be conductive, but even more sturdy than the rails since the armature must withstand the Lorentz force accelerating it, the pressure of the air pushing against it in and outside of the barrel, and must not melt or break under that heat or pressure. These pressures can easily reach hundreds of megapascals.
For most experiments the armature is designed as a ‘Y’ or ‘U’ shape, as shown in figure 3, since this shape is the best combination of conductivity and strength, however as a weapon we would want something more like a traditional bullet\(^{15}\). A combination of both is used for the both the BAE System and General Atomics’ railgun’s payload and armature as shown in figure 4. The same ‘Y’ shape is used however the length of the armature is designed to break away after the whole armature/projectile exits the barrel of the railgun, allowing the much more aerodynamic interior projectile in figure 5 to travel the rest of the way to the target.

Finally, the most important component of an Electromagnetic Railgun, the ‘electro’ part of the operation, the capacitor bank. In actuality you could rip the battery out of your car and hook it up to a homemade railgun, however the battery cannot produce the amount of current needed, or discharge it quickly enough to get a projectile up to Mack 7. For this we need a large capacitor, and a whole lot of them wired in parallel. The General Atomics’ 32 megajoule railgun draws from a capacitor bank the size of a twenty-foot-long shipping container\(^8\). This system is designed to discharge millions of amps as rapidly as possible to generate enough current to produce a large enough Lorentz force\(^{15}\).

Moving away from a human’s ability to destroy something, the massive range and speed of a railgun can also be used for more constructive means; such as cheaply launching items into space. Space flight and transport at the present moment is not exactly cheap, costing around 10,000 USD to put a pound of anything into space according to NASA\(^1\). The fuel cost, the requirement to re-create those non-reusable components, in addition to the maintenance of those parts that we can reuse only serves to keep the cost of space transport high. There have been several advances in technology helping cut costs as well as private ventures such as SpaceX entering the space transport market and their effort to create reusable rockets; which has made space transport slightly more affordable, but still not for daily use\(^{17}\). However amazing and wonderful these advances may be, railguns could potentially make the cost per pound to send things into space even cheaper\(^{13}\). Currently the main issue preventing railguns from undercutting the space transport marketplace ironically is their speed. The escape velocity for Earth is about 11.2 kilometers per second, or slightly below Mack 33; our most powerful railguns can only produce a velocity of Mack 7.5 on its best day\(^6,3,8\). However we can mathematically and literally increase a railguns speed. Because the Lorentz force that is propelling our payload is, on its simplest level, only generated by the product of variables, there is no real limit to what we can scale our railgun to do, other than exceed the speed of light. The first is the most obvious thing we can do is pour even more amps of current into a railgun. Secondly, we could increase the length of the rails themselves. Now mathematically if we increase the strength of \(B\) our magnetic field we could also squeeze more power out of our railgun; however, in reality since the magnetic field is generated because of the current applied to the rails, our first change does this for us.

Another problem to our space cannon railgun is that the designs actively being used and tested by the military are designed, for the military; with their needs in mind. A civilian space railgun has a much different set of needs and would need to utilize a much different design\(^{13}\). The military railgun first and foremost needs to be able to aim at something, having to move a 100,000 ton Nimitz class aircraft carrier is no easy feat; having to rotate it and wait for it to vertically be moved up and down by the ocean just to line up the next shot is about as worthwhile as trying to get information out of a black hole, and ask Steven Hawking how that all works out. Iain R McNab has proposed an alternate design for a space railgun that make use of...
“An externally traveling wave augmented railgun” called the UTSTAR railgun\textsuperscript{13,5}. Before we can understand what that means we have to look into how his design differs from that of BAE Systems or General Atomics. Factoring that the military railgun needs to aim, there is only a single point where the current is applied to the rails, usually near the gun’s breech, this works great until you need to make the rails 1600 meters long. At this length, we have a much greater force yes, however, the loss of current due to the inherent resistance in the rails becomes too great to be useful\textsuperscript{13}. In the UTSTAR design the 1600-meter-long barrel of the railgun is surrounded in augmenting magnets, that only activate when the projectile is near them, acting like a booster, so the current load flowing through the armature and the rails is significantly reduced\textsuperscript{13}. When the UTSTAR design is fired, the first set of magnets at the base of the railgun are energized starting the payloads motion; very similarly to the military design. However, once the payload reaches the next augmenting magnet, that magnet is turned on with a similar power to increase the payloads velocity\textsuperscript{13}. This process continues along the whole length of the barrel. To reiterate, the major difference between the UTSTAR design and the military design is that once the payload is a distance away from the breech where it was loaded, those magnets are powered off, and in sequence with the location of the payload the preceding magnets are powered on, and then off as the payload enters and leaves their area\textsuperscript{13,5}. A visual of this design is shown in figure 9.

The primary benefit of this railgun system is that it is much cheaper to produce, maintain, and use, Ian R. McNab, the creator of the UTSTAR design, estimated that the whole system: railgun, power plant, and other supporting structures, would cost approximately 1.32 Billion dollars to produce, and each load would cost about $528 USD per kilogram sent into orbit\textsuperscript{13}. Which greatly undercut the current going rate\textsuperscript{13}. However, a drawback is that this system does not have any supporting infrastructure in place, on Earth or in space. Theoreticality you could just get a huge cone shaped projectile and fill it full of nothing but chocolate and launch it with the UTSTAR into orbit no problem, however this is the same as closing your eyes throwing a football into a canyon at night. Unless you time the position of either the International Space Station or some other orbiting body, that delicious payload will just drift in space orbiting the earth until it either hits something, or reenters Earth’s atmosphere; which would be a massive waste of chocolate. The obvious solution is to put everything in a satellite-like container that can be maneuvered via conventional means. This is probably what would be done, however that increases operating costs. Another solution would be to use the railgun as the propulsion method for a space elevator connected to something in a geosynchronous orbit with the Earth, but that discussion begins to drift outside the scope of this paper.

From science fiction to science fact, the electromagnetic railgun is on the bleeding edge of weapons and transport technology. The incredibly high speed and acceleration allows for many uses of this technology; from ship to ship combat, air defense, long range land bombardment, or as a method to transport things into space\textsuperscript{6,3,13}. Sadly the technology present today is not quite able to fully realize the demand for reusability and power, however the technology is improving day by day\textsuperscript{8,14,13}. This beautiful application of electromagnetism and Neutonian forces is perhaps one of the most exciting ways physics can be shown in the real world. Personally, I am fascinated by this piece of technology, and elated that I can see something I have only known from science fiction come to life, in a beautifully devastating in an awesome way, just don’t stick your face in front of one when it sparks off.
Figures

Figure 1

![Figure 1](https://upload.wikimedia.org/wikipedia/commons/d/d3/Railgun_usnavy_2008.jpg)

Shows a basic graphical representation of a railgun and its two rails (the gray blocks on top and bottom) and armature (the yellow wedge). The Lorentz Force and how the current (I in blue) and magnetic field (B in green) produce a Force (F in red).

Figure 2

![Figure 2](https://upload.wikimedia.org/wikipedia/commons/d/d3/Railgun_usnavy_2008.jpg)

Shows the US Navy Railgun milliseconds after firing, the tailing fireball is the result of the intense heat pressure and friction creating plasma and igniting gasses as well as pieces of the railgun/armature around the projectile. The silvery cone at leading the projectile is the shockwave from the projectile breaking the sound barrier.

Figure 3\textsuperscript{15}

![Image of armature design](image)

Shows the basic design of an armature used in experiments. The design is called a “C” shape, but honestly it looks more like a Y or U to me. The lower image shows a simulation of the pressures on the armature experienced in launch.

Figure 4\textsuperscript{11}

![Image of BAE Systems armature/projectile design](image)

Shows the BAE Systems Armature/projectile design for its 32 megajoule EM Railgun. The outer blocky part is the armature, using a similar design to figure 3. The interior cone is the actual projectile that is released when after the combo leaves the barrel of the railgun, after which the armature breaks into four pieces.

Figure 5\textsuperscript{3}

![Image of literal projectile](image)

The literal projectile from figure 4. This piece is incased in the armature from figure 5.
This is John P. Hartke’s RLC circuit diagram for a railgun. \( L_o \) and \( L_r \) are inductances, \( R \) is the resistance, \( C \) is the capacitive source, \( E \) is the resulting energy expended.

John P. Hartke’s circuit diagram for a individual capacitor bank.
Allan S. Feliciano’s full circuit diagram for the 100 kilojoule railgun used in his thesis paper.

Figure 9\textsuperscript{13}

Ian R. McNab’s design for a small segment of his UTSTAR design to propel a projectile into orbit.
References


Using Carbon Dating to Retrace Human Kind

Cierra Clement

April 18, 2017

General Chemistry II, CHM 152

Professor Olander
ABSTRACT:
In 1988 one of the most intriguing yet mystifying pieces of historic significance was tested with hopes of placing some amount of fact to the enigma that the Shroud of Turin has posed for centuries. This artifact is one of the most highly debated topics both in scientific and religious communities alike. As questions seem to be the only thing this relic has people agreeing on, many were anxious to use the process of carbon 14 dating to get some answers behind this mystery. Radiocarbon dating, also known as carbon 14 dating, is the use of the isotope carbon 14 found within organic materials to date an object back in time. Due to the radioactive property of the decay of the carbon 14 isotopes it became a means of historic dating and revolutionized the field of archeology in the mid 1900’s. This paper will discuss the chemistry behind the process of radiocarbon dating and the significance that the results have on various fields of study.

Scientists implore various methods when dating things today. These dating methods are most commonly broken into one of two ways, relative and absolute. Relative dating is a way of measuring time based on its relevance around things that have already been given a date. One such dating method that falls into this category is stratigraphy, which utilizes layers of rocks or deposits to determine when something occurs. The idea assumes that the first things to occur or be were first and then proceeding layers came after. (12)

Absolute dating differs from the former method by an obvious reason suggested in the name itself; this method is more absolute. (2) However, many scientist are weary of using the term absolute as it implies a more certain tone than is warranted. This method does give a numerical date for events and things, with there still being a window of error and uncertainty. (12) Within this scope of dating methods there are various ways that scientist go about dating things. It is within this field of study that the method this paper will explore is included.

Absolute dating covers a variety of methods however there are three methods that constitute a large sum of what scientists generally will use when referring to this dating method. The three common forms are tree ring dating, the use of recorded records of civilizations, and radioactive dating. (2) Tree ring dating utilizes the fact that trees such as oaks and evergreens will produce one ring around its trunk every one year. Scientists can use this fact to date trees from different areas of the world and work backward to discover events that were happening at that time. (6) Records of civilization help historians date events and people groups and is very useful as many accounts give actual dates. (7). In addition to these methods scientists have discovered a unique attribute found amongst specific isotopes that allow them to quantifiably date certain materials found on earth today. This dating method is commonly referred to as radiometric dating and it is through this chemical process that radiocarbon dating was discovered.

To understand the science behind radiocarbon dating one must first understand radioactivity within various elements. An element is known to be radioactive if the nucleus of the element is so unstable that it changes itself into another element. This process is known as radioactive decay and can occur in multiple ways. When a given element decays, it emits particles and/or energy from the nucleus of the atom. This was first discovered in the early 1900’s as a joint effort from French physicist’s Henri Becquerel and husband and wife Pierre and Marie Curie. (8)
Ernest Rutherford, a British physicist, went on to continue their research by identifying three types of radioactive emission. Alpha ray was coined as the name for the stream of positive particles that would be released during an element's radioactive decay. Secondly, a beta ray, was named after the negative particle emission and gamma ray was the name given to an emission that did not consist of particles, however was a stream of energy. Rutherford separated these various emissions experimentally by passing them through an electric field. It was at this point that the original element was given the name parent atom and that the element that followed from the parent element's radioactive decay was named the daughter element.

It was through this study of radioactive nuclides that Hans Geiger could find a way to measure radiation. As Rutherford was working on discovering and separating the different radioactive emissions, Geiger was figuring out a way to measure how many emissions were occurring. He first started by filling a tube with a certain gas that ionizes when struck by radiation. A pulse of electricity will generate from the ionized gas particles and this electricity can be converted to an audio signal heard as a click. The number of clicks are counted and this is given as the number of atomic nuclei undergoing radioactive decay. Even today scientists are using this same basic machine for measuring radiation and the device is known as a Geiger Counter.

Upon learning that the amount of radiation coming from a particle could be quantifiably measured, this opened the door for the possibility of radiometric dating. Radiometric dating utilizes the fact that the nuclide of an atom will decrease at a consistent rate. As scientist began to study the decay rates of elements undergoing radioactive decay they noticed a trend that was not by coincidence. They observed that the amount of time required for a sample to decrease by half its original amount was the half-life of the sample, and that this half-life does not change as time goes on.

It is this fact that allows scientist to implore the use of radiometric dating. As they have a type of clock found within these elements that undergo radioactive decay. When these various elements are found within materials that they wish to date it becomes a matter of simply calculating the amount of original element within the material and comparing this value with the amount of daughter nuclides found within the sample. Some elements have varying isotopes, which are different configurations of the neutrons and protons in the nucleus of the element. Carbon has three common known isotopes. These are carbon 12, 13 and 14. Normally carbon would consist of 6 protons, 6 neutrons, and 6 electrons. This would be the makeup inside the most common found version of carbon, which is carbon 12. Neutrons act as the glue that keeps the protons and electrons together. However, found within the other two known “versions” of carbon are a varying number of neutrons, those being 7 and 8 respectively. These isotopes still have the same number of protons and electrons in their nucleus, however the amount of “glue” that is holding them together is different.

Both carbon 12 and carbon 13 are stable isotopes, they don’t experience any change and will remain as they are. However, carbon 14 is unstable, making it what we call a radioactive isotope. This isotope will change and turn into another element over time. The way in which this
happens is that inside the isotope’s nucleus one of the neutrons will turn into a proton, emitting an electron in the process. This emission of an electron is what makes it radioactive, as it is a form of energy and commonly referred to as beta decay. (10) A chart showing this process in more detail can be found by looking at Figure 1 located in the Figures section at the end of this paper. Now the carbon has gained a proton and in the process, changed the very nature of itself into a new element. That new element being nitrogen 14. (3)

This process of carbon 14 “turning into” nitrogen is the very reason that scientists can use this isotope to date certain things. Just as mentioned early carbon can come in various forms. Within all living things there exists certain amount of carbon consumed from either simply breathing it in or by the means of photosynthesis. When plants consume carbon during the process of photosynthesis it is found in the plant in two forms, carbon 12 and carbon 14. Fast forward years down the road when the plant has experienced a lack of water to the point of it dying. (4) At the time that the plant dies and stops ingesting more carbon into its system the amount of carbon 12 and carbon 14 in the plant is the same. However, as the plant begins to decay the carbon 14 will begin the process that was discussed earlier, its radioactive decay. The carbon 14 neutrons will begin to turn into protons and start changing the carbon 14 into nitrogen. The rate at which the carbon 14 decays into nitrogen 14 is known as the decay rate. (13) This decay rate has what scientists call a half-life. A half-life is the amount of time that it takes for half of any quantity of a certain isotope to decay. Each individual isotope has a fixed half-life and is usually an exponential. A graph of this general decay rate is given in the Figures section under the graph labeled Figure 4.

This half-life is what scientists use as a clock for their dating method. Since the original amount of carbon is preserved in the form of carbon 12, scientist can compare the amount of original carbon with the amount of carbon 14 that is left at the time that it is being investigated. By determining the amounts of both forms of carbon experts can use the ratio of carbon 12 to carbon 14 and determine how many half-lives took place within the time frame between the material dying and present day. (9) This is how archeologists date many of the findings. However early on one can see how this form of dating is limited to organic materials, or materials that had been living at one point.

Carbon dating has been used by scientists all over the world to date a variety of things. The use of this dating method can range from forensics at a crime scene to better understanding the cycle of draughts that California has experienced. (9) One field that relies on the use of carbon 14 dating most heavily is the study of our own human history and the field of archeology. Radiocarbon dating can be used on any organic material, which means that it can include in its scope the remains of human or animal bones, old scrolls which that are written on plant-based material, and has even been used to date charcoal from ancient campsites.

The discovery of carbon 14 dating was a revolution for the field of archeology and many would say is the most common and widely used dating method today. (13). First discovered as a means of dating in 1940 by Willard Libby this new-found method was researched and perfected extensively over the last half of the 20th century. (16) Scientist explored what worked and what didn’t work, how effective and precise the dating method was and the extent that they could use this new tool to piece together the history of our world.
What they have found is rather extraordinary. Since carbon 14 has a smaller half-life than much of the radioactive isotopes that were already being used to date rocks and materials, this isotope opened the opportunity for scientists to date things that have occurred at a point in our human timeline much closer to current time than, say, the beginning of earth itself. Carbon 14 was found to have a half-life of 5730 years, allowing it to be used to date materials as old as 60,000 years old. (13) As mentioned before when scientists first discovered the use of carbon 14 as a means of dating it revolutionized the science in the field of archeology and gave rise to renewed fascination in relics that had become a mystery to historians for centuries in the past. One such relic was the infamous Shroud of Turin.

As science as evolved over the centuries there have been more and more a rise in the conflict between the sciences and the people of faith. However, there is no doubt that one of the most controversial has surrounded a garment known as the Shroud of Turin. This 53-square foot piece of cloth bears the image of a man that had been tortured, leading many to believe that this cloth was used to wrap the body of the crucified Christ nearly 2000 years ago. At first glance the naked eye can only see the faint outline of a human body, however in 1898 an Italian photographer made a discovery that sent a renewed interest over the cloth around the world. When Secondo Pia analyzed the reverse negative of his photograph in the darkroom what he found was unmistakably the image of a bearded man, bearing the marks of crucifixion and beatings. (14)

The history of the Shroud first began in the 1300’s and was passed throughout Europe until it ended up in Turin, Italy owned by the House of Savoy. It was here that the photograph was taken by Secondo Pia and the world become reinterested in the authenticity of the Shroud. (14) The mystery is not only wrapped up in who the image of the Shroud depicts but how it got there in the first place. For many the possibility of the image bearing the outline of the Messiah Himself is far-fetched, however it still begs the question how did it get there? This is the question that scientist have been attempting to answer ever since the negative image of cloth was taken at the end of the 19th century. It is here that the usefulness of carbon dating became a possible answer to the question that had many at a loss.

As the use of radiocarbon dating became more and more useful in the mid 1900’s many scientists were pushing for shroud to be testing, as knowing the date of the objects origins could shed much light on the history behind the fiercely debated question of authenticity. However, when the carbon dating was first introduced in the science community the amount of material needed to test was so great that the Catholic church would not even consider the shroud to be given to have tests performed on it. However, when the use of accelerator mass spectrometry was introduced in the late 1970’s it allowed for minute sized amounts of material to be tested opening up the door for the testing of the shroud. (13)

Now that the date of the Shroud could be tested scientists and scholars alike were anxious to get the approval of the Vatican to test the ancient relic and put to rest the quarrelling that had erupted from this piece of unknown history. The Vatican agreed to have a small piece of the cloth used for carbon 14 testing and it was sent to three separate testing sites, Arizona, Oxford, and Zurich. (14) Now historians, regardless of their view on the Shrouds authentic were finally
going to be able to arrive at some conclusion about the history of this relic, that being the date that it was created.

What scientists found was a large percent of original carbon 14 in the material, implying that Shroud was not very old. Scientists then used this amount of carbon 14 to trace backwards, using the isotopes half-life to determine an actual quantifiable date for the cloth. Each testing site conclusively found that the piece of cloth was about 660 years old, putting the date of the cloth at around 1300 A.D. Interestingly this is around the same time that the Shroud surfaced in history. However, this date would mean that the notion that the image bears the imprint of Jesus of Nazareth is completely out of the question, as his death would have been at the beginning of the first century, needing the cloth to have been around 2000 years old.

Since that test was done in 1988 many scientific critics have disputed the date given by the carbon 14 testing that was done. Many explanations have been offered as to why the test could have been erred and what could have gone wrong in the testing procedure for the date to be given for the cloth to have been skewed. The validity of radiocarbon dating was called into question and many different theories were made as to why the date could have been wrong and where scientist could have gone both in the lab and in their assumptions about the original amount of carbon in the cloth.

The process of radiocarbon dating relies heavily on being able to know how much original carbon was in the sample. The explanation of how scientist came to this conclusion dates all the way back to the original man who discovered that carbon 14 could be used to date organic material. Willard Libby first uncovered the fact that carbon 14 was radioactive at the University of Chicago about 1950. He and a team of chemists had found that carbon 14 was produced when cosmic rays, high intensity atoms that hit Earth’s atmosphere at almost the speed of light. Although much about these rays are still a mystery to scientists, Libby could conclude that when these rays hit Earth’s atmosphere they reacted with nitrogen to make the radioactive carbon. This different form of carbon was then converted into CO2 and taken up in the process of photosynthesis. (13)

As plants breathed in the masked form of radioactive carbon in the process of photosynthesis, they now became carriers of this “timeclock” that would later be used. This process is depicted in the figures section labeled Figure 3. With carbon 14 now being in an integral part of the earth’s ecosystem it wouldn’t be long before this radioactive element would be ingested by animals and then later, humans. This process can answer the question of how the radioactive timer became a part of all organic material, however the question remains, how did Libby know the amount of carbon 14 that was set as the “original” amount when compared to in the testing experiments.

As mentioned earlier some materials have sources of carbon 12 and carbon 14, and it is by measuring these two quantities and the difference between them that a date can be found. This method works great with certain materials, such as wood, which can retain its original amount of carbon so well that it can be extracted and measured. (9) However, this was not the method that was first discovered by Willard Libby. His answer was hidden in the fact that it is only through the atmosphere that the plants can inhale any of the carbon 14. Libby would figure and base
much of the whole process of radiocarbon dating on the assumption that the only source of carbon 14 was through the photosynthesis process, and that being from earth’s atmosphere. Based on this assumption, if the amount of carbon 14 in the atmosphere could be tested and found, then that same number would have to be the amount of carbon that would have been in any plant, animal, or human at the time right before it died, supplying him with the original amount of carbon 14. As means of example, a grazing animal would ingest carbon 14 from the plant that had received it from the process of photosynthesis. (13) When the animal died, the carbon 14 would begin to decay, since it is unstable. However, because the animal will no longer have any means of gaining anymore carbon 14 once the animal has died, this point will begin the start of the timer that scientists will later use as the beginning of the decay process of carbon 14.

This leads to the first reasoning behind the critics of the 1988 carbon dating tests. Many believed that the assumption that the amount of carbon 14 in the atmosphere may not be the same value as the initial amount of carbon 14 in the Shroud. If the Shroud was from an ancient time, perhaps around the time of Christ which would have been close to two thousand years ago, the amount of carbon in the atmosphere could have changed from now until then. (11) Libby could prove that the amount of carbon 14 is consistently even in the atmosphere all around the world. As this is sufficient evidence to prove that one could test anything form any part of the world, critics will argue that this is not sufficient evidence to prove that an artifact from a different time in history will have had this same amount of carbon influence from the atmosphere, as atmospheric levels of radioactive carbon could have changed since say, 30 A.D.

Along with this assumption, it must be taken that all living things will have the same amount of carbon 14 as that of the atmosphere. Libby and his colleagues made this assumption based on the idea that there would be no other source for the individual living organisms to intake any amount of radiocarbon other than that from the atmospheres CO2 levels. This assumption was not based on blind ideas. Almost 99% of the world's naturally occurring carbon isotope is found to be in the form of Carbon 12, with little over 1% to be found as the stable isotope of carbon 13. This leaves a miniscule amount to be left in the form of carbon 14. Figure 4 found in the Figures section can give a visual representation of this ration. Carbon 14 is so small that it only forms one trillionth of all modern carbon, to put this in perspective western civilianization has not even existed for one trillion seconds. (13) So, the assumption that it would only be from the atmosphere that all organic materials would gain their levels of carbon 14 was and is still very reasonable of an assumption, however an assumption it is.

A second discrepancy that commentators suggested was that the final amount of carbon 14 had been altered. (11) This critique rests and calls to questions the accuracy of the dating method due to the third assumption that is made within the process of carbon dating. This third assumption being that the remaining amount of carbon 14 in the material being tested is only there as a result from the decay of the original amount left over from the individual’s life. To better understand the critique let the case of the Shroud be examined. As scientists took a small piece of the material to be taken to the lab to be tested they measured the amount of C14 that was found within that one small piece. However, critics would argue that if this piece of the cloth could have been altered over the case of the cloths recent history to contain more C14 than that which would have remained from the original amount at the point of harvesting to make the cloth itself. (13)
Commentators suggested that bacteria could have attached itself to the cloth. Bacteria being a living thing, will retain a certain amount of C14, thus if this bacterium was included in the dating process could be a reason for why such high levels of C14 were found in the dated sample. Another possible source of outside C14 was suggested to have come from the restoration process that the cloth has undertaken in its recent history. Since the Shroud had become a phenomenon and a renewed interest was in taken to it not only from religious leaders and from the public and science community, efforts has been taken in recent years to restore the cloth to a prime condition. Commentators suggested that new amounts of C14 could have been added to the Shroud during one such restoration, again skewing the results of the carbon dating technique to point toward a younger age than was true.

Radiometric dating is one of the most highly used and highly reliable methods that we have today for dating materials not just on planet earth but our universe itself. Since the discovery of radioactive elements in the late 1800’s, scientists have advanced in their use of them to the point of being able to use them and their decay properties to follow time backwards and give an actual qualitative date for our earth and objects on earth. One of the radioactive elements that is used in this dating process is carbon 14. Unlike other radioactive isotopes, carbon 14’s half-life is shorter, allowing scientists to use this element to calculate accurate dates on objects that have shorter histories compared to that of Earth of our universe.

Since uncovering the reliability and accuracy of carbon 14 dating in the mid-20th century scientists and archeologists alike have been experimenting and attempting to put answers to the some of the big questions about our past as human kind and the history that we have had on earth. Many historic periods of time and ancient artifacts have been carbon dated and much has been learned through this chemical process. The Shroud of Turin, a cloth with the image of a crucified man imprinted on it, is one such item that historians and scientists were very anxious to carbon date in hopes of shedding some light on where this cloth could have come from and who’s image is imprinted on it. Too many this cloth holds the image of Christ the Messiah; however, many doubt the Shrouds authenticity, creating controversy between the religious and scientific communities.

It was in 1988 that scientists were finally given permission to date the relic, however upon reaching a conclusive date the answer did nothing to help calm the quarreling that had been surrounding this enigma. The date given would have placed the relics originating at around 1300 A.D, far too early to be from the era of Christ. The reliability of the radiocarbon dating process was almost immediately called into question from many and has since been disputing. This leaves many to wonder at how accurate this chemical process in attempting to uncover the history of the past. As we have seen, there are multiple assumptions when applying the process of carbon dating to an item. I have concluded that while there is sound and thorough evidence for this chemical process, it is still reasonable to assume that there can be errors found within the results of this dating method.
Figure 1 depicts the actual molecular change in the carbon atom when it undergoes decay.

http://upload.wikimedia.org/wikipedia/commons/thumb/7/70/Carbon_14_formation_and_decay.svg/800px-Carbon_14_formation_and_decay.svg.png CC BY-SA.

Figure 2 shows the relative amount of the different forms of naturally occurring carbon isotopes in our atmosphere.

https://www.esrl.noaa.gov/gmd/outreach/isotopes/chemistry.html
Figure 3

The figure above shows the process that carbon 14 goes through within our atmosphere and ecosystem, showing how it is later found in all organic material.


Figure 4

The table above shows the very basics of how the half-life of an isotope declines. As a certain fixed amount of time goes on, half of the element is converted into another form. As this fixed time does by a second time, another half of the element converts.

References

Mars: A Brave New World

Ben Conrad
Astronomy 101
March 22, 2017
Professor Prause
Mars: A Brave New World

Orbiting just 225 million kilometers from Earth, Mars has long fascinated and enthralled mankind. Because of its proximity to the earth and its bright red color it has been sighted since 2000 B.C. Most ancient cultures, including the Romans, Egyptians and Chinese included it in their pantheons. (“Early Times”). The Planet is commonly referenced in literature; in fact it’s a plot point in Jonathon Swifts “Gulliver’s Travels”, H.G Wells “A War of the Worlds” and Edgar Rice Burroughs “A Princess of Mars”. As our knowledge of the universe has grown, our focus has not shifted; since the year 2000 there have been more films involving Mars than the rest of the solar system put together. (“List of Films Set on Mars”). While its apparent magnitude and propinquity have made it so prevalent to the ancient world, it is featured so much in the modern era because as the only other planet in the Goldilocks Zone it is the most likely to be settled, and as such represents a Nuevo Mundo, or a new world. Mars small size has resulted in a weak magnetic field and a thin atmosphere which have made the planet’s surface alien and toxic to humans, with no liquid water, lethal amounts of radiation and an atmosphere which is composed of mostly carbon dioxide. Despite these differences the planet is still the most similar to Earth; while it’s only about half the size of Earth, because it lacks any oceans it has approximately the same landmass. Like earth it once held liquid water, and its average temperature is the closest to Earths, at only a 50-degree average variance; for comparison the next closest planet Jupiter; is almost 250 degrees colder than the average temperature of Earth. (Car et al). For these reasons the planet is the most likely to colonized, standing as a “city on a hill”, a monument to humankind’s eventual expansion into the stars. Granted, such an undertaking would encounter several obstacles: the radiation level is lethal, supplying water remains difficult, and the weaker gravity could cause lasting harm to cosmonauts. With that
being said, landing on Mars does have some advantages; the soil can sustain crops and oxygen can be extracted from the poles. In any case mankind will eventually have no choice; by the end of the century the population is expected to peak to 12 billion, as the planet becomes more and more polluted we must research a way to colonize other planets (“Gerland et al”). As the closest planet, the forces which shaped Mars made it hostile but malleable to human habitation, and as such represents a way to escape the constraints of the earth.

Mars is similar to Earth in that both were formed by accretion, the process where gravity pulls pieces of soil together until it finally becomes large enough to constitute a planet. It is commonly believed, but not proven that Mars was formed far enough from the Sun that it was subject to the gravitational pull of Jupiter (“Why is Mars so Tiny? Blame Jupiter”). This caused it to form at approximately 40% of the size of Earth, and this small size means that it can’t keep a large atmosphere. The core is too big to form a strong active magnetic field, both factors contributed to many of the differences between it and Earth. The atmosphere was once stronger, but the weak magnetic field was unable to protect the planet from the solar wind, and the atmosphere was gradually eroded (“Kenneth Chang”). Without an atmosphere, Mars turned into the wasteland we know it for now, as the temperature was lowered to below freezing, the solar wind irradiated the surface and the planets weak gravitational field was not able to hold on to liquid water.

When the atmosphere was eroded, the planet was eventually covered in radiation, having an average amount of 300 mSv or as much radiation as an average human receives in a year. If improperly prepared any colonists could potentially get cancer, or radiation sickness, which entails dizziness, fatigue hair loss and potentially death (“Symptoms”). As disastrous as this sounds, it’s not an insurmountable problem. Several ways to protect cosmonauts have been
proposed. The problem has been to ensure that there is enough material to shield a settlement while still being light enough to be transported. NASA has proposed using hydrogen for shielding. Since this element is found in water, it would have to be brought to the planet or extracted from the planet. The only problem would be that it would not be sufficient protection from solar flares. To that end, a radiation shelter would be needed in case of a solar flare, which will still be more cost effective than using heavier materials to shield the entire site (Frazier).

Like the radiation, the smaller gravitation appears to be more of an obstacle than it really is. As human’s biology is designed for the earth it was worried that the human body would be unable to function and astronauts would have trouble eating because the food would travel through the digestive system in a different way. While the lower gravity is not really a boon, as Burroughs envisioned, the danger is not as pressing as previously believed. The Astronauts aboard the ISS did experience several problems with muscular dystrophy, but even long term exposure to a low gravity environment still allows the human body to function. To that end, NASA has developed a series of exercises and vitamins which will mitigate the effects of low gravity (Marwaha).

While Mars is unable to hold water now due to the low temperatures and has no life, it still has frozen water at its poles and in the soil. The University of Alabama has proposed, and they have successfully been able to heat up soil and extract water vapor from the surface in simulated experiments (Lewis 2015). The soil provides yet another advantage: it is possible to grow plants which will require less resources to be transported to the planet and enable sustainability. Dust storms could block greenhouses, and the soil would have to have chemicals added in. Despite these conditions, NASA currently believes that it is possible to plant crops on the surface, which will also provide oxygen to the colony (Jordan).
Mars is a hostile landscape, with toxic amounts of radiation, a lethal atmosphere and it largely consists of a barren wasteland which puts the Sahara to shame. That being said, it is possible to extract water and plant crops, so a sustainable colony is not impossible. The planet's environment possesses many dangers such as the radiation and the effects of gravity, but the effects have shown to be exaggerated or solutions have been designed. Besides, there really is no other option; Mars is the only planet that is both reachable and in the goldilocks zone. It is far from ideal, but as the Earth’s resources are being stretched further and further it comes more apparent that this planet is not capable of sustaining upwards of seven billion people forever. The future is not as bleak as it may appear; NASA has proposed a landing by 2030 and a Dutch group has proposed landing by as early as 2017. Mars smaller size has many adverse effects but there is nothing that would stop us from landing on this brave new world.
Works Cited


“Early Times”. *Mars Exploration, NASA*,


“Why is Mars so Tiny?”. Blame Wandering Jupiter.” *NBC News. Com.* Jun./2011,

This paper describes both the history and the physics of how the radio operates. Starting from the men and theories of electromagnetism, Oersted and his electromagnetic fields, Ampere and his characteristics of these fields. Faraday and his discovery of inductance, and Maxwell who could connect ideas of other and see how electromagnetism behaves. Moving to Electromagnetic radiation with Hughes being one of the first to document detecting radio signals, And Hertz proving Maxwell’s theories of electromagnetism to be correct. This all leading to Marconi and Fessenden to pioneer radio broadcasts that would evolve to what is used today.

The radio is without a shadow of a doubt one of the most influential and important devices invented by man. With its abilities to transmit incredible distances, it revolutionized how communicates, and views the world. The base of the radio is how it manipulates electromagnetic (EM) radiation to transmit data to a receiver. This electromagnetic radiation has become a very integrated part of modern society, from transmitting sound, images (television, and radio), to data as in radar, GPS, Wi-Fi and mobile phones. These devices operate in the radio wave spectrum of the electromagnetic spectrum. To understand the physics of how the radio works, first how it came to be must be explored.

First there was electromagnetism, which is the relationship between electric fields and magnetic fields. The first to observe this phenomenon was Hans Christian Oersted, when in April 21, 1820 as he was giving a lecture and demonstrating current passing through a wire, when the wire came close to a magnetic compass, and the needle moved towards the wire \[1\]. This discovery fascinated Oersted and for the three months experimented with this phenomenon and then wrote and published his findings in his paper “Experiments on the Effects of a Current of Electricity on the Magnetic Needle” \[1\]. In this paper Oersted describes how a wire with electrical current flowing through it effects a magnetic needle, and through these experiments describes how the magnetic field of the wire acts as well. Oersted was one that believed that all the forces of nature were connected together, this mindset was influenced by the philosopher Immanuel Kant. During his experiments he discovered that this force between the wire and the needle could be an attractive or repulsive force \[2\]. He did this by noticing how one pole would be attracted to the wire and the other was repelled by this force \[2\]. Oersted figured out that the direction of the magnetic field was dependent on the direction of the current, and the location of the wire in relation to the needle \[2\]. This was discovered because Oersted noted how the needle would change direction if the wire was either above or below the needle, and if the current through the wire was revered \[3\]. Though these discoveries Oersted wrote how the magnetic field created form direct current (DC) through a straight wire behaves, this is called Oersted’s Law. This law is states that the magnetic field lines will encircle the wire, and these lines will be perpendicular to the wire \[4\]. The direction of the magnetic field is dependent on the direction of the current, and the magnitude of the field is dependent on the magnitude of the current \[4\]. And lastly the strength of the field is inversely proportional to the distance from the wire \[4\]. An easier way of imagining how the magnetic field behaves is to use the right-hand rule, which is if you point your thumb in the direction of the current and then wrap your fingers around the wire they will show the direction of the magnetic field.

Oersted first discovered electromagnetic fields, and form his experiments more discoveries are made some directly from his work with the magnetic needle such as Ampere. The same year Oersted published his paper describing the effects electric current and magnetic fields Andre Ampere was shown his experiment and wished to know more. In September of 1820 Ampere was attending a demonstration of Oersted’s experiment at the French Academy \[5\].
Fascinated, just as Oersted, by this phenomenon Ampere began to conduct his own experiments and make his own discoveries that would go on to influence scientists, and discoveries for years. During his experiments, Ampere discovered a that if two wires, in parallel, have current flowing in the same direction will be attracted to one another, and repelled if the currents are opposing [5]. As Ampere continued his experiments he came to a discovery and created a law based on it. He discovered that for a closed loop the sum of the length times the magnetic field along that length is equal to $\mu_0 (4\pi * 10^{-7} \text{ N/A}^2)$ times the current in the loop [6].

$$B = \frac{\mu_0 I}{2\pi r}$$

where

B = Magnetic field
I = electrical current
r = distance of magnetic field from wire.

This discovery would go on to influence other scientists and would lead to discoveries in electromagnetism that would change the world.

Faraday a man that started life as a book binder with a healthy amount of curiosity came to be a great scientist that went on to make countless discoveries that influence so many. The discovery of Faraday’s that this paper will focus on is his discovery of electromagnetic induction. Faraday working at the Royal Institute of Great Britain discovered that if a magnetic field is constantly changing it will create an electrical charge [6]. This alone is a key piece in what makes modern life, with this kinetic energy is able to be converted into electrical energy [6]. This is the principle that generators work on today providing the power that make modern life possible. Faraday’s Law of Induction states that the voltage induced from a magnetic field is dependent on a changing magnetic field. The equation for the average electromagnetic force (EMF) is [7]

$$\varepsilon = -N \frac{\Delta \Phi_B}{\Delta t}$$

where

E = EMF (Volts)
N = number of turns of a coil
$\Phi_B$ = Magnetic flux flowing through an area ($W_b$)
t = time (seconds)

Faraday’s discovery of induction along with the discoveries, of Oersted, and Ampere where the stepping stones that lead to the creation of Maxwell’s Equations.

James Clerk Maxwell is most famously known for his equations that describe how electrical current, and magnetic fields influence each other. Maxwell published these four equations in his text book “Electricity and Magnetism (1873)” [8]. The first of these equations is Gauss’s Law of Electricity, which states that electric flux leaving a closed surface is equal to the electric charge inside divided by $\varepsilon_0 (8.85 * 10^{-12} \text{ F/m})$ [9].
\[ \phi_{\text{electric}} = \frac{Q}{\varepsilon_0} \quad [11] \]

\[ \phi_{\text{electric}} = \text{Electric Flux} \ (W_b) \]

where

\[ Q = \text{electric charge} \ (C) \]

\[ \varepsilon_0 = 8.85 \times 10^{-12} \ (\text{F/m}) \]

Maxwell’s second equation is Gauss’s Law of Magnetic Fields which states that the total magnetic flux flowing through a closed surface is equal to zero \([10]\).

\[ \oint \vec{B} \cdot d\vec{A} = 0 \quad [11] \]

where

\[ B = \text{Magnetic field} \ (T) \]

\[ A = \text{area} \ (m^2) \]

Maxwell modified Faraday’s law of induction to create his third equation, Faraday’s Law which is described as, the changing magnetic flux through a surface will induce a voltage, and a changing magnetic field will create a circulating electric field \([10]\).

\[ \oint \vec{E} \cdot d\vec{s} = -\frac{d\phi_B}{dt} \quad [11] \]

Lastly Maxwell derived his final equation of Ampere’s Law which states that either an electric current or a changing magnetic field through an area will create a circulating magnetic field \([10]\).

\[ \oint B \cdot ds = \mu_0 I + \frac{1}{c^2} \frac{d}{dt} \int E \cdot dA \quad [11] \]

These equations are the equations that describe how electricity and magnetism effect each other, and change how scientists look at the world. In 1862 Maxwell was able to calculate the speed of light using equations for electromagnet fields, form this he concluded that light comes from electromagnetism \([8]\). Later to be called electromagnetic radiation it was not until Heinrich Hertz was this confirmed to be true. These discoveries and equations are what form the base of later scientists who study the phenomenon electromagnetic radiation (radio waves), which is the fundamental principle that radio work use to transmit data.

One of the first documented studies into radio waves was done by a physicist called David Hughes. Hughes was already an accomplished physicist having created several devises and discovering principles. But one in particular is interesting, for at first it was dismissed and so he did not investigate further, this is his discovery of radio waves \([12]\). In 1879 while experimenting with magnetism Hughes had noticed that there was a loose connection in his device that would suddenly spark, and create a noise. He decided to set up an experiment with a coherer, a tube of iron fillings used to detected electrical signals from a distance \([13]\), and a generator that would send electrical pulses through an induction circuit \([12]\). He was able to detect these impulse up to five hundred yards away \([12]\). His findings were dismissed as being merely induction, only later when Hertz does his experiments will Hughes findings be realized.
At this point Maxwell’s equations are merely a theory, and instances of capturing electromagnetic (EM) radiation is dismissed as being induction, this is all about to change. In 1885, at the age of 28, Heinrich Hertz becomes a professor at the University of Karlsruhe where he beings his experiments to prove Maxwell’s theory to be true\cite{14}. Hertz was offered a chance to conduct research into Maxwell’s theories earlier but turned the off away to focus on other projects, but now he would have his chance, and in 1886 he did it\cite{14}. While giving a lecture about sparks Hertz was infatuated by the sparks, and devised an experiment to test Maxwell’s theory that these sparks are, in fact, creating EM radiation. Hertz devised a devise that would generate a high voltage alternating current (AC) spark, this consisted of a spark generator connected to two metal spheres a distance apart to generate a spark, and a metal ring that had a gap between the two ends\cite{15}. When the generator created a spark across the spark gap it would generate EM radiation and the ring would act as a receiver sparking if it captured some of this radiation. This radiation would induce an electrical current in the ring proving that this was indeed EM radiation and not simply induction. When Hertz began his experiment, in November of 1886, he generated sparks across the gap and the ring, that was 1.5 meters away, had sparks arc across its gap\cite{14}. This success was a monumental discovery, having transmitted power across open air with no physical connections, Hertz had discovered electromagnetic radiation. Moving forward, Hertz continued his experiments with his discovery proving Maxwell more and more right. He showcased that this energy did behave like light being “reflected, refracted, produce interference patterns, and produce standing waves just like light”\cite{14}. Hertz was able to confirm Maxwell’s theories, that of EM radiation, and that it and light are one and the same. This discovery would be one that would change the world forever.

The work Hertz did inspire and fascinated scientist everywhere as they began to conduct their own research. One such scientists were Guglielmo Marconi, who would receive the first paten for the wireless telegraph, and be a pioneer in the field of radio communication. In 1895, in his father’s estate, Marconi began his experiments with wireless communications sending and receiving signals at greater and greater distances\cite{16}. These experiments would lead Marconi to travel to England where he would apply, and receive, a patent for a wireless telegraph system\cite{16}. This would be the beginning of a long line of patents and achievements that would change how the world communicates. In 1900 Marconi would patent a device that would allow multiple stations to communicate simultaneously without disrupting each other, this is done with his “tuned or syntonic telegraph”\cite{16}. In the same year he would send the first signal across the Atlantic from England to Canada, proving that radio waves would not be effected by the curvature of the Earth\cite{16}. Marconi is a titan in the radio telegraph industry, creating systems to send and receive signals at incredible distances, and practically. These systems are the stepping stone to the countless advancements that lead to the devices make modern life.

Marconi may be regarded as the father of the radio, but modern radio would not be possible if it was not for Fessenden. Reginald Fessenden was the first person to transmit voice and music in an amplitude modulated (AM) broadcast\cite{17}\cite{18}. Unlike the transmitters used by Marconi and others, that used spark gaps to generate a signal. Fessenden came to the understanding that if voice and other sounds or signals wanted to be transmitted a new system would have to be devised a system that would have to utilize a continuous wave form. Fessenden is most famously known for his 1906 Christmas Eve broadcast which was the first AM broadcast. This broadcast was received by ships in the Atlantic Ocean that were equipped with his new receiver that was capable of un-modulating a signal\cite{17}\cite{18}. This receiver, called a Liquid
Barretter, consisted of a platinum wire that would just break the surface of an acidic solution creating an electrolytic capacitor, with the antenna could tune to a specific frequency \[^{[18]}\]. On December 24 1906, Fessenden changed how the world would look at how radio waves are used. That day Fessenden and his assistants played Christmas music and sung over their transmitter and were received by ship from the US Navy and the United Fruit Company \[^{[17]}\]. This along with achievements in new AC alternators, to create continuous sin waves, Fessenden is truly the father of modern radio.

To transmit sound over radio waves it first has to be modulated. Today there are two main ways of doing this the first is with amplitude modulation (AM). This is the process of changing the amplitude of a carrier wave to in tune to a signal wave. The other is frequency modulation (FM), this is changing the frequency of a carrier wave in tune to a signal wave. Modulation is a necessary process for transmitting data reliably. AM and FM have their own characteristics that make them more capable at certain tasks compared to the other. AM, because it uses a lower frequency, and there for a longer wavelength, is able to travel longer distances, and through objects. Unfortunately, AM is susceptible to forms of interference, such as the sun or other cosmic radiation that is able to distort these signals. FM does not have this issue due to how this type of modulation does not occur in nature. FM does run in to a problem when it comes to reception distance, because of its higher frequencies it is unable to travel the vast distances that AM is capable of. Distance and obstacles are an issue for FM broadcasts because of its shorter wavelength these signals are unable to move through object, such as mountains. Amplitude and frequency modulation have their places in both history and modern life.

As stated early amplitude modulation is varying the amplitude of a carrier wave to that of a signal wave. This can be expressed mathematically with the equation \[^{[19]}\]

\[
s(t) = A_c [1 + k_a m(t)] \cos (2\pi f_c t)
\]

Where

\[
m(t) = \text{baseband signal}
\]

\[
f_c = \text{carrier frequency}
\]

\[
A_c = \text{carrier amplitude}
\]

\[
k_a = \text{modulation index}
\]

There are a several ways of doing this with electrical components, but one of the simplest would be one that varies that the amount of voltage an oscillator, set to the carrier frequency, receives due to an audio signal. To receive this, signal a RCL (resistor, capacitor, inductor) circuit connected to the antenna and is tuned till its resonance frequency is that of the signal \[^{[20]}\]. Once it is received it can be de-modulated with a low pass filter, to remove the high frequency carrier wave, and amplified \[^{[20]}\].

Frequency modulation is the frequency of a carrier wave that is modulated by some signal that is wanted to be transmitted. This is expressed by the equation

\[
\cos(\omega_c t + B \sin \omega_m t) \[^{[21]}\]
\]

where
The principle behind how to make a frequency modulated transmitter is that some carrier wave need to have its frequency to change along with a signal. This can be accomplished by creating a LC oscillator that is connected to an audio signal. When the audio signal is added to the oscillator they are combined creating a FM signal.[22]

The radio is undoubtedly one of the most important inventions of the twentieth century as it has led to the creation of many new tools that make modern life. Throughout this project I have learned a great deal about not only the physics behind the radio but also its history. Both I believe are important aspects that should be explored. For instance, how it was not just one person that created the radio, but several that laid the ground work for others to improve or prove. Such as Maxwell and the scientists before him that lead to his understanding of how electromagnetic radiation behaves. And that of Hertz that proved Maxwell to be correct in his theories. And with the theory proven people like Hughes and Marconi blazed a path for radio engineering till Fessenden changed everything and spoke over the air waves. History is an important part of understanding not only where we came from but where it may lead us to.

The physics behind the radio going back to the works that inspired Maxwell to how information is transmitted today, was an undertaking and eye opening. A quick search of the physics of the radio rapidly sent me down a rabbit hole that I knew there was more here than I could understand and write in this paper. Some, if not most, of the mathematics behind these ideas are beyond my scope, but nonetheless was interesting finding explanations for these equations and what they meant for the future of the radio. This is not a complete description of the physics behind the radio, but merely a brief history and explanation of what each of these scientists discovered and how that lead to what we us today.
Works Cited


Arsenic: The Silent Killer
Savannah Cummins
April 20, 2017
General Chemistry II, CHM152
Professor Olander
ABSTRACT:

Acceptable levels of arsenic in drinking water as set by the USEPA and the WHO are between 0.010 mg/L and 0.050 mg/L (Smith et al. 2000). With levels ranging in Bangladesh anywhere between 0.050 mg/L and 5.000 mg/L (Asadullah and Chaudhury 2011), extreme arsenic poisoning can be seen throughout the developing country. Arsenicosis, the poisoning of a being by toxic and inorganic arsenic, is being seen throughout many rural areas of Bangladesh. However, these symptoms are vague and are rarely directly linked with arsenic poisoning. Treatments, expensive and often hard to come by in rural Bangladesh, are every rarely offered due to the lack of knowledge about the poisoning occurring throughout the country. Steps are being taken by the government of Bangladesh to reduce toxic tube wells from being used for drinking water by the public. Test are run on a majority of tube wells already put in place and determined to be safe or not based upon acceptable concentration levels of arsenic set by governmental entities, however many of these wells continue to be used by families without another option. New deep tube wells are drilled in attempt to replace ones that have been deemed “contaminated”. These wells are placed in areas where clean water is already available and areas of true need are skipped over. Government politics continue to stand as more important than the lives of people in these rural communities. These contaminated water sources also have begun to contaminate food crops for the people, only adding to their daily intake of toxic arsenic. With options available to reduce toxin levels and potentially eliminate the use of contaminated water sources, there is hope for these citizens to come. Steps need to be taken to protect the lives of rural areas of Bangladesh.

Arsenic exposure by way of drinking water obtained through tube wells continues to plague citizens of Bangladesh more than 20 years after its detection. This naturally occurring, inorganic, highly toxic element contaminates much of the groundwater and well water throughout the United States, Mexico, Chile, Mongolia, Taiwan, India, and Argentina (Uddin and Huda 2011). However, the highest levels of contaminated water continue to be found throughout the Ganges and Brahmaputra rivers that originate in the Himalayas. These rivers flow into the “13th largest river basin in the world”, the Ganges-Brahmaputra river basin (Smith et al. 2000). The Ganges and Brahmaputra rivers flow from the Himalayas, transporting arsenic rich dirt and soils along the way, eventually depositing the toxic metalloid into groundwater and aquifers of the delta. It is estimated that upwards of 77 million people of Bangladesh have been exposed to these extremely high levels of toxic arsenic through tube wells used for drinking water (Jiang et al. 2012).

Arsenic is a natural element, formed with the creation of our planet, and the 20th most ample element found on the earth’s surface (Singh 2006). This metalloid is “released into the environment through natural processes such as weathering and volcanic eruptions, and may be transported over long distances as suspended particles and aerosols through water or air” (Singh 2006, 599-600). Much of the groundwater contamination with Bangladesh is linked to these weathering processes of the Himalayan mountains, however there are other concerns as to how this arsenic rich water has come to be a problem. The oxidation of arsenic rich pyrite (FeS2), a sulfide mineral, can also be linked to high levels of arsenic in groundwater sources (Singh 2006). This solid and relatively stable mineral oxidizes into arsenate by way of the mineral manganese (Mn), freely existing within the surrounding sediment (Stollenwerk et al. 2007). Other factors, such as overflow of organic carbon sources as a result from uncontrolled pumping of...
groundwater for irrigation of rice fields, release As (arsenic) into the surrounding environment and associated crops from the phosphate rich fertilizers (Jiang et al. 2012). Each of these scenarios contribute to the overall arsenic contamination of source water throughout South Asia.

Acceptable levels of arsenic in drinking water as established by the United States Environmental Protection Agency and the World Health Organization have been set to 0.010 mg/L (parts per million), which is equivalent to 10 µg/L (micrograms per liter) (Smith et al. 2000). These levels have been set as a standard for developed countries, while third world and developing countries have been permitted limits around 50 µg/L (micrograms per liter) such as in Bangladesh (Uddin and Huda 2011). However, it is estimated that over 77 million people of Bangladesh have been exposed to arsenic levels of around 500 µg/L, 50 times greater than the WHO permitted levels throughout the world. Between 2000 and 2003, an estimated 5 million wells throughout Bangladesh were tested for arsenic levels, around 2 million of which contained levels of over 50 µg/L (UNICEF 2008). Another 200,000 tube well have also gone untested, the majority of which are expected to exceed the established limit, resulting in approximately one fifth of wells providing unsafe drinking water to the Bangladesh people (UNICEF 2008).

Tube wells found with safe levels of arsenic are painted green and tagged with labels displaying accurate values of these levels (see Figure 1), while wells with high and toxic levels of arsenic are painted red and labeled with warnings, many times leading to the affected wells to be disabled. While painting and labeling these wells is very beneficial to the health and knowledge of the population of Bangladesh, these paints and labels eventually wear off due to the constant sun and weather beating down on the wells (see Figure 2). This degradation of warning signs on wells can eventually lead to citizens continuing to use contaminated wells, in the end deeming all testing ultimately useless. In fact, such efforts put forth by the national screening services have also diminished since 2006. Much of the drinking water has returned to prior quality with high levels of toxic arsenic since 2013 (Human Rights Watch 2016).

Due to these extremely elevated levels of arsenic within the well water of Bangladesh and the surrounding areas, arsenicosis is exceptionally prevalent within the rural populations. Arsenicosis is a term that refers to effects on a person or population of arsenic poisoning over an extensive period of time, usually 5 to 20 years (Asadullah and Chaudhury 2011). This exposure is believed to have effected between 37 million and 77 million citizens of rural areas of Bangladesh, where better alternative water sources are not readily available (Smith et al. 2000), though only about 40,000 residents with symptoms have been identified where medical attention is few and far between (UNICEF 2008). Long term effects of this natural, but extremely dangerous toxin include: skin lesions (see figure 3), external and internal cancers, diabetes, cardiovascular diseases, hypertension, developmental delays, pulmonary disease, as well as many neurological deficits (Smith et al. 2000; Human Rights Watch 2016) (see Figure 4). Signs and symptoms of arsenic poisoning include: metallic taste in mouth, excessive saliva production, problems swallowing, blood in the urines, cramping of the muscles, hair loss, stomach cramps, convulsions, excessive sweating, garlic smelling breath, vomiting, and diarrhea (Gilbert 2016). Arsenic poisoning most commonly effects the skin, liver, lungs, and kidneys. Depending on the severity of the symptoms of the poisoning, seizure can ensue, leading people to go into shock, ending in a coma and commonly death.
Arsenic affects the body by disrupting the creation of adenosine triphosphate (ATP), the nucleotide in charge of moving energy through the cells, completely blocking and competing with ATP. By blocking these pathways within the cells, systems within the body begin to fail: neurological, cardiovascular, muscular, and immune (LiveScience 2010). Not only does arsenic compete to shut down organ systems within the body by way of transmitters, it will also attack the organs separately and specifically. Arsenic is one of very few toxins that actually attack multiple organs rather than just one. By focusing on the entirety of the body, arsenic can much more effectively and quickly destroy the functions within it.

The main concern for persons exposed to continuously high levels of ingestible arsenic are the potential cancerous effects. Based on a study conducted in 2000 on a sample of 65,876 people, it is suggested that there is “at least a doubling of the potential cancer burden in Bangladesh due to arsenic exposure” (Chen and Ahsan 2004). Results of this study can be seen in Figure 5. Inorganic and organic versions of arsenic have been classified by The International Agency for Research on Cancer (IARC), The US Department of Health and Human Services (DHHS), and US EPA as a known human carcinogen and can be exceptionally lethal in even small doses, 0.14 grams (less than an eight of a teaspoon), having the potential to kill a child and even a healthy adult (Gilbert 2016). While arsenic can cause cancer, in controlled amounts and specific treatments, arsenic can also be used in the treatment of cancer. These treatments have seen to be very effectful and can even put cancer into remission. High levels however, such as levels seen in Bangladesh, are more harmful than helpful. Next to cancers of the kidney, lung, bladder, and skin, peripheral neuropathy is one of the most serious conditions resulting to chronic exposure to arsenic (Flanagan et al. 2012). Neuropathy occurs when the body sustains damage to peripheral nerve endings, resulting in numbness and pain in the extremities (see Figure 6). One’s hands and feet become weak and unable to accurately send touch information to the rest of the body. This results in intense and inaccurate burning or freezing sensations, along with sharp pains and agonizing throbbing of the extremities. This phenomenon also occurs in cancer patients who have received extremely high doses of powerful chemotherapy drugs, causing extensive and unrepairable nerve damage throughout the body. Although these health effects of arsenicosis are quite extreme and often times life threatening, they appear slowly and are quite generally vague symptoms that do not point directly to arsenic poisoning. Because of the unclear symptoms, diagnosis and treatment are quite rare.

Because of the sheer number of inhabitants being affected by this contamination, it has been deemed the “largest mass poisoning of a population in history” (Smith et al. 2000). During the early 1990’s and into 2003, researchers sampled shallow tube wells throughout the country testing for severely toxic levels of arsenic within the drinking water. With 35 percent of wells containing concentrations over 50 µg/L and 8.4 percent of wells containing concentrations over 300 µg/L, it was apparent that actions needed to be taken to protect the citizens (Smith et al. 2000). Along with painting and placing warning signs on these shallow wells, it became necessary to dig new, deeper wells. These new tube wells tap into deep aquifers around 200 meters deep. At this depth, water supplies were mainly arsenic free (Jiang et al. 2012).

Due to these extreme numbers, steps have been taken by governmental entities to dig these deeper wells. Between 2006 and 2012, government officials began installing approximately 125,000 new deep tube wells. “Although deep wells can reach groundwater of better quality, government programs to install new wells [did not] make it a priority to install them in areas where the risk of arsenic contamination is relatively high” (Human Rights Watch 2016).
Politicians of Bangladesh were seen to divert government funded deep wells away from rural areas that needed them most, and into areas of their political supporters. Of wells installed by governmental programs between 2006 and 2012, 36 percent were installed in areas where safe drinking water was not available, 63 percent were installed in areas where safe drinking water was available to more than 50 percent of the surrounding citizens, and 10 percent were installed in areas that contained absolutely no arsenic contamination whatsoever (Human Rights Watch 2016). Not only did these new deeper and safer wells get placed in areas where they were not needed, it has also been seen that more than 5 percent of these wells newly installed are also contaminated well above governmental and worldwide standards (Human Rights Watch 2016).

There are also potential problems with drilling deeper into the earth for clean well water. Immediate contamination of deep aquifers is a potential complication. Without precise and careful drilling, contaminated groundwater has the opportunity to seep down into these aquifers, ultimately polluting this fresh water with arsenic rich materials (Smith et al. 2000). If deep tube wells are drilled successfully, without introducing arsenic filled debris, there is also the high probability that these arsenic rich minerals will eventually seep further into the earth and lead to contamination within the fresh, deep aquifers (Stollenwerk et al. 2007). The unfortunate truth is that much more will need to be done in order to provide this population with arsenic free, completely safe drinking water.

Drinking water is the main problem for citizens of Bangladesh, but crops are a major source of poisoning as well. “Rice is the most important cereal and staple food for the people of Bangladesh, who eat an average of 450 grams of rice a day” (Jiang et al. 2012). In Bangladesh, soil arsenic levels reach to 83 µg/g, contributing to As levels within the crop. The wet season brings anywhere from 15 to 50 feet of water per year, naturally watering crops without introducing added arsenic levels. However, during the dry season, rice paddies are flooded with the only source of water the people of Bangladesh have, arsenic contaminated groundwater. Over 1,000 tonnes are irrigated into rice fields every year (Jiang et al. 2012). With the consumption of more than 450 grams of rice a day, containing 0.08 µg/g As, the people of Bangladesh add an additional 36 µg As to their systems every day. Most crops do not absorb much arsenic from the soil and water, however, rice is a strong exception to this standard. Although rice and vegetables are sources of extremely high amounts of arsenic within the diet of the Bangladesh people, seafood far surpasses these levels. Seafood, mostly shellfish, have very high levels of arsenic within their meat, the difference though is that the arsenic within the fresh seafood is much less toxic and harmful to the body (Gilbert 2016). Therefore, groundwater is the most worrisome aspect of Bangladesh’s arsenic contamination, but not the only worry.

Though digging deeper tube wells has been the main problem solving method for reducing arsenic levels in source water throughout Bangladesh, quite a few other methods have been proposed. One such method has been presented as providing each household individual filter systems. A candle filtration system is one option in this realm (Smith et al. 2000). These clay filters are used to remove toxins and suspended materials from drinking water and are used as storage for filtered water. This filtration system requires no energy and is a safe system for the removal of toxic arsenic form the water of Bangladesh. Unfortunately, this method has its drawbacks, one being the candle filters must be disposed of very regularly and replaced, and may not filter out all toxins. This poses be a problem in rural areas where availability of these filters is extremely limited.
The second option is a mixture of powdered chemicals supplied to the people of Bangladesh. These chemicals are added to batched of well water and left to sit for a day (Smith et al. 2000). By adding the powdered chemicals to the water, the toxins will be drawn out of the water and collected on the bottom of the container in a thick sludge. Chemically treating the groundwater one batch at a time is a viable short term solution that is relatively inexpensive, but over time could potentially cause more hazards. This sludge created from the toxic arsenic and chemical mixture must be disposed of after every treatment, leaving one wondering where the now toxic sludge could be discarded. While the treatment would quickly and successful create nontoxic drinking water, the aftermath of the treatment may be potentially just as dangerous. And finally, an option posed by many people is the use of surface water with filtration and the addition of chlorine as a means to purify the water (Smith et al. 2000). This option, however, just as many of the ones before, comes with its risks and drawbacks. The addition of chlorine to already potentially contaminated surface water is not a guarantee that the water will be of a quality that is passable by WHO or other world organizations that monitor water supplies. By adding more chemicals to attempt and withdraw toxins within a liquid, the chemicals used will be entering the body instead. Each method of reducing arsenic contamination or finding other sources are valiant efforts, but further problem solving and idea making will highly benefit the people of this toxin controlled area of the world.

In a country ravaged by toxic arsenic contamination and arsenic poisoning, relief and realistic solutions should be the number one priority of not only the government, but of everyone in the world. Over 50 million people of Bangladesh have been exposed to extremely high levels of arsenic within groundwater just in the last year alone. Around 40,000 citizens will die early deaths every year due to natural poisoning of their drinking water and crops, from illnesses such as: cancer, hypertension, pulmonary diseases, and cardiovascular disease. So much is known about the contamination and methods to prevent further population for being poisoned, however the proper steps must be taken in order to save lives. Simply creating temporary preventative measures is not enough.

Deep tube wells must be dug in rural areas of the country, where the population is the most susceptible to ingesting this toxin. Regulations and safe drilling methods must also be implemented. With proper drilling methods and care taken to not contaminate aquifers further than they already have, quality drinking water can be made available for all people of Bangladesh. As discussed, aquifers also run the risk of becoming contaminated with arsenic by the metalloid seeping through layers of soil and gravel, but this could take many decades. The most reliable and realistic solution at this time is to create sources of usable drinking water, while government and worldwide officials come together and create a long-term solution for the ground water contamination. In years to come Bangladesh may begin importing clean and safe drinking water for all their citizens, but no progress will be made until greater awareness is established and discussions into solutions begin.
FIGURES:

Figure 1 (right): Green Painted Well in Bangladesh

This photo shows one of them many tested wells within Bangladesh. This particular one was painted green to depict safety for the public to use the well for drinking water. Wells tested and found to contain under 50 µg/L arsenic are painted green and labeled with arsenic level values.


Figure 2 (left): Disconnected Toxic Well in Bangladesh

This photo depicts a disconnected well in Bangladesh. This well was tested and found to have levels above the allowed 50 µg/L. This well was also painted red to deter use, but has since faded due to exposure to the elements.


Figure 3 (right): Skin Lesions due to Arsenic Poisoning

This photo shows the effects of arsenic poisoning on the extremities. Skin lesions such as these will appear, covering the chest, arms and legs, palms, and soles of the feet. Changes in the pigmentation and the growth of wart-like skin lesions are clear characteristics of skin lesions resulting from arsenic poisoning.

Smith AH, Lingas EO, Rahman M. Contamination of drinking-water by arsenic in Bangladesh: a public health

Figure 4: Arsenic’s Effects on the Human Body

A large majority of the major health effects arsenic poisoning have on the body are depicted within this photo. Not only does arsenicosis effect health years after exposure (5-20 years), but immediate effects can also be seen within young children and adolescents.

Figure 5: Lifetime Excess Morality Risks (per 100,000 Population) From Bladder, Liver, and Lung Cancers: Bangladesh

This table exemplifies the findings of Chen and Ahsan during their 2000 study on the effects of chronic arsenic exposure in Bangladesh. Drastic increases in cancer risks can be seen in high exposure situations, resulting in almost a doubling of cancer risks over a lifetime.

Figure 6: Peripheral Neuropathy

This figure shows the causes of peripheral neuropathy. Nerve endings become severely damaged, causing information being sent throughout the body to be interrupted and misinterpreted.

PERIPHERAL NEUROPATHY. Sound Pain Solutions. [accessed 2017 Feb 14].
http://soundpainsolutions.com/peripheral-neuropathy/
REFERENCES:


3D Printing

Angela Davis

Physics 112
Dr. Durandet
November 16, 2016
ABSTRACT

The technology of 3D printing has grown exponentially over the last few decades. While every printer employs the same basic principle of adding a material layer by layer they differ widely in the techniques in which the layers are applied. From thermoplastics to laser beams the physics behind these machines has enabled an entirely new realm of innovation. 3D printers have many promising potential future applications in almost every field imaginable.

At its most basic 3D printing is the application of taking a digital file and turning it into a physical structure. Despite its relatively recent popularity, the technology first emerged in the early 1980’s and was pioneered by innovator Charles “Chuck” Hull. Although the first method of plastic based three dimensional printing was invented in 1981 by a Japanese scientist through use of a special polymer and a photo-hardening process, it was prematurely abandoned for lack of perspective on future business implementations. Yet, in 1984 Hull patented the technique of stereolithography and continued his investigations in conjunction with the newly formed corporation 3D Systems. Together they produced the first fully-fledged prototype 3D printer using the approach of stereolithography which in turn has become one of the most well known and widely used printing formats. Since then the technology has progressed at an astounding rate leading to the discovery of a multitude of new printing methods as well as an impressive expansion of material mediums.

Regardless of the approach all 3D printers have some common features including a means to input the desired objects printing specifications electronically and a medium or “ink” which is applied in layers to construct the object. This is referred to as additive manufacturing and although all printers rely on this basic concept they differ in the ways in which the layers are applied. Since 2010 the American Society for Testing and Materials (ASTM) has created standards which classify 3D printing into seven categories based upon their functional processes. These categories are: Vat Photopolymerisation, Material Jetting, Binder Jetting, Material Extrusion, Powder Bed Fusion, Sheet Lamination, and Direct Energy Deposition. Yet none of these printers would be able to function without the ability to turn electrical energy into mechanical energy. This is accomplished through the use of a stepper or servo-motor which drives the printhead and tabletop into the correct location. The interaction between magnetic and electric fields is a fundamental property of all motors and is induced when an electric current moves through a coil creating a magnetic field which in turn interacts with a permanent magnetic field. In this way when a coil is charged and allowed to rotate it will align itself with the magnetic north of the permanent magnet and cease its rotation, then when the charge is cut off the coil will continue to rotate. By turning the electric current on and off the coil will continue rotating which will then rotate a shaft and move an object. The types of motors used in 3D printer apply the same basic physics but are so precise that they can move the object, such as the printhead or tabletop, accurately up to 0.00001 meters. The ability of a step motor to govern such precise movements is key to any 3D printer regardless of the printing method that is used. Such a specialized motor also requires a specialized driver in order to make the motor turn. The motor specifications can be used to determine the current requirement of the driver through the use of Ohm’s Law which is given below:
V=IR

One of the original printing methods, Vat Polymerisation, consist of a container which is filled with a polymer resin and the use of a UV light which hardens the resin. The most common technique for this process is stereolithography which uses the laser to harden the resin one layer at a time. As shown in Figure 1, for each layer the laser traces a pattern on the surface of the liquid resin which then solidifies it and joins it to the layer beneath it. After the initial pattern has been completed the container descends by the thickness of one layer, usually 0.05 mm to 0.15 mm, and a resin filled blade re-coats the surface with fresh material. Then a new pattern is traced and joined to the previous layer. This process is repeated until a complete three dimensional object has been formed. However, this method requires the use of support structures to hold the object in place as it floats in the liquid resin and the supports are removed manually after the object has been finished. Recently the fledgling 3D printing company Carbon has developed an exciting new technique which is based on the same principles of Vat Polymerisation. CLIP, which stands for Continuous Liquid Interface Production, still relies on liquid resin and a photo hardening light source, however, due to an oxygen window the printing can take place up to 100 times faster and without interruption or visible layers. By using a digital light projector to emit a continuous sequence of uv images through the oxygen window at the bottom the layers are formed continuously rather than one at a time. Because of this, CLIP is able to produce objects with almost no visible layers whose structural integrity far surpasses that of the original resin hardening process. Additionally, Carbon has designed a diverse selection of resins which were created to meet some of the common engineering requirements. Among these include a polyurethane elastomer which retains the resilience of an injection mold as well as resins with a temperature resistance similar to glass filled nylon. The structural integrity, variety of resins, and considerably shortened construction time allow CLIP the potential to change the way 3D printing can be used in mass production by alleviating many of the issues associated with traditional Vat Polymerisation.

Material Jetting is a process similar to the way a common inkjet paper printer works. A medium is applied layer by layer in droplets through a small nozzle print head. The nozzle moves horizontally across the build platform depositing the material which is then hardened by uv light. Figure 2 depicts this process with a printhead for both the support and build materials. The material is then jetted onto the surface of the platform through an oscillating nozzle. The droplets are charged and positioned through the use of charged deflection plates. Because the system is continuous it allows for a high degree of positioning and control. Any drops that are not used are then recycled back into the printing system. This is referred to as Drop On Demand (DOD) with more drops being added until the object is complete. However, because of this the viscosity of the material used must be able to form the droplets and is limited to polymers and waxes. Unlike continuous printing, the drops are only dispensed when needed and are released by a pressure change in the nozzle by a piezoelectric or thermal actuator. A thermal actuator can deposit the drops at a faster rate yet the piezoelectric actuator is preferred because it allows for a wider range of mediums to be used. Additionally there is also a need to refill the medium reservoir often which can affect the print speed. The advantages of this process are that it is highly accurate and therefore accrues little waste product as well as that it can deposit multiple colors at a time. However, support structures are almost always required and the mediums are limited to polymers and waxes which are capable of forming droplets.
First developed in 1993 at MIT, Binder Jet printing employs both a powder based medium and a binding liquid. The binder acts as an adhesive between the particles of the powder material. As seen in Figure 3 the build material is stored in a container at the bottom while the binding material is added from above. The printing nozzle moves horizontally across the platform and deposits alternating layers of the powder and glue over the x and y axis. After each layer is printed the build platform lowers by the thickness of one layer and a roller replaces a layer of build material until the object has been completed. This process allows for multiple color printing and can use a variety of mediums including metals, polymers, and ceramics. This technique is faster than many of its counterparts and print speed can be increased through the use of additional printer heads. Despite the relative speed of printing this method of binding creates objects whose material characteristics lack in structural integrity. The post processing that is needed to remedy this can add a significant amount of time to the overall printing process as the binder needs an adequate amount of time to set. However, by using a two material approach a variety of binder-powder combinations can be made by changing the ratios and individual properties of the two materials.

Fused Deposition Modeling is the most common type of Material Extrusion process and was trademarked by the company Stratasys after its development in 1988 by Scott Crump. This is the most widely used technique in personal desktop 3D printers. A nozzle is used to heat a medium, typically a plastic or metal filament, which can then be applied in layers in both the horizontal and vertical positions. The material hardens directly after being extruded from the nozzle and is controlled by a computer aided manufacturing software. The software that comes with the desktop printers will automatically generate support structures if required and dispenses a separate disposable material for the structure. This setup is depicted in Figure 4 and due to its relatively straightforward approach parts can be bought individually enabling the customer to build customized versions of this printer. The elementary mechanical components these printers include a print bed, filament, extruder, motion control, and a motor. While the electrical components consist of a power supply, motherboard, and step drivers.

The print bed is the surface on which the object is printed and can vary in size and function. For example, material extrusion printers which often print in plastics include a heating element to prevent the object from warping. When the plastic cools and shrinks due to thermal contraction it warps upward and peels off of the base, however a heated bed keeps the object warm in order to prevent this. Additionally a thin film is also added to the bed which allows the plastic to stick during printing but is easily removable when finished. The filament is the medium which is used by the printer and comes in the form of a spool. Filaments are typically a plastic or metal although many different materials are available. The extruder is the core of the printer and is where the filament gets drawn in, melted, and pushed out. The extruder consists of a cold end where a small motor draws the filament in and pushes it through and a hot end where the filament is melted and discharged. One type of extruder which is called a direct drive has both the hot and cold ends stacked together and the filament goes straight down from one end to the other. Another type of extruder is called a Bowden setup in which the hot and cold ends are separated. In this set up the cold end is stationary and bolted to the frame while the filament is fed through a tube to the hot end and pushed out which decreases the amount of weight that the printer has to move around. Because the hot ends to not use plastic insulators in their construction they can reach a much higher temperature allowing it to print a wider variety of materials. However, to prevent heat creep where heat travels up the plastic and melts it
prematurely hot ends require active cooling through the use of a specialized fan. The nozzle is the piece in which the melted filament is dispensed. Nozzles come in various sizes with small ones for finer detail and larger ones for faster printing. However, nozzles often get clogged and require frequent cleaning which is a common issue for desktop 3D printers. The motion control for the nozzle is controlled through either a delta or cartesian setup. The cartesian motion control system has either one or two motors along both the x, y, and z axis. The extruder is attached to a platform while the motors control the direction of the platform. A delta printer has three arms that join together in the center to suspend the extruder above the build platform or print bed. The delta set up also uses the cartesian coordinate system to move around in, however, instead of moving one motor per axis at a time all three arms move at different rates to adjust the nozzle precisely through triangulation. The final mechanical component is the step motor which rotates in increments rather than continuously like a DC motor in order to give the printer more precise control over its movements.

One of the main electrical components in a home printer is the power supply. This device takes the 120V AC power from the wall socket and converts it into a low volt DC supply for the printer to use. The most common power supply is an ATX which is used in desktop computers and is often repurposed for the printers. These power supplies are beefy and efficient and can provide a variety of different voltages. It is critical to be aware of the printers voltage system when replacing components as some printer run of 12V systems and some off 24V systems. Another key electrical component is the motherboard which is essentially the brain of the printer. The computer gives the specific commands and the motherboard orchestrates them. The motherboard contains all of the circuitry needed to run the motors, read sensors, and relay to the computer. The last important electrical component is the step driver which is used to run the step motors. This driver controls the firing sequence of the coils which allows the motor to move in increments. Often the step driver is built into the motherboard and can balance the power delivered to each coil which divides the steps into further increment. This is called micro stepping and allows more control over the motor than is normally possible. Additionally the step driver controls how much current is supplied to the motor.

Powder Bed Fusion printers use a technology called Selective laser sintering (SLS). This method uses a powerful laser to fuse small particle of metal, ceramic, plastic, or glass powder into the shape of the desired object. These printer are made of three main components including a heat source to fuse the material, a mechanism to control the amount of heat, and a device to add new layers of material over the previous layer. This type of printer benefits because it does not need any type of additional support structure as the powder base provides enough support to the object throughout the build process as seen in Figure 5. The build platform which chambers the powder material is temperature controlled and kept at a temperature a few degrees below the melting point of the material that is being used. This in effect reduces the dependency of the laser to fuse the layer together. Nitrogen within the chamber is used to maximize oxidation which in turn increases the end quality of the object. Once completed the object will require a cool down period to ensure the highest quality of fusion. Electron Beam Melting (EBM) is another type of Powder Bed Fusion in which an electron beam is used to melt metal powders. This type of printing creates objects with increased strength and structural integrity due to the even temperature distribution during fusion. In fact EBM creates object whose standards compare with those used in airplanes or medical devices. However, post processing requires the removal of excess powder as well as some CNC work for a finished product. Ultimately this type of
printing is relatively inexpensive and can work with a wide range of materials, yet the process is time consuming and can only produce objects of limited size.

Sheet Lamination printing includes techniques such as ultrasonic additive manufacturing (UAM) and laminated object manufacturing (LOM). The UAM process employs sheets or ribbons of metal which are bound together by ultrasonic welding. The finished product must be CNC to remove any unwanted excess material. Some of the metals that can be used are copper, aluminum, stainless steel and titanium. By using ultrasonic frequencies and pressure to join the layers rather than melting the process can bind many different metals requiring relatively little energy to do so. LOM uses a similar approach but uses paper instead of metal and glue instead of welding. Unlike UAM, LOM uses a cross hatching method during the process which allows for the easy removal of unwanted material from the finished product. An example of an LOM printing setup is depicted in figure 6. Once all of the layers have been successfully joined a knife or laser is used to cut the final shape. This occurs at a very low temperature which allows for internal shapes to be made. The objects made from sheet lamination are often used for aesthetics or visual models and are not suitable for structural use.

Direct Energy Deposition (DED) is primarily used in the high tech metal industry. Often this printing process is used to repair or add additional material to existing parts. As seen in Figure 7 a typical DED printer consists of a nozzle mounted upon a multi axis arm which deposits the material onto the build platform where it solidifies. The material used in this printer is in either powder or wire form and is melted upon deposition by either a laser, electron beam, or plasma arc. Although this process can use polymers and ceramics it is most commonly used with metal mediums. This type of printer makes efficient use of its material but requires some degree of post processing to achieve the final object. Despite the potential uses for this type of 3D printer the process still requires quite a bit of research until it can effectively become mainstream.

All of the approaches depicted above are still being researched and improved upon today. However, the technology develops rapidly and new printing methods have been already discovered. Hailed as the third industrial revolution the scope of 3D printing continues to grow and expand into a variety of different fields. Some of the most promising areas include customized consumer products, custom parts replacement, and numerous medical applications. In 2013, the Chinese scientists began to print human ears, eyes, and kidneys, based on the use of living tissue as a working medium material. Additionally, the researchers of Hangzhou Dianzi University have developed a 3D-bioprinter called Regenovo. The lead developer of Regenovo, Xu Ming, predicted at the time that functional print organs can be created over the next ten to twenty years. In the same year, the researchers from the University of Hasselt in Belgium have successfully printed a new jaw for an 83-year-old Belgian woman. These laboratory investigations allow researchers to print the organs and implant them to the body of laboratory mice and possibly one day humans. The researches of the University of Missouri managed to apply a special bio-gel clumps, based on the living cells of a given type. This type of printing research is investigating the cultivation of high-grade organs. Already, 3D printing has proven to be profitable in prosthetics and the manufacturing of implants, which include the fragments of the skeleton, skull, bones, and cartilage. Not long ago, the German scientists invented a technology for producing the human skin from gel obtained from the donor cells. In 2011, scientists were able to reproduced the living human kidney via 3D printing.
The technology of 3D printing can be used successfully in the foundry industry as well. The design of 3D printing construction, based on transparent material, allows engineers to see the work of the mechanism from inside. This tool was used by Porsche engineers when studying the oil pressure in the transmission system before the development of the engine. One of the most significant spheres of 3D printing implementation is the sphere of daily routine. 3D printers allow to design both little things for home and numerous complex, mass, durable, and low-cost systems. For instance, the unmanned aircraft Polecat of the Lockheed company were designed by a high-speed three-dimensional printing technology.

Another significant sphere of 3D printing implementation is in development and architecture. For instance, in architecture, 3D printing allows the specialists to create three-dimensional layouts of the buildings, the entire neighborhoods, and the infrastructure, including squares, parks, road, and street lighting. Due to the use of a cheap gypsum composite, the ready-made models appear to be at far lower costs than traditional layouts made by hands. Besides, the contemporary 3D printing appliances used in architecture allow for the models to be in more than 390,000 color shades.

3D printing technology is one of the most progressive industries in today's market. It can be applied to the great variety of scientific and business spheres, including the development, architecture, food industry, machinery, space craft, foundry industry, medicine, jewelry, and home appliances. As for the perspectives of the technology, it will eliminate mass production and reveal the sense of small batches and customized objects to be designed per individual request. Using 3D printing technologies implies the rise of Maker Movement around the globe, since the consumers become the true makers of the things they need. The process of education will also benefit using 3D printing, as it allows students to make their projects real.

One of the most important outcomes of 3D printing technology is that it allows to create a country of makers. Today, there is a distinct difference between the producers and the consumers in the sphere of manufacturing. 3D printing eliminates this difference, as it is able to turn the consumers into creators of the things. Due to the impact of 3D printing, the new Maker Movement can be seen across the world. The movement of makers helps to spur the technological innovations and transform the whole business structure between mass production and consumer printing. 3D printing technology development eliminates the influence of mass production.

Aeronautics, medicine, and development are the main spheres, where 3D printing technologies can be applied most effectively. The researchers of these fields are interested in improvements, low-cost, and fast decisions to design and evaluate their projects. As well as 3D printing technology transforms and develops constantly, it can provide the most progressive and innovative approaches to design and construction today. The technologies based on 3D printing gain brilliant future perspectives, eliminate mass production, and help to realize the customized needs and fast decisions as the most foreground technology today.
Figure 1: Vat Polymerization printing setup

http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation

Figure 2: Material Jetting printer setup

http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation

Figure 3: Binder Jet printer setup

http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation

Figure 4: Material Extrusion printer setup

http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation
Figure 5: Powder Bed Fusion printer setup
http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation

Figure 6: Sheet Lamination (LOM) printer setup
http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation

Figure 7: Direct Energy Deposition printer setup
http://www.lboro.ac.uk/research/amrg/about/the7categoriesofadditivemanufacturing/vatphotopolymerisation
Literature Cited


9. Thermosetting Polymers Plastics Engineering Jean P Pdf - Ebooks Download. Pdf -
Ebooks Download. [accessed 2016 Nov 16].

10. Laminated Object Manufacturing. Rapid Prototyping Center: Laminated Object


Web. 16 Nov. 2016.
The Optimization of Weightlifting with Basic Physics
Shane Davis 4/09/17
Physics 112 11247
Professor Michael Swingler
Physics can be effectively used to optimize the weight training of an individual in the pursuit of both aesthetics and strength. A torque (distance)(mass)(gravity) allows forces around a pivot point to be calculated and explains the design of quality Olympic equipment. Work, the energy required to produce a force over a distance, creates an easy way to measure intensity based on the energy expenditure of the muscles. Power, defined as work over time, is an even more accurate representation of intensity as it takes into account the time it takes to do work and can be used to determine an average energy output throughout a workout. More power generated will correlate to a higher amount of calories burned and shows that weightlifting plays a much more important role in fat loss than previously understood. The musculoskeletal system is made up of many different levers that allow the body to move and produce force. The effort that must be produced by the muscles is greatly affected by their insertions on the skeletal structure in accordance to the function of the lever. A bicep muscle in some cases must produce an effort that is 7 times the amount of the load providing resistance on the end of the lever, in such a case the palm of the hand. These concepts and more provide clear guidelines on how to use physics to the advantage of a lifter and optimize their performance and training.

Hitting the gym is gaining momentum in the US and around the world. People from all different walks of life are realizing the benefits of exercise and incorporating it into their schedules. Increased energy, fat loss and muscle gain only scratch the surface of what an individual will experience if they decide to trade out a few hours a week of free time for a few hours in the gym. While some people simply want to maintain their weight and enjoy a looser diet, there are some dedicated gym goers with physique and strength goals. These dedicated lifters model their training after professional bodybuilders and weightlifters depending on their goals. Bodybuilding and weightlifting officially started roughly in the 1900’s and “on the 16th of January 1904, the first large-scale bodybuilding competition in America took place at Madison Square Garden in New York City.” During this time, weightlifting and bodybuilding were performed almost exclusively by professionals. Both of these practices involve training with weights and performing movements such as the squat, bench press, deadlift and overhead press. “In the 1970s, bodybuilding had major publicity thanks to Arnold Schwarzenegger, Franco Columbu, Lou Ferrigno and others in the 1977 film Pumping Iron.” Bodybuilding and weightlifting have since the 1970’s become much more commercial and recreational and today the amount of amateur lifters and bodybuilders is quite extensive. The range of participants in lifting and bodybuilding today can be likened to baseball where both professional players on the top and little league players on the bottom exist. With more and more amateurs exercising techniques of professionals, some variability in effectiveness and proper execution is evident. The recent explosion of so-called “Gym Fail” videos supports this claim. While this is not an instructional guide on proper form and technique, understanding the correlation of physics and lifting weights can shed light on the reasoning behind the correct form and technique.

The involvement of physics in lifting weights is and should be obvious to most, but some may see a lifter performing a deadlift and think it is mindless and brute. The very design of the modern equipment used, Olympic barbells and weights, is derived from the laws of physics. An Olympic barbell, seen in Figure 1, is 7.2 feet or 2.2 meters long and weighs 44 lbs or 20 kgs. In most cases, 45lb or 20.4kg weight plates act as the largest increments for increasing the weight
on the bar. This increment was not accidentally chosen, it was calculated. Most squat cages, bench presses and squat racks support the bar at two locations 46 inches, or 1.16 meters, apart. Weight plates, which are .0317m wide are placed on the sleeves as close to the pivot point as .122m. This results in the first plate applying its force .137m from the pivot point assuming the plate’s center of gravity is in its direct center. Variations in the center of gravity of plates would create trivial changes in calculations. The second plate placed on that same sleeve would apply its force .168m from the pivot, the third at .2m and so on. Applying the basic rules of torque, lever arms and gravity, it can be determined that three of these 20.4 kg plates can be safely placed on one side of an Olympic bar if the supports are 1.16m apart. Even an extra 4.5 kg plate could be slid on without worry. The torque in one direction of an object must be equal to the torque in the other direction if the object is to be balanced. The forces that combine to equal the torque on one side are calculated by multiplying the kilograms of the force times the gravity constant, 9.8, times the distance in meters from the support or pivot point. When placing weight on only one side of the bar the support closest to that side acts as the pivot point. The right sleeve and support loaded with weight are shown in Figure 4. The weights create a force in the clockwise direction while the bar, since its center of gravity is on the other side of the pivot point, creates a force in the counterclockwise direction. With the support .58m away from the center of gravity, the bar creates a force of about 113 N*m. On the opposite side of the pivot point, three 20.4 kg plates and a 4.5 kg plate create an opposing force of about 111 N*m. The next heaviest Olympic plate weighs 11 kg and if used instead of the 4.5 kg plate would cause the clockwise force to reach 113 N*m and flip the bar. This setup ensures that lifters can safely rack and un-rack their weight plates without worrying about the bar flipping, causing a scene and possibly causing injury as well. It is rare a lifter will need to increase the weight on each side of the bar by over 60 kg as calculated here and as a result there is plenty of cushioning between the weight it takes to flip the bar and the weight that will frequently be placed on each side at a time. Non-Olympic barbells range from 1.5 to 2.2 meters long and are between 7 to 10 kg in weight. These bars can be difficult to work with and often an individual with a flawed understanding of basic physics will make the mistake of creating too much torque on one side of the bar. A professional lifter would rarely be seen using this type of equipment because it was not designed according to the laws of physics.

It is beneficial to realize the physics behind the design and calibration of weights and barbells but even more so is understanding how physics can optimize the usage of these tools. While it is unfortunate, there exists a plethora of misconceptions and twisted ideals regarding lifting and working out in general. Coaches and trainers adopt different training methods and styles and emphasize on what they believe is most important. This creates a mixing pot of claims and opinions that are constantly changing and can be easily mistaken for scientific facts. Fortunately physics does not change and some important constants can be defined and used to measure efficiency. Anytime a lifter creates a force over a distance, they are performing work. Work is mathematically defined as force multiplied by distance. In a simple exercise such a bench press, the distance can easily be measured as well as the amount of force required to move the weight used in the exercise. If a lifter was to lower 100 kilograms a measured distance of .5 meters in a bench press and then push the weight back into the starting position the work done would equal 980 Joules. This is because force is being exerted over the .5 meter distance during the eccentric (lowering) phase and the concentric (pressing) phase. This concept may be difficult for someone who has performed a bench press before to understand. To a lifter, the concentric
phase is the most difficult and feels as if more work is being done than in the eccentric. The reason for this lies in the muscles of the human body which can produce more force during this eccentric phase than the concentric. Eccentric force was found to increase “by around 25 – 50% when measuring strength in living humans.” Because human muscles can create more force when they contract this way, the lowering phase of the bench press requires less effort but not less work because the weight feels lighter to the muscles. Work is represented by joules and should be converted to kilocalories for practical purposes. Kilocalories are the typical measurement used to determine how much energy muscles are using or how much energy food contains. 1 kilocalorie is equal to 4184 joules. There is more information needed to accurately calculate calories burned from an activity and work on its own does not provide very much information. But power, which is defined as work divided by time, is a very useful value for measuring intensity. The first application of power would be the level of intensity that a lifter is performing a certain exercise. If a lifter executes a 150 kg deadlift and the total distance the bar travels in one repetition is .6 meters, then the amount of work done is 1764 joules per rep. If it takes the lifter 5 seconds to complete one repetition and they completes 5 in total, then the total amount of work is 8820 joules over the 25 seconds. Power then is equal to 352.8 J/s. The power equation, work/time, illustrates how a lifter can then increase intensity. There are three applicable ways to do this: increase the weight, reduce the time taken to complete the reps or increase the number of reps. It should always be the goal of a lifter to increase the value of power produced during an exercise and ultimately over the duration of their entire workout. The power equation is such a valuable tool because it is so flexible and can be applied to any workout no matter what the goals of the athlete are. If the purpose of a workout is for strength, the weight used will be heavier and the repetitions will be lower. If the purpose is for muscle hypertrophy or muscle endurance, the repetitions will be higher and the weight will be lighter. In both of these cases however, the power should be roughly the same. The aforementioned lifter performing a deadlift had a power value of 352.8 J/s. If that same lifter performed a deadlift with 95 kgs, and did 8 repetitions over the same 25 seconds, their power value would be almost the same at 357.5 J/s. This means that although the lifter was handling a lighter load, 95 kg instead of 150 kg, they performed the exercise at a faster pace and thus did more reps to achieve the same level of intensity.

A lack of intensity is most commonly seen when an individual claims to be performing high repetitions on an exercise. They will usually be seen doing a set of 20 plus bicep curls with poor form and a weight much too light for them while texting on their phone or watching the news. Performing a higher number of repetitions is great for developing muscular endurance, but it is not supposed to be easy. The power equation can be used to ensure that the same level of intensity is being reached even though the weight is lighter. Performing a bicep curl with 20 kgs traveling in a .5 meter arc will generate the equivalent 196 joules of work. If it takes a lifter 3 seconds to perform a bicep curl with 20kgs and they do 5 their power output would be 65.33 J/s, with a total of 980 joules over 15 seconds of work produced. If that individual wanted to do 20 repetitions at a pace of 1.5 seconds per rep, they would need to use 10 kg to achieve the same power level of 65.33 J/s during the exercise. In this case power is used to compare the intensity of two sets. Average power output over the duration of an entire workout can be calculated as well and would take into consideration rest intervals between sets. This is important because even if an athlete is working hard during their sets, it is not optimal to sit around for extended periods of time in between the sets. This of course depends on the exercises being performed and
how. If a lifter is doing single, max effort deadlifts, they will need to rest more in between their
sets than someone who is doing sets of pushups. However, a lifter will find it difficult to attain a
high amount of power output when doing max effort single reps. Performing 3 to 8 repetitions is
much more optimal than singles if training for strength, and the power equation reflects this.
Another conclusion that can be drawn from the power equation is that some exercises have
obvious differences in power. This varies due to which muscles the movement involves and how
much weight can be handled. A deadlift involving primarily the legs and back can be performed
with much more weight than a bicep curl that involves only the biceps brachii in isolation.
Therefore because a deadlift involves more muscles and more weight than a bicep curl, more
power is being generated. While power does not give raw values of calories burned, power does
represent energy usage of muscles. The more muscles the body is using the more energy required
to sustain them. Usually, when an individual is looking to get in shape and lose fat, they turn to
 crunches, isolation exercises like the bicep curl and the treadmill. While cardio is an important
part of fat loss and increased fitness, crunches and isolation exercises are nearly worthless.
“Research suggests that using free-weight, multijoint exercises such as the squat, bench press,
shoulder press and bent-over row maximizes the number of calories burned compared to
machine exercises or single-joint isolation moves.” The muscles involved in these exercises are
small and do not have high energy expenditure or high power output. “Too many women (and
men, for that matter) shun the weights when their goal is to drop pounds and lean out, but in
doing so they’ve eliminated one of the best ways to achieve their objective.” They do not think
that performing compound lifts like the deadlift or squat will help them lose weight. The truth is,
fat cannot be spot reduced. In other words, doing crunches will not burn belly fat, and exercising
the arms will not burn fat on the arms. The body distributes fat as needed and reducing the total
amount is the only way fat can be lost. Compound lifts allow an individual to produce more
power than any other exercises and therefore are very efficient in burning calories if proper
repetition ranges are used.

The entire skeletomuscular system is comprised of levers. A lever can be defined as “a
simple machine that can be regarded as a rigid bar that can turn about a pivot.” The
understanding of levers explains the leverage advantage and disadvantage of different
movements and reveals strengths and weaknesses of the human body. The bicep joint is simple
and illustrates the concept of a lever very well. A lever system is comprised of an effort, fulcrum
and load. As seen in Figure 2, the bicep muscle forms the effort, the elbow acts as the fulcrum
and the load is located in the hand. The type of lever here is a third class lever. “In a third-class
lever, [shown in Figure 3] the effort is applied between the load and the fulcrum. These levers
are speedy and always operate at a mechanical disadvantage.” Unfortunately most of the joints
in the human body are third class levers, and that means the muscles must create much more
effort in comparison to the load. For example, the distance between a person’s elbow and hand is
.35 meters and their bicep inserts .05 meters from the elbow. If they were to hold their arm at a
90° angle with 15 kg’s in it, their bicep muscle would have to create a force of 1029 N,
equivalent to 105 kg’s, to match the force of the load! If it were not for the huge lever
disadvantages between the muscle and bone of human bodies they would be much stronger. In
fact, better leverages are one of the reasons why most animals are stronger than humans,
especially when compared to their body weight. The equation to determine the effort in
accordance with the load in a lever is Fe=(Fl x Dl)/De. The force of the effort (Fe) is equal to the
force of the load (Fl) multiplied by the distance between the load and the fulcrum (Dl) all divided
by the distance between the effort and the fulcrum (De). The determining factor for the leverage of the biceps brachii is the location of the insertion of the muscle on the radius. Small changes in the location of this insertion point can have drastic effects the effort required by the muscle. By looking at the equation, $F_e = \frac{F_l \times D_l}{D_e}$, it can be seen that the larger the value for $D_e$, the smaller the value for $F_e$. Therefore if the insertion point is further down the arm away from the elbow, it will create a mechanical advantage and reduce the effort needed to move a certain mass. If the insertion point moves the other way, toward the elbow joint, then it creates a mechanical disadvantage and requires more effort to move a certain mass. For instance, take two lifters with same arm lengths but different insertion points. In the arm of one lifter the bicep attaches .038 meters from the elbow ($D_e=.038m$) and in another it attaches .030 meters from the elbow ($D_e=.030m$). They both have an arm length of .33 meters and hold 10 kg’s in their hand while maintaining a 90° angle at the elbow. Each bicep must produce a different amount of effort ($F_e$) to hold this position. The first lifter ($D_e=.038m$) must produce a force equal to 788.5 N or about 80 kg’s ($F_e=788.5N$) to maintain this position. The second lifter ($D_e=.030m$) must produce a force equal to 998.8 N or about 102 kg’s ($F_e=998.8N$) to maintain this position, more than 20 kg’s more! This small difference of .008 meters caused a 20 kg difference in effort required by the bicep muscle. A more common variation between lifters is limb length ($D_l$). In the leverage equation, $F_e = \frac{F_l \times D_l}{D_e}$, a higher value for $D_l$ correlates to a higher value for $F_e$. This means that those who have longer limbs are likely to be at a mechanical disadvantage compared to others depending on the location of their insertion points. If comparing two lifters with the same insertion points but different limb length, the one with shorter limbs will have a mechanical advantage over the other. For example take two lifters that have an insertion of .038 ($D_e=.038m$). One lifter has an arm length ($D_l$) of .35m and the other an arm length of .40m. The lifter with the longer arm would need to produce 12.5% more effort ($F_e$) than the lifter with the shorter arm. This explains why those with smaller stature are usually stronger than others who are lengthier even if the two persons carry a similar amount of muscle and bodyweight.

An expansion of the previously explained concepts leads to another important factor to keep in mind when lifting weights. Why is it that an object will feel heavier and heavier the further it is held from the body? The application of torque and the center of gravity of the body will answer this question. The further someone holds an object from their body, the further it is from their center of gravity. This creates a greater force of torque on the muscles that keep the body upright. The center of gravity, or center of mass, of the human body can be calculated using only a scale and the equation $F_1 \times L_1 = F_2 \times L_2$. To find a person’s center of mass, they must situate themselves in the pushup position with their hands on a scale directly under their shoulders and they must know their total body weight. $F_1$ is equal to the force in kilograms under the person’s hands in the pushup position. $L_1$ is equal to the distance in centimeters from the palms to the toes in the pushup position. $F_2$ is equal to the bodyweight of the person and $L_2$, is the distance from the pivot point, the toes, to the center of gravity. If a subject that weighs 85 kilograms creates a force of 62 kilograms on a scale and the distance from their toes to where their palms touch the scale is 147 centimeters, then their center of mass would be 107 centimeters from their toes. This location, depending on the individual, will be located around the belly button, either above or below it. This example and calculation is shown in Figure 5. Knowing that the center of mass is near the belly button, it makes sense that the further an object is held from that location, the greater the amount of torque it produces on the muscles that are working against it to keep the body in the desired position, usually upright. A deadlift is a perfect example of how proper
technique means keeping the weight as close to the body as possible. When handling a heavy weight, a few centimeters further or closer to the body can make significant differences in how heavy that weight will feel to the muscles handling it. For most lifters, one of the main goals of working out is increasing strength and using more weight in their lifts. One of the best ways to increase the weight used in an exercise is to exploit mechanical advantages and avoid mechanical disadvantages to the greatest capacity. In Figure 6, the proper form of a deadlift is shown in three stages. Notice that the bar is in contact with the lifter in all three stages of the lift. The bar is also traveling in a straight line throughout the lift, not zigzagging across his body. This is important as well because the shortest path between two points is a straight line. A lifter will reduce the amount he is able to lift if the bar is not moving in a straight line because he will have to do more work to move the weight from point “A” to point “B”. “There is no rationale to produce a horizontal force upon the earth. Any horizontal vector will be wasted energy that could have been added to the vertical vector.”

Therefore a perfect deadlift is executed with the bar as close to the body as possible with the hips and knees situated so that they do not get in the way of the straight path of the bar. Because there are so many muscles involved in the deadlift, the change in force of the weight felt by each will be different. Because of this the center of mass of the body is conveniently used. This works similarly to the increase of force each weight plate put on one end of a bar when placed further than another. If a load of 150 kilograms is held .1 meters from the center of mass, it will create a force of 147 N opposite that center. If that same load is held .127 meters from the center of mass, the force would be 186 N at the center of mass. This is a big difference for such a small amount of variance in distance. This means that even a slight loss of control during the movement could make it suddenly much more difficult and cause the athlete to fail the lift. Not all of the muscles are located at the center of mass, in fact many are not. Usually they situate on the outer edges of the body on either side. The muscles primarily used in the deadlift, the hamstrings and back are on the opposite side of the body that the load is being held which creates an even larger mechanical disadvantage for those muscles. This makes it even more important to keep the bar tight against the body during the lift for maximum performance.

Clearly it can be seen that applying physics to your lifting has many more advantages than simply impressing your science geek friends. Although this is a definite bonus. At the very least an appreciation for the thoughtful design of olympic equipment should be adopted. The understanding of torque and forces makes us think twice before mindlessly loading up 20.4 kg plates on one side of the bar for the guy on the football team with oversized genetics. Olympic equipment really is the most optimal to use because of its standardized weights and barbells and conveniently it is widely available to those wishing to use it. In accordance to the power equation (Power = Work/Time) to be more powerful, lift as fast and as heavy as possible to generate the most power, while of course maintaining good form. At the same time, training like this will help burn fat and speed up your results. Reducing rest periods to maximize your power generated will provide even more fat burning benefits as well as challenge and increase your stamina. Leverages help us understand that some mechanical advantages and disadvantages, like limb lengths and muscle insertions, cannot be changed. Others that can be controlled we make an effort to exploit as much as possible as in the case of deadlifting to keep the bar as close to the center of mass as possible.
As bodybuilding and weightlifting continue to evolve it is my hope that an average individual’s training routine will become more and more optimized. Obviously not everyone can afford the highest grade equipment and best trainers, and many don’t have the desire to put that much effort into their fitness in the first place. Basic fitness information however should be less cluttered with misleading advice and consist more of solid, science backed concepts. Many people are slowing or even halting their progression by following advice from people who simply don’t know what they are talking about. The concrete belief that lifting weights is exclusively for bulking up and running only for slimming down is misleading and should be let go. A combination of both with good programming will provide results that surpass anything that could be achieved with only one of the two.
Figures:

Figure 1: Olympic Barbell Measurements

Original Material

Figure 2: Bicep Joint

http://www.simplescience.info/science
Figure 3: 3rd Class Lever

Engineeringtoolbox.com

Figure 4: Comparison of Torques

Original Material
\[
F_1 \times L_1 = F_2 \times L_2 \\
62 \text{ kg (147 cm)} = 85 \text{ kg (L_2)} \\
9114 \text{ kg (L_2)} \\
85 = 85 \times \frac{L_2}{L_1} \\
L_2 = 107 \text{ cm}
\]

**Figure 5: Center of Mass**

*Original Material*

**Figure 6: Proper Deadlift Form**

[https://stronglifts.com/deadlift/](https://stronglifts.com/deadlift/)
References


6. Stoppani J, Weubben J. Lift to burn: blast away fat by hitting the weights the right way. 2011 [accessed 2017 Apr 14]. go.galegroup.com/ps/i.do?p=AONE&sw=w&u=mcc_main&v=2.1&id=GALE%7CA260874641&it=r&asid=c7c9a7c8e59168c1243ee44b94dab758


The Human Heart: The Physics of the Cardiovascular System, Causes of Lost Pressure and the Maintenance of Homeostatic Balance

K. Dennis-Mohler
April 20, 2017
Physics 112
Dr. Casey Durandet
Abstract

Physics occurs in everything from cars being driven to the blood traveling through the human body. The heart and cardiovascular system, are responsible for delivering blood, hormones, nutrients, and water to all of the organs the body depends on to sustain itself. Without this, the body would not exist. In order to accomplish the daily goal, the heart has to withstand different pressures, blood must be able to flow at any given angle and electrolytes needs to maintain a homeostatic balance at all times. This paper will discuss how the heart and cardiovascular system responds to certain situations using physics.

The heart and its function

The heart is one of the main, if not the main, organ responsible for keeping humans alive. It has been noted that the Greek philosopher Aristotle said, “The heart is the only organ of all organs that cannot endure injury. This is expected because when the main power source is destroyed, no additional force can be brought by the organs that depend on it.” 1.

The heart is a muscular organ that is apart of the cardiovascular system. Its contractions are responsible for moving blood throughout the body and is important for key life sustaining necessities, such as hormones, immune functions, nutrients and oxygen to the cells. This transportation system carries wastes like carbon dioxide away form the cells to all other organs responsible for its elimination from the body2. It can move more than five quarts (or 4.73 liters) of blood through the body each minute; this is equivalent of approximately 2,000 gallons (or 7,580 liters) per day2. The average heart beats approximately 72-80 beats per minute2, or 103,680-115,200 beats per day (37,843,200-42,048,000 beats per year). It can weigh from 250 to 350 grams2. A female hearts weight can range from 250 to 300 grams (9 to 11 ounces) and a male heart can weigh 300 to 350 grams (11 to 12 ounces)2.

The heart is so unique it has its own circulation system, and muscle tissue, cardiac muscle. The individualized system responsible for blood flow within the heart is the coronary circulation. This circulation is considered the functional supply of the heart3, its how the heart is nourished without pulling from the “transported nutrients” the blood receives. The myocardium that encases the heart is very thick, it is too thick for diffusion, another system the body uses to deliver nutrients to parts of the body, to deliver the nutrients the heart needs. The coronary circulation is considered the functional supply of the heart3. These vessels within this circulation system bring oxygenated blood to the heart tissue itself2. The cardiac muscle is unlike any other muscle within the body. The muscle fibers are striated and contract by sliding filament mechanism, like skeletal muscle fibers. However, skeletal muscle fibers are long, cylindrical, and multi-nucleated. The cardiac muscle fibers are short, fat, branched and interconnected3. The intercellular spaces have endomysium (connective tissue matrix) filled in between them. This matrix is connected to the cardiac skeleton allowing the cardiac cells something to pull or exert their force against; acting as both tension and insertion3.

The cardiac muscle has to be different from the other muscle fibers in the body due to the stress and work it endures. These fibers respond to a rhythmic contraction enabling the muscle tissue to respond to the electrical impulses that regulate the beats of the heart. This muscle does not depend on the body’s nervous system for the electrical impulses, it has its own system. Autonomic nerve fibers supply a healthy heart so that it can alter its rhythm as needed. For
exampled, transplanted hearts are cut from their nerve supply and placed in a new body completely severed of its nerve supply, and the heart will continue to have its own rhythm. The intrinsic cardiac conduction system consists of noncontractile cardiac cells specialized to initiate and distribute impulses throughout the heart, so that it depolarizes and contracts in an orderly, sequential manner. There are several cells called the cardiac pacemaker cells (autorhythmic cells) that are responsible for the electric impulses within the cardiac conduction system. The sinoatrial (SA) node, atrioventricular (AV) node, atrioventricular bundle (also called the bundle of His), the right and left bundle branches; and subendocardial conducting network (also known as the Purkinje fibers) are the cells that make up the cardiac pacemaker cells.

The first electrical signal, the SA node, causes the atria to contract, and that contraction sends the blood from those two chambers into the two ventricles. This node typically generates impulses about 75 times every minute. This node sets the pace for the heart, its the hearts pacemaker, and gives it the sinus rhythm. The signal then passes down to the AV node via gap junctions and the internodal pathway. These cells are located at the base of the right atrium, right above the tricuspid valve. These nodes response is delayed by 0.1 second to allow the atria time to respond, and complete their contraction before the ventricles contract. The AV node has been liked to a traffic light because it must await the response of the atria before it can proceed. The impulse then travels to the AV bundle, found in the superior part of the interventricular septum. The AV bundle is the only electrical connection between the atria and the ventricles, there is no other connection. The right and left bundle branches are in charge the impulses through the interventricular septum, and the subendocardial conducting network depolarizes the contractile cells of both ventricles. The electrical activity of the heart is measured by an electrocardiography, also called a ECG.

The cardiac cycle is considered one full heartbeat that is divided into two phases: systole and diastole. If the blood flow is followed coming into the heart, it would start at the right atria receiving deoxygenated blood. Systole is the first phase started in the atria followed by diastole of the atria, followed by ventricular systole and diastole. As the blood moves from the atria it must pass through one of the two valves in the atria (the right has the tricuspid valve and the left the mitral valve). The systole phase occurs while the heart is contracting. This contraction is the shortening of the myocardial muscle fibers. Systole of the ventricles cause blood to surge out of the heart and into the aorta and pulmonary artery. The diastole phase causes the heart to relax and the myocardial fibers to lengthen. As the heart expands, or dilates, the cavities fill with blood, and this continues slightly before the ventricles go through the diastole phase.

Blood flow starts from the heart, and is pumped by force to various parts of the body. The arteries carry oxygenated blood away from the heart (left side of the body) and the veins bring deoxygenated blood back to the heart (right side of the body). The heart is often referred to as a “double heart pump” that is separated into four chambers; the right atrium and ventricle, and the left atrium and ventricle.

The left ventricle chamber is compressed by heart muscle, $1.7 \times 10^4$ kPa≈128 mmHg, and blood is pumped through the aorta into the body’s largest cycle, the systemic cycle. The systemic cycle supplies approximately 84% of the body’s blood volume, while the remaining 16% is divided between the pulmonary cycle and the coronary circulation system (10% pulmonary and 6% coronary circulation). Any fluid driven by a pump through a circuit of closed channels operates under pressure, and the closer the fluid is to the pump, the greater the pressure exerted on the fluid. Blood flow is in a closed circuit system and would be viewed as such. The blood
leaving the aorta will experience the greatest pressure because it is closer to the heart, which is why it is important that the aorta is one of the largest vessels in the body. Blood flows through the vessels along a pressure gradient, always moving from higher to lower pressure areas. Figure 1 shows the pressure at the aorta is increased and the further blood gets from the “pump” the lower the pressure. The highest point on the graph is generated by ventricular contraction and this contraction is known as systolic pressure and it averages about 110 mmHg to 120 mmHg in healthy adults. During the diastolic pressure, the aortic valve closes, which prevents blood from flowing back into the heart. The walls of the aorta (and other elastic arteries) shrink in a recoiling like fashion, maintaining sufficient pressure to keep the blood flowing to the smaller vessels.

Physics of blood flow

The relationship between blood flow, pressure and resistance all relate to the physiology of blood circulation. Blood flow (F) is directly proportional to the difference in blood pressure (ΔP) between two points in the circulation, hydrostatic pressure, gradient. When ΔP increases, blood flow speeds up and when ΔP decreases blood flow declines. Blood flow in inversely proportional to the peripheral resistance (R) in the systemic circulation; if R increases, blood flow decreases. The relationship can be expressed as: F = ΔP / R.

The resistance (R) of the flow (F) is determined by the pressures (ΔP) between two point in a tube: 5R = ΔP [mmHg] / F [L/s]

The pressure in the aorta drops to its lowest level, approximately 80 mmHg in healthy adults, and this is measured as the diastolic pressure. As the blood flows from the aorta through the major arteries, the small arteries, the capillaries and the veins to the right of the atrium; the pressure drops from 100 mmHg to nearly zero. If the flow rate is approximately F ≈ 95 ml/s the total resistance is:

R_{total} = ΔP / F = 100 mmHg / 95 ml/s = 1.05 x 10^4 Pa / 0.095 L/s = 1.4 x 10^5 Pa·s/L

The resistance of a system of tubes can be calculated using the tubes in series and tubes in parallel. This arrangement is due the different number of arteries and veins the body has that branch off into many different areas and the organs being arranged in parallel (Figure 2); one tube in series (Figure 3), tubes in parallel (Figure 4).

Hydrostatic pressure is the force exerted by a fluid across a surface of unit area. This definition applies to blood traveling through the different vessels. There is an exchange that happens in the of blood and interstitial fluid that takes place in the capillaries. The blood cells that travel through these microscopic tubes push out the interstitial fluid into the interstitial spaces and causing the blood pressure to be lowered. Pressure can be expressed mathematically as P = x h x g; where P = pressure (Pa), x = density (kg/m^3), h = height, g = acceleration due to gravity (m/s^2); g = 9.8 m/s^2.

When a fluid is in motion it is said to be streamline, or laminar. Every particle that passes a particular point moves along exactly the same smooth path followed by previous particles passing that point. Fluids are substances that are incapable of preserving their form in the presence of a force. Blood is a fluid but in that fluid is a mixed of other materials and the
The combination makes it viscous. In order for blood flow to occur there must be a pressure gradient along the tube, Figure 5. In order to calculate the hydraulic resistance found in the blood vessels the Hagen-Poiseuille equation (Figure 6) can be used. It shows the constant relationship between flow and pressure or that the flow is a linear function to the change in pressure. Two streamlines could not cross each other’s path and continue this steady stream. If the flow becomes irregular or if the velocity of the fluid changes at any point, it is considered turbulent. This is very important dealing with blood flow. Any type of sudden change or interruption to the steady stream could cause a person serious damage to their body.

Work performed by the heart

Blood flows while a person is at any angle this is due to gravity and also to laminar flow. The heart has to work harder when a person is standing, exercising or standing upside down; but when a person is sitting or lying down, it does not work as hard. When the body is in motion, the heart puts in work to maintain steady pressure. Work is defined as work done only if an object is moved through some displacement while a force is applied to it. It can be define mathematically as \( W = F \cdot d \), where \( F \) is the magnitude of the force acting on the object and \( d \) is the magnitudes object’s displacement. This equation could be used since the heart is a muscle and it is experiencing force and blood travels or is being displaced. However, since the body is a closed system and the heart is essentially a pump, the work done on the heart could also be solved using \( W = P \Delta V \), where \( P \) is pressure, measured in mmHg and \( \Delta V \) is \((V_2 - V_1)\) or volume, volume of blood pumped during each compression measured in liters. This equation can be derived from the earlier equation. As the heart contracts, the volume of blood decreases in the respective chamber. The increase in pressure closes the inlet valve and opens the outlet valve.

Losses of blood pressure

There are several instances that can cause a loss of blood pressure to occur within the body. The majority of those instances would start with turbulent flow. Unlike laminar flow, fluid in motion moving stream like, turbulent flow is the opposite, resembling waves or swirls. When the velocity of a fluid increases above a critical velocity, the laminar flow becomes turbulent flow (Figure 8). The critical velocity for the turbulent flow in a tube depends on the viscosity, density and average velocity of the fluid as well as the radius of the tube. The flow condition is characterized by Reynold’s number. The flow will be laminar if the Reynold’s number is less than 2000, it will be turbulent if it is greater than 3000. The region between 2000 and 3000, the flow is unstable, meaning the flow can move in a stream like motion, but any small disturbance will cause its motion to change to turbulent flow. If there is a constriction in a vessel, blood will no long be able to flow in a stream like manner. This causes the blood to move faster through the constricted area, both speed and resistance will increase with the lower cross sectional areas. The cost of increased speed requires more force and strong contractions. If there is an aneurysm within a vessel, the vessel is stretched and the walls thinned due to the increased pressure in that area. There is laminar flow toward the center of the vessel and the flow is slower, but toward the thinned section of the vessel; turbulent flow occurs creating a vortex within the vessel (Figure 9). The turbulent flow puts additional pressure on the vessel, which cause the heart to work harder. An aneurysm is one of many
examples that causes loss of blood pressure. It would be the same for atherosclerosis or any
disease that would restrict blood flow.

Heart failure simply means, the heart is unable to keep up with the workload to sustain
the body. The clinical definition for congestive heart failure is: a chronic, progressive condition
in which the heart muscle is unable to pump enough blood through to meet the body’s needs for
blood and oxygen. When the heart is unable to keep up with the its “workload”, it will
compensate for it in other ways. It will do so by enlarging the heart chambers, developing more
muscle mass and pumping faster (or working harder). The body will try to compensate by
narrowing the blood vessels or diverting the blood away from less important tissues and organs,
like the kidneys, to the heart and brain.

Electrolytes Maintain Balance

Electrolytes play a pivotal role in the maintaining heart function and homeostatic balance
in for the body general. They, electrolytes, are charged, chemical compounds that dissociate in
water, that conduct an electrical current in solution. They are responsible for conducting the
electrical current that causes the heart to depolarize and contract. The electrolytes commonly
found in the body are in the form of salts such as NaCl, CaCO3 (calcium carbonate), and KCl
(potassium chloride). The most common found in the body is sodium, potassium, magnesium
and calcium.

Sodium (Na+) is the primary mineral in electrolyte balance and overall homeostasis. It is
one of the most important in heart and renal functions. The regulation of Na+ balance is
inseparably link to blood volume and pressure, changes in these two variables trigger a variety of
neural and hormonal controls that regulate total body Na+ content. Blood volume is carefully
monitored and regulated to maintain blood pressure and cardiovascular function. Because Na+
content determines fluid volume and fluid volume determines blood pressure, the baroreceptors
indirectly monitor Na+ content. Drops in systemic blood pressure led to reflex constriction of
systemic arterioles, which reduces filtrate formation and urinary output (in the kidneys) and
increases systemic blood pressures. The heart and kidneys are closely interrelated and any
disorder of either of the two organs may induce dysfunction in the other organ in a spiral fashion
leading to cardiorenal syndrome (CRS). These two organs act in together to regulate blood
pressure, vascular tone, diuresis, natriuresis, intravascular volume homeostasis, and peripheral
tissue perfusion.

Potassium (K+) is the second most abundant mineral in the body. It is considered the
chief intracellular cation and is required for normal neuromuscular functioning, as well as for
several essential metabolic activities. K+ excess will cause the extracellular fluid (ECF) (internal
environment) to decrease its membrane potential, causing depolarization, often by reduced
excitability. Too little K+ in the ECF causes hyperpolarization and nonresponsiveness. K+ depletion
causes diastolic dysfunction, while high K+ protects against hypertensive endothelial
dysfunction. K+ mediates vasodilation via strong inwardly The heart is sensitive to either too
little K+ and too much K+ (hypokalemia and hyperkalemia). Both can disrupt electrical
conduction in the heart, leading to sudden death.

Magnesium (Mg) plays a role in many enzymatic processes, and it is an important
component in the mitochondrial structure and function; it modulates cellular K+ permeability and
affects calcium update and its distribution. Hypomagnesemia can potentially induce
hypercoagulability\textsuperscript{9}, can increase cardiac glycoside toxicity and can cause sudden death in patients with congestive heart failure.

Calcium is important for the electrical impulses that are responsible for heart contraction. Intracellular calcium released from the sarcoplasmic reticulum (SR) is required for cardiac muscle contraction\textsuperscript{10}. If the SR is unable to send a signal that releases calcium, it would impair contractility as the contraction of heart muscle is directly determined by the level of calcium elevation during systole\textsuperscript{10}. A similar defect would happen during diastole, it would decrease cardiac relaxation, which is critically important in that it allows the heart chambers to refill with blood in preparation for the next beat\textsuperscript{10}.

**Conclusion**

The heart and the cardiovascular system enables us to test our bodies through any activity whether its exercising, riding a rollercoaster or doing a sirsasana (headstand yoga pose). It will still “work” to keep up with the body’s demands. All of the cells in the heart, blood vessels (arteries to veins), and electrolytes play a role in maintaining a homeostatic balance. If one area is out of the normal rhythm, it will disrupt that balance. The severity of the disruption could prove fatal for someone.

Before researching this topic, I thought I was well versed as to how the heart functioned. I knew about the direction the heart pumped blood through the aorta, the exchanges that take place in the capillaries, and the contractions performed by the cardiac muscles to help the veins send blood back to the heart for oxygenation. What I did not know was the physics played a major role in that as well. The amount of force it takes to pump blood through a person’s body can depend on the size of the person, but it is truly amazing is each heart is specifically designed (and wonderfully made) for that person. It performs exactly the same as anyone else’s. Each heart has the “lub-dub” rhythm that beats 24/7/365.
Figures

Figure 1: Systemic blood pressure in various blood vessels

![Blood pressure graph](image1)

Figure 2: (left) Gas exchanges that occurs during the blood flow; (right) direction of blood flow follows in series and parallel throughout the body

![Blood flow diagram](image2)

Figure 3: Example of a blood vessel in a series and the equation for a tube in series

![Blood vessel diagram](image3)
Figure 4: (top and bottom pictures) Example of a blood vessels in parallel and the equation for a tube in parallel:

\[ R = R_1 + R_2 + R_3 \]

\[ \frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} \]

\[ R = R_1 \cdot R_2 \]

Figure 5: Example of vessels with laminar flow. The larger the flow without obstruction the easier the flow can move.

Figure 6: Hagen-Poiseuille equation; used to calculate the rate of flow within a cylindrical tube.
Figure 7: Illustration of vessels of the circulatory system:
A the artery, B the arteriole, C the capillary, D the venule, E the vein;
This represents the an example of resistors in series and parallel.

Figure 8: Graph of a change in blood flow:
This illustration shows the difference laminar to turbulent; as pressure increases, the speed increases.
Figure 9: Vortex caused by turbulent flow within a blood vessel\(^1\).
References


6. Heideman, P. Circulatory system physics of blood flow in vessels. [Internet] [Cited 2017 April 20] Available from https://www.youtube.com/watch?v=r0Glum6o1sA


11. What is heart failure? American Heart Association. [Internet] Dallas, Texas. [Cited 2017 April 20] Available from http://www.heart.org/HEARTOR/G/Conditions/HeartFailure/AboutHeartFailure/About-Heart-Failure_UCM_002044_Article.jsp#.WPa2zFLMzdQ
The Science Behind Laser Hair Removal

Chandler Dillabo

Physics 112
Dr. Casey Durandet
April 20, 2017
ABSTRACT

Hair is designed to help regulate the body’s temperature by keeping it warm and cooling it down. As humans have progressed and have come to be more civilized, the necessity for hair to keep warm or to cool down becomes less and less. Hair has now become more of a fashion sense in society rather than an evolutionary biological component is relied on to survive. Since ancient times, the act of hair removal has taken place. With today’s advancement in science and technology, hair removal can be obtained by lasers. As laser hair removal seems like a simple zap, it is far more. Laser hair removal incorporates the understanding of both human biology and optical physics.

HISTORY

The largest organ of the human body is skin and in skin, are millions of little follicles that produce hair. Hair is found all over the human body, whether it is terminal hair or vellus hair. Vellus hair is very thin, short, and has a very little to no color to it; making it almost unnoticeable. Terminal hair is the hair that most people think of when they think of hair. Terminal hair is much thicker than vellus hair, rich in pigment, and grows to much longer lengths when compared to vellus hair. Hair is designed to help insulate body heat and to protects the skin from harmful ultra violet light. Furthermore, hair functions to also cool down the body; it does this when the sweat that covers the skin and the hair sheaths evaporates. At the beginning of the human race, hair served its great purpose. As time progressed and the thrilling invention of clothes came along, the function of hair became less important as there were now cloths and fabrics to put over the external body to not only preserve body heat but to also cover oneself. Since then, as time progressed, the evolutionary and biological function of hair became less. As it became less of an importance to maintaining a homeostatic balance, removing it seemed to have little to no effect, other than a symbol or state of fashion in a given society.

The first documented civilization to practice the act of hair removal were the ancient Egyptians. Both males and females in ancient Egypt would remove the hair from their heads and also other parts of their body using razors. Hair removal for the Egyptians was practiced for personal hygiene purposes, as well as fashion in their society. The removal of hair in Egyptian society came to be extremely helpful during the hot summers and also was a way to reduce the incidence of contracting lice. Another historical group of people who practiced hair removal were the Romans. The romans would shave their heads before battle in order to prevent their hair being pulled in battle. In the aforementioned societies, hair removal was primarily practiced when hair was removed from the head; today, that is not the case. Today, hair removal is practice much more of a societal norm and serves as a fashion statement. Males and females practice hair removal utilizing the methods of shaving, waxing, tweezing, and lasers. The act of shaving to remove hair has been around and utilized since the Egyptians and waxing has been around almost as long. Laser hair removal is a much newer method of hair removal that came about in the 1960s. The term laser is not actually a word, rather, it is an acronym for “Light Amplification by Stimulated Emission of Radiation”. On July 7, 1960, Theodore H. Maiman was the first person to create a laser and he did so by utilizing a ruby as
his lasing medium and intense, high energy, flashing lights. Since then, scientists have taken Maiman’s creation, advanced it, and will continue to advance it.

THE PRACTICE OF LASER HAIR REMOVAL

Laser hair removal allows for the hair to not only be removed at the time of treatment but to be removed permanently. In order to fully understand how laser hair removal works, there needs to be a basic understanding of the integumentary system. The integumentary system is one of the eleven body systems and it is composed of the skin, hair, and nails. For the purpose of only understanding how hair is removed by a laser, only the skin and hair will be necessarily discussed in regard to the discussion of understanding the integumentary system. The skin is the largest organ of the human body. The skin covers the body in the effort to protect the body from harmful things that could damage it. The skin is also utilized to regulate body temperature and allow the body to feel sensation.

The skin is divided up into three layers; the epidermis, the dermis, and the hypodermis. A detailed diagram of the skin can be visualized in Figure 1. The epidermis is the most superficial layer of skin and is composed of five different layers, with each layer serving its own purpose and particular cells. The first, most superficial layer is the stratum corneum, the second layer is the stratum lucidum (which is primarily located in the palms of the hands and the soles of the feet), the third layer is the stratum granulosum, the fourth layer is the stratum spinosum, the fifth, most deep, layer is the stratum basal. These five layers hold important cells such as melanocytes, keratinocytes, Merkel cells, and more. The next layer of the three layers of skin is the dermis. The dermis is composed of two different layers; the papillary layer and the reticular layer. The dermis is occupied by nerve endings, hair follicles, sweat glands, sebaceous glands (also known as oil glands), and blood vessels. Those most deep, or lowest layer of skin is the hypodermis. The hypodermis mainly composed of adipose tissue (fat), as it is a storage center for it.

The dermis is where the hair follicles reside and the hair is made. Common knowledge of the hair is that is has two parts, the hair shaft and the hair follicle; however, that is incorrect. The hair follicle has five different parts associated with it; the infundibulum, the isthmus, the inferior segments, the hair follicle, and the hair follicle bulb. The infundibulum is the portion that goes from the hair follicle opening to the sebaceous gland, that way, the hair can be lubricated by oil to prevent the hair from becoming brittle and breaking. The isthmus is the portion that is in between the opening of the sebaceous gland and the point where the arrector pili muscle sits. The arrector pili muscle allows for the hair to stand up straight. The inferior segment is found between the arrector pili muscle and base of the hair follicle. The hair follicle bulge the lower part of the hair follicle that extends down. Lastly, the hair follicle bulb is the bottommost fragment of the hair follicle and is made up of hair producing cells such as matrix cells and melanocytes. Melanin is what gives hair is color but is also known as the hair’s chromophore. A chromophore is a material that is found in a specific tissue. In hair, the chromophore is melanin and melanin is what makes laser hair removal possible.
Laser hair removal is a process that involves a photothermal reaction from the dermis-epidermis matrix that causes damage to the hair follicle, while leaving the epidermis in a viable state. Even though the laser does not cause intense damage to the epidermis, it will still inflict a certain level of pain. Before laser treatment, the skin is cleansed and then a numbing topical ointment or gel is applied to the skin for about 30 minutes prior to treatment in order to make the treatment as painless as possible. As previously discussed, melanin in the hair is a chromophore. The melanin in the root of the hair is what gives the hair is pigment or color. The more melanin a body produces, the darker the tissue will be. In hair, melanin in the root of the hair is what absorbs the photons from the laser. When the melanin in the hair root absorbs the photons form the laser, it then heats up to the point where it will explode, damaging the hair follicle. This can be seen in Figures 2 and 3. It is important to note that there is also melanin in the skin, thus giving different skin color and various skin tones. Melanin has the ability to absorb different wavelengths. Melanin’s ability to do this allows a range of different lasers to be utilized. It is more difficult to perform laser hair removal on those who have similar quantities of melanin in their hair as they do in their skin.

The best case for laser hair removal is for the patient to have lighter or fairer skin (smaller amounts of melanin) and darker hair (larger amounts of melanin). In this case, the laser will have minimal impact on the patients skin due to the fact that there is little melanin in the skin to absorb the laser. The laser in this case will make the darker hair the target as it is what has the most melanin to absorb. The melanin in the hair absorbs the photons from the laser, heats up, and explodes. Those who have light hair and light or dark skin, have lower success rates when it comes to removing hair via a laser. This is because there is a lack of melanin in the hair to absorb the photons. Depending on the patient’s skin type, different lasers will be used at different intensities in order to give the best results. The chart in Figure 4 shows the various kinds of skin types and what is incorporated with each skin type. Despite the type of skin or the type of hair one has, laser hair removal is not 100% effective after one treatment for several reasons. One reason laser hair removal may not be completely effective after the first treatment may be because the intensity of the laser may not have been equally distributed to the hairs in the area that the laser was in. Additionally, depending on the patient’s skin pigment and hair pigment together, the laser may not have been strong enough to kill the hair follicles without causing physical harm or damage to the epidermis.

Furthermore, hair goes through different cycles of life and depending on the area that was treated, all of the hair may not have made it to the surface at the time of treatment. This kind of scenario would occur if the treatment was performed in areas that are shaved, such as the armpits or legs. When the shaving method is used, not every single hair in that area is cut to the same, uniform length; some of the hairs may be missed and not even get cut after shaving. After the first laser hair treatment, the hair will be reduced about 10-40%, where three treatments will show a 30-70% reduction of hair, and further, repeated treatments can reduce the amount of hair to about 90%.
THE PHYSICS OF LASER HAIR REMOVAL

Since the 1960s, more lasers have been created in order to better the laser that it is proceeding. Lasers come in many different intensities and utilize different equipment and mediums to fulfill their desired function. Despite the kind of laser, all lasers are composed of four prime components. Each laser has a laser medium, the optical cavity, the power supply, and the delivery system. The medium in the medium in the laser is typically a solid, liquid or gas. The optical cavity is located around the laser and allows for the laser to amplify, while the power supply works to excite the atoms and create a population inversion, meaning more atoms are in a higher excited states and lower states of energy. The delivery system is what delivers the light to the desired target. Some of the effective laser hair removal lasers are ruby laser (694 nm), the alexandrite laser (755 nm), the diode laser (800 nm), and a neodymium:yttrium-aluminium-garnet laser (1064 nm).

All lasers are unique in that they all utilize different kinds of mediums and are able to reach different levels of intensity. What they all have in common, it that they are all utilizing laser light. Laser light has three interesting qualities to it; they are monochromatic, and hold the characteristics of coherence and collimation. Lasers being monochromic means that the laser is made up of one, single wavelength. Coherence in a laser refers to the way that the waves in lasers move uniformly together as they travel. Collimation in lasers refers to the transmission of light; light in lasers transmits in a parallel manner without diverging from the central beam.

When a laser is being transmitted to skin tissue, the factors of irradiance and energy fluence need to be considered; both are extremely important. The irradiance is a factor that will determine the laser’s ability to vaporize the tissue or coagulate it. Irradiance can be calculated using the formula below.

\[ I_r = \frac{\text{Laser Output (W)\times100}}{\pi r^2 \text{ (of the laser beam)}} \]

Units: W/cm²

The amount of energy that the laser delivers in a single pulse is referred to as the energy fluence. The greater the number is for energy fluence, the greater its effect is. Energy fluence can be calculated using the equation below.

\[ EF = \frac{\text{Laser Output (W)\timesexposure time (t)}}{\pi r^2 \text{ (of the laser beam)}} \]

Units: J/cm²

When the energy from the laser is transferred to the hair follicle, the temperature from the laser’s energy is so great that it vaporized the tissue of the follicle causing tissue death to the follicle. With the follicle tissue dead, hair will no longer be able to produce there.
Another important calculation that physicist should make when designing and testing lasers is the LITCIT (laser induced temperature calculation in tissue) \(^9\). This calculation has been incorporated into a software system so that scientists are able to calculate the heat distribution and the absorption from the laser to the tissues. This new laser software allows those scientists who are working on new laser equipment to be able to calculate data utilizing simulations that is much more accurate than running multiple tests. The LITCIT software utilizes the equation below \(^9\).

\[
\Delta T (x, y, z) = \frac{\Delta E (x, y, z)}{\rho c_p V_{voxel}}
\]

Units: W/cm\(^2\)

In the equation, the \(T\) represents temperature and \((x,y,z)\) represent the changes in temperature. \(E\) represents the converted energy and the \((x,y,z)\) represent the changes in converted energy. Furthermore, \(\rho\) represents the density, \(c_p\) represents the heat capacity of the tissue, and \(V_{voxel}\) represents the volume of each tissue element \(^9\).

With all the information above, it is very clear to see how powerful lasers are and why one has to be trained on how to properly use them. Knowing the laser and its different intensity levels and when to use them is vital to patient safety. Using the wrong intensity on a patient could lead to a poor treatment in the sense of the laser intensity was not strong enough for that patient or that the laser intensity was too strong for a patient. Improper use can lead to patients experiencing an array of symptoms; symptoms such as burning, blistering, scarring, hyperpigmentation, hypopigmentation, and even infection \(^5\). Laser hair removal is not completely painless as it is advertised. Laser hair removal has slight side effects that typically include temporary discomfort and slight redness and swelling. Since laser hair removal requires several treatments, these treatments still need to be spread out far enough so that the skin has time to heal well enough to endure another treatment. Any kind of laser treatment should be performed by a professional who understands the machine they are using, how to determine the treatment level of intensity based on the patient, and the knowledge of how to execute the procedure.

CONCLUSION

Researching the topic of laser hair removal has been very interesting. It is very interesting to see how the scientific field of physics collides with the scientific field of biology. This research paper also included bits of sociology as well. I did not know that hair removal was performed by the ancient Egyptians but learning about it made sense. The climate in Egypt seems to get quite warm and thick heavy hair can probably come as a disadvantage during the summers. Some articles informed me that even though the ancient Egyptians shaved off their hair, the still would wear wigs. I also saw that it had a positive effect on preventing lice.

Learning about the integumentary system seemed to be a little tedious at first but then I came to realize how important it is to know the basics of skin and hair because it makes learning about the lasers much easier when there is the idea and image in mind of what exactly
is going on when the hair is being shot with a laser beam. Researching and learning about the physics of laser hair removal was also more interesting than I thought it would have been. Learning the parts of a laser was helpful in understanding the mechanics of it. It was also helpful to see the equations regarding lasers and their intensities as it gave a better understanding on how the various components work together and how they can affect one another.

Throughout this research paper, I really learned how there is a whole entire different way of looking at something when you have more knowledge regarding that subject. I made sure to incorporate every part of laser hair removal as much as possible; with only having a basic knowledge of skin and hair, fully understanding laser hair removal is less exciting.

To see how far technology has come from the first laser hair removal machine from the 1960s is truly astounding. It is amazing to see how other scientist can take something and continuously work on it to better it. It has been almost 60 years and we have gone from Theodore H. Maiman’s laser to lasers that not only remove hair better but can resurface skin, remove scars, and more. Lasers are not only used to destroy things like hair and skin, they are also used for other methods. Lasers are even used in hospitals on almost every patient because lasers are utilized to read the oxygen in blood. The pulse ox that is placed on a patient’s finger not only reads the patient’s pulse but it also uses a laser that is able to penetrate through the layers of skin to blood cells in order to read how saturated the hemoglobin is with oxygen. I know that the exploration for new uses of lasers is happening currently and will continue. With the kind of technology advancements that we are experiencing today, there is no doubt that the next new thing is just around the corner. I think that future of lasers could be used to possible be used to more accurately remove cancer cells and even help repair tissues. Honestly, the options are endless because every single day new things are being discovered, researched, and tested.
Figures

**Figure 1:**

http://humancoloringbook.blogspot.com/2015/11/integumentary-system.html

**Figure 2:**

Chilled tip cools and protects the upper levels of the skin and enhances treatment comfort.

The laser beam travels through the skin and is absorbed by the hair shaft and converted to heat.

This effectively damages hair follicles which significantly impedes its ability to re-grow.

http://www.usa.lutronic.com/img/hair_removal_follicle.jpg
Figure 3:

The Science of Laser Hair Removal


Figure 4:

<table>
<thead>
<tr>
<th>Skin Type</th>
<th>Skin Colour</th>
<th>Hair Colour</th>
<th>Eye Colour</th>
<th>Description</th>
</tr>
</thead>
<tbody>
<tr>
<td>I</td>
<td>White or very pale</td>
<td>Blonde</td>
<td>Blue, Grey, Green</td>
<td>Always burns, never tans</td>
</tr>
<tr>
<td>II</td>
<td>Pale white with beige</td>
<td>Chestnut or Dark</td>
<td>Blue</td>
<td>Always burns, sometimes tans</td>
</tr>
<tr>
<td></td>
<td>tint</td>
<td>blond</td>
<td></td>
<td></td>
</tr>
<tr>
<td>III</td>
<td>Beige to light brown</td>
<td>Dark brown</td>
<td>Dark Brown</td>
<td>Sometimes burns, always tans</td>
</tr>
<tr>
<td>IV</td>
<td>Light to moderate brown</td>
<td>Black</td>
<td>Brown</td>
<td>Rarely burns, always tans</td>
</tr>
<tr>
<td>V</td>
<td>Medium to dark brown</td>
<td>Black</td>
<td>Brownish black</td>
<td>Rarely burns, tans more than average</td>
</tr>
<tr>
<td>VI</td>
<td>Dark brown to black</td>
<td>Black</td>
<td>Black</td>
<td>Never burns</td>
</tr>
</tbody>
</table>

http://www.colaz.co.uk/fitzpatrickchart.html
Cited References

The Chemistry within Drinking Water

Hannah Dixon

April 9, 2017

General Chemistry II, CHM 152

Professor Olander
Abstract:

Water is responsible for all that exists and supplies the world with life, without water all would be lost. This paper supplies the reader with information regarding how the human water table operates. It begins with knowledge of how wastewater is treated and continues to summarize the process of drinking water. This supplies enough information to then follow the paper as it continues to explain in greater detail the chemistry that is currently applied along with the future of more sustainable treatment methods. This paper includes images of local treatment facilities as well as diagrams that help the reader to visualize the facilities operations. The goal of this paper is to inform the reader and encourage one to gain greater knowledge and appreciation for the future of our most valuable resource.

Water is the most important resource on the planet. Water gives life and freshwater specifically gives life. Yet it is one of the most wasted resources around the world. It is estimated a single household can waste up to 3,000 gallons of fresh water every year. With roughly 4 trillion cubic meters of drinking water used around the world annually the freshwater resources need to be a global concern. As a world reaching to eliminate this global footprint more knowledge and research produces new and innovative ways to recycle. Daily American’s recycle anything from paper plates to plastic bottles but the effort to recycle water has yet to become a common routine. Human’s sit under the false impression the world will never run out of drinking water.

A known fact remains the earth’s surface is 70% covered with water, but 95% of that is salt water. Now the desalination of water or removing the salt from the water can be a very simple process through filtration, the task of finding a place and purpose for the hundreds of tons of salt is a task not yet concurred. So that leaves the world with only 5% of freshwater left. It can be said the solution of water conservation can be left for the future, unfortunately such is not the case. Benjamin Franklin spoke the words “You learn the value of water when the well runs dry.” It should be a known fact that Arizona faces such issues and should encourage the effort to learn more about wastewater and how it can become fresh water.

This topic should hit close to home, Lake Mead is the main water source for much of Arizona and is suffering through a 30 year drought. Due to the lake of snow pack in the Colorado mountains Lake Mead received less than 30% of its normal intake flow last year. Lake Mead along with Lake Powell which run through the Hoover Dam contain around 50 million acre feet of water when full. Today they only fill a quarter of that capacity while the demand for freshwater grows. The town of Cave Creek that is solely dependent on water from Lake Mead can no longer supply water meters for new homes and has restricted building in the area. Humans need water to survive and have a right to a portion of the world’s water. But they don’t have a right to consume all they want and not assist the earth in regenerating more water. It is a fact we will run out. So one must wonder if so many other resources can be fully recycled why not water. Why can’t one flush the toilet and when they go to wash their hands they have that same water?

This goal of recycling water starts with understanding what is in water, it starts with the chemistry of drinking water. Wastewater comes from the earth and our bodies leading to the fact its main components are earth and human based. So returning it to its original state should not be
a simple process yet a step so commonly left out in the treatment of water. Most wastewater treatment plants produce what is commonly referred to as non-potable water, not quite drinking water but safe for agricultural use. This is all made possible by the knowledge of chemistry. This same knowledge has already been used in full swing naturally by the earth. Every drop precipitated is filtered before returning to the earth, in the same marshes naturally treat and purify wastewater. The earth has already supplied humans with the ability to recycle water, all they have to do is become motivated enough to start.

Begin by understanding the process of treating wastewater and then will further uncover the chemistry behind the process. Wastewater treatment plants can follow a series of different designs but a few things remain the same. The chemistry in all treatment facilities remain constant. Most plants start with a lift station or collection tank where all the wastewater is typically gravity feed to the start of the plant. The fact most sewage lines can be gravity feed is an important topic when talking about making treatment facilities more sustainable. The only chemistry that takes place is that of the natural world. Common elements found in the lift station are typically ammonia $NH_3$ and nitrate $NO_3^-$. These two ions make up major components of the treatment system and dictate the plants net pH level and will overall determine many of the treatment methods.

The next stage of the facility could be one or many aeration basins. The main goal of this process is to separate the solids from the clean water. There are two stages starting with air being added to the tank. There are many microorganisms within treatment facilities and can be viewed as the main operators of the facility and the performance of the facility is directly associated with the performance of the microorganisms. The types of microorganisms that exist presently depend on the age of the sludge and whether free oxygen is added to the tank. When there is a presence of oxygen bacteria that gets energy from dissolved oxygen are able to live and multiply. They operate at a high level and consume specific impurities while creating carbon dioxide and water. The aeration of tanks can also be used to remove hydrogen sulfide $H_2S$ which is harmful to both humans and the environment.

While still with in an aeration basin aerobic bacteria will fall into a stage of sleep or unawareness when the air is turned off. Now anaerobic bacteria are able to live and reproduce. They consume heavier compounds like nitrates and sulfates. Imagine yourself after thanksgiving dinner. Because the anaerobic microorganisms metabolize these heavier compounds they are slow moving which means many tank require a longer amount of time for the anaerobic bacteria to work verses the fast pace aerobic microorganisms. When the air is eliminated from theses tanks the heavy solid will settle at the bottom of the tank. The raw sludge at the bottom is then able to be pumped to a separate tank referred to as the digester. Now the “clear water” is pumped off the top and into a tank called the contact chamber.

The chlorine contact chamber is where the clear water now is treated with chlorine $Cl_2$. It is important to note the fact most of the treatment was done by microorganisms which require no external energy and are completely biodegradable and reusable. The goal is to create a plant with sewage in a state of equilibrium. In order for the aerobic and anaerobic bacteria to effectively do their job where must be a healthy form of sludge. Young sludge contains microfilaments which live short lives and leaves sludge cloudy unable to settle and be eliminated. Mid aged sludge contains cilia leaves sludge foamy and can easily turn septic. Septic sludge is routed like expired milk. Rotifer is the desired bacteria and allows the sludge to properly separate from the clear
water. Operators have to work hard to achieve this equilibrium and have a plant properly with the environment.

To be able to understand the chemistry happening within a chlorine contact chamber one must first understand the components in chlorine and how it reacts in water. Water is the main source of life and understood to be where most chemical reactions take place. This makes it extremely important to know how chlorine with react in the water and what will be the factors that can potentially change and effect of this reaction. Now when chlorine which is made up of two chloride ions $Cl^-$ is added to water $H_2O$ underchloric acid is formed. This reaction can be expressed by the following equation. $Cl_2 + H_2O \rightarrow HOCl + H^+ + Cl^-$ We are able to identify this as an acid based on the $H^+$ ion. When adding chlorine to the water in the contact chamber one must understand the $H^+$ ion is an agitator that can change the pH along with the chemical reaction. A high pH will increase the expression of underchloric acid to hypochlorite ions. This reaction can be followed by this equation.

$$Cl_2 + 2H_2O \rightarrow HOCl + H_2O + Cl^-$$

$HOCl + H_2O \rightarrow H_3O^+ + OCl^-$

This equation will then over time turn in to $OCl^- \rightarrow Cl^- + O$.

The chlorine in the contact chamber can be used in three different forms. Gas, liquid, or solid tablets. The difference between the three only depends on the size of the system. Larger plants typically use gas chlorine to maximize exposure while liquid and solid forms are used for smaller treatment facilities. Chlorine is a disinfectant and has the capability of separating the elements that make up the undesired bacteria and chemicals.

After the water has been separated and chlorinated it is ready to run through several filtration methods. The most common is sand filters. Sand filters have the ability to remove the rest of the suspended solids that did not settle to the bottom in the aeration tanks. Part of the chemistry to wastewater comes from knowing what elements categorized on the periodic table. The 10 metals required to be removed in wastewater treatment facilities are arsenic $As$, barium $Ba$, beryllium $Be$, cadmium $Cd$, chromium $Cr$, mercury $Hg$, lead $Pb$, selenium $Se$, and thallium $Tl$.

These metals are naturally heavy in nature and will settle with the sludge in the second treatment tank and most will be pumped out to the digester. The metals left behind will gather with the rest of the suspended solids which is a term referring to the sludge that is young and lose and does not settle to the bottom of the tank but still must be removed. The suspended solids and heavy metals will then be caught and removed by the sand filters. All plants require proper testing for each of these metals and if there still remains high levels further filtration maybe needed, membrane and Nano filters are a few examples. It is important to remove these metals and the bacteria not only to create sustainable and recycled drinking water but also to help reduce the possibility of contaminating the earth along with the surrounding future and current ground water.

Once the wastewater has completed the many stages of the facility it is eliminated in a few common forms. One common form is for agricultural use such as watering golf courses and watering livestock. It can also be drained into what is referred to as a lagoon which consists of a manufactured pond with a tarp like material used to prevent the water from leaching back into underground reservoirs and instead precipitated. Water that did not go through a membrane or
micro filter still contains suspended solid and is not potable. It is eliminated through sand drying beds with the sludge and left in large concrete basins and precipitates back into the water table.

The heavy sludge that was originally pumped from the bottom of the aeration tanks and sent to the digester requires different conditions. Within the digester the same gravity based settling process takes place. This method is excellent in the push to create a more sustainable world. The clear water at the top of the tank is then pumped back to the head of the treatment facility where it is then sent through the plant again for purification. This method insures operators maximize the amount of clear water they are able to produce from the plant. The same chemistry is reapplied and more clean water is generated.

The thick of the thick sludge that has repeatedly settled to the bottom throughout the process can no longer be treated. It is important to note that the more the sludge settle to the bottom it acted like a filter catching and dragging the heavy metals listed above along with the exoskeletons of the microorganisms that gave their lives to help purify the water. This sludge is then pumped and hauled to be discarded in specially design landfills. These portions of the landfill are design to dry the sludge without allowing it to leach back into the ground. At this point the sludge is considered hazardous since it contains all of the bacteria and elements we aimed to remove because they were unsafe for humans and the environment.

A new wastewater treatment plant located northern Tuscan treats the wastewater to a small community fully with membrane filters and then uses the water to irrigate crops. So what can be done to instead continue the water through drinking water treatment facilities and recycle the used water completely? This is an interesting topic with regards to isolated rule areas such as African communities and native islander who are miles from the nearest fresh water source. More than half of the wastewater that comes into any facility can be purified for potable water uses.

Drinking water treatment facilities are different from wastewater in the sense less natural and organic methods are used and the start of human influence increases. Water treatment is also different because it is heavily influenced by location. Every plant struggles through the process that no two drops of water are alike. In turn operators must be well prepared for whatever type of water comes their way. Water treatment operators must also be geologist to be able to anticipate the type of water they will be treating.

For example most of the ground water treated in the north east phoenix areas contains a high level of mercury $Hg^+$ this is simply because of its geographical location and the contents of the soil there. Where if you were to travel south to Tuscan you will find high levels of lead $Pb^+$. The fact water treatment changes based on locations makes operation at times difficult. The treatment of drinking water also provides different challenges based on whether or not the water is coming from the surface meaning the water came from either a river or lake or if the water is ground water which is the term used or water coming from underground reservoirs. Each of these different types of naturally located water sources come with different types of treatment methods.

Now drinking water treatment facilities resemble many similarities in the chemistry to wastewater. The dangerous metals and bacteria most of the time exist in a lesser concentration. Many times surface water is more difficult to treat. This is because surface water if exposed to every environmental contaminate possible. This has provoke the practice of storing drinking
water in the underground aquifers. This method has also been applied to recycled wastewater waiting to be treated as drinking water. It is an excellent opportunity to allow the earth to practice its own natural treatment methods before we handle the water.

Drinking water treatment facilities and their chemistry start with the well known chlorination process. Similar to the process before this chlorination of the water is important because the chlorine chemical has the ability to break apart the chemical bonds that form bacteria along with many liquid metals such as iron and magnesium. The water waiting to be treated is typically pre-chlorinated which increases the water detention time meaning it increases the time the water has with the chlorine. This allows more time for more molecules to be broken up by the chlorine.

The next phase is again similar in nature to the wastewater when the water enters an aeration tank. Most plants aerate the water to allow the microorganisms that live with air to digest and decompose. Then the air is shut off to allow anaerobic or anti air organisms to work and all the exoskeletons and remaining heavy solids will settle to the bottom. This method can be referred to as gravity filtration. Many times in drinking water treatment facilities a thickening agent is added to the water to increase the amount of suspended solids to drop to the bottom. Then will also commonly use a method called coagulation or flocculation and a polyelectrolyte is added to the water to “thicken” the floc or suspended solids on the top of the water that can be skimmed off.

The process then process the same through filtration and a second round of disinfection. Drinking water plants commonly use micron or Nano filters along with sand filters because the microscopically small screening has the capability of catching and restraining tiny particles of bacteria along with liquid and solid forms of metals. One new form of disinfecting drinking water is by using ozone $O_3$. This unique but futuristically common method is similar to the chlorination method in the sense the added chemical has the ability to break apart the chemical bonds that form the undesired elements. It is important to know most levels of disinfectant agent are removed through the filtration process and though the chemicals help operators to treat the water the high levels of disinfectant required are harmful as well. Once the drinking water has been fully treated it is sent to your home.

Now that some review has been made regarding the process of treating both wastewater and drinking water we will cover more of the chemistry behind drinking water.

Before the facility the water travels down the sewer line and an important chemical reaction takes place. Most of the nitrogen in the water undergoes an anaerobic reaction since there is a lack of dissolved oxygen in the pipes, meaning there is no air added to the water. The raw natural nitrogen converts to ammonia through the process referred to as ammonification. Since this reaction like many takes place in water it is also under the definition of hydrolysis. This type of reaction can be described in the following equation.

$$NH_2CN + H_2O + 7H^+ \rightarrow 3NH_4^+ + CO_2$$

The next chemical reaction to take place is described as nitrification. This reaction is organic and converts ammonia to nitrate. It is important to understand operators prefer when this reaction is stable and predictable so they may be able to anticipate whether the water required the nitrification process to eliminate high levels of ammonia or in different a denitrification process.
to eliminate a high concentration of nitrate. This reaction can be displayed in the following equation along with the equation that identifies the basic buffer required.

Alkalinity Buffer

\[ H_2O + CO_2 \rightarrow H_2CO_3 \rightarrow HCO_3^- + H^+ \rightarrow CO_3^- + 2H^+ \]

Nitrification

\[ NH_4^+ + 1.50_2 \rightarrow 2H^+ + 2H_2O + NO_2^- \]

\[ NO_2^- + 0.50_2 \rightarrow NO_3^- \]

\[ NH_4^+ + 1.83O_2 + 1.98HCO_3^- \rightarrow 0.021C_5H_7O_2N + 0.98NO_3^- + 1.041H_2O + 1.88H_2CO_3^- \]

\[ NH_4^+ + 1.90_2 + 2HCO_3^- \rightarrow 1.9CO_2 + 2.9H_2O + 0.1CH_2 \]

This example is helpful for operators to be able to calculate in pounds the amount of ammonia \( NH_4^+ \) is converted to nitrate \( NO_3^- \). These types of chemical equations are used often especially the nitrification one because nitrogen that is not converted to ammonia will never convert the nitrate and the operators would not be able to remove the nitrogen from the plant. In turn follows the equation for denitrification.

\[ 6NO_3^- + 5CH_3OH \rightarrow 3N_2 + 5CO_2 + 7H_2O + 6OH^- \]

In the final stages the reading explained how the water both non-potable and potable come in contact with chlorine as a form of disinfectant. The below equations explain how chlorine in its hydrolyzed form breaks apart unwanted elements such as sulfur dioxide and sulfite salts.

\[ SO_3^{2-} + HOCL \rightarrow SO_4^{2-} + Cl^- + H^+ \]

\[ SO_3^{2-} + NH_2Cl + H_2O \rightarrow SO_4^{2-} + Cl^- + NH_4^+ \]

Chlorine is capable of pulling apart the element, once in its pieced apart form it is typically no longer dangerous to humans and the environment. Above are a few of the common chemical reaction that take place while treating drinking water, many reaction vary based on locations and it critical that not only operator but consumers as well know the chemistry that is performed on drinking water before it enters the home.

Some of the commonly used mathematical formulas used to calculate and understand the chemistry include the reaction rate formula, change in concentration over change in time. Operators also commonly use the change in concentration with time both first order and second reaction questions. They can be expressed below.

\[ \ln[A] = -kt + \ln[A] \]

\[ \frac{1}{[A]} = kt + \frac{1}{[A]} \]

Graphs are widely used in this field as well and provide not only a great visual understanding of what is taking place in the facility but also to help revile information to other works of current or future issues. Operators also use and understand equations similar to the Henderson-Hasselbalch
equation. Knowing how elements might react with each other and create possible buffer is important because it might lead to a false treatment of water assuming the pH of the water describes what acids and bases it contains when there could really be to conjugate pairs.

Wastewater and drinking water treatment facilities both use an advanced knowledge of chemistry and how it interacts with the world. As discussed in the beginning of this reading water is a depleting source. One reason for the continual loss of the earth’s freshwater source is the fact most is being used and wasted without the proper knowledge of where it came from and where it is going. It is hard for anyone to care for and preserve something they have no knowledge of. Education is the first time to recycling water. The state of Arizona and its water source would greatly benefit from now technology that allows water to follow a complete recyclable system. In San Diego they have a waste water treatment facility where the return water that is treated by the facility is stored in both underground and above ground containers. This method of allowing the water to return to the earth before reentering the drinking water system was a method designed to allow the water to reabsorbed good bacteria and nutrients.

In conclusion I hope this paper can supply information that will intrigue the reader to further research and understand the methods used to treat water. Fresh water is so valuable and so much of it is wasted and lost in our current treatment methods. In Arizona alone so much fresh water is wasted and this waste is causing us to consume more than our share of the Colorado River. When we were children we learned to turn off the water while brushing our teeth and take shorter showers in an attempt to conserve water, but our problem has grown so much larger today and requires a solution that will change how we use water.

I feel because wastewater is such an unspoken process it is often put last when it comes to new and innovative technology. Only a small percentage of American even know how or where the process takes place and this lack of knowledge leads to uneducated decisions on how water is used day to day. If more people were made aware of how the treatment plants operate they would be able to treat water more knowledgeable and conservative. The wastewater and drinking water facilities need to be on the front of new technology to make them more efficient along with more environmentally friendly. The only way this will become a reality is if more people understand and appreciate the process.

It is time for humans to know the world they live in. time to know the life it was supplied for them and even more importantly know how to protect that life. Like all things in nature which fall back to their original state water is always move toward the process of becoming pure but we will find a day when water has been pushed to its limit, unable to be used. The only way our generation will be able to find new ways to keep our water system flowing is by learning about it and becoming educated humans who supply more than they consume.
Figures

The below figure shows Lake Mead, the white “toilet bowl” ring can be defined from sudden absence of water due to drought and over consumption.

This next figure demonstrates the use of an aeration tank at a wastewater treatment plant located in phoenix.
This next figure is a simple diagram that summarizes a basic wastewater treatment facility.

![Diagram of wastewater treatment facility]

The last figure is Tempe Town Lake which is an example of an above ground drinking water reservoir.

![Image of Tempe Town Lake]

IMAGINE ME EMPTY
References


EPA. [accessed 2017 Apr 20]. https://www.epa.gov/


Fission and Fusion

Vesna Djukic
April 20th, 2017
Physics 112
Dr. Durandet
Abstract

Fission and fusion are really important in today’s world. These two energies are completely different than one another. One uses energy and the other one does not. There are many advantages and disadvantage that are seen within these two. They both work differently and are approached differently. Therefore, fission, particularly, can be used as therapy for cancer patients.

What are Fission and fusion? The most individual’s will get from these two terminology is that they are some type of energy. Many will even question themselves and ask, “How does this apply to the world?” Indeed, this essay will explain, what are fission and fusion, the advantages and disadvantages of each, how fission and fusion works, and where is fission can be used.

Fission is known as, when a large atom is being split into two or more smaller atoms. Fusion is when two are more atoms come together to become one large atom ("What is fission and fusion?"). On this particular website, it mentions that fission does not occur in the nature and it contains highly radioactive particles. Therefore, it requires high speed of neutrons. Also, it does not take too much energy to split the atoms in the fission reaction (Libretexts, 2016). With all this said, it can be seen that Fission does not require too much work for it to occur. What this means is, it will not take a lot of energy for it to split its atoms into two or even more atoms. On the figure page, it will provide an equation for fission, along with an image of fission and how it looks like when a larger atom is split into two atoms. Furthermore, fusion will occur in such places as stars and sun. Fusion produces only few radioactive particles compared to fission. However, if it is triggered by fission, then it will result radioactive particles. In contrast, fusion requires high energy for it to bring two or more atoms together (Libretexts, 2016). With this being said, it can be seen that, since it requires a lot of energy for the atoms to come together, it can be predicted that it will also release a lot of energy because it requires a lot of energy to bring the atoms together. To add on, fusion can be easily adjusted. What this statement means is, since it produces only a few radioactive particles, it just being triggered a little bit by the fission, then it will completely result radioactive particles just as fission does. On the figure pager, it will provide an equation for fusion, along with an image of how it looks like when the atoms are being conjoined. Overall, it can be seen how fission and fusion are defined as and what they are exactly. They are different from one another by certain aspects. Just as the previous statement mentioned, fission will split a large atom into two or more, and it contracts with fusion because fusion will fuse two or more atoms into one large atom. Therefore, the amount of energy is completely different from one another and this is really important. Reason why is because it shows how much energy they will take and how much energy it will then be produce in the end. Just for an example, as mentioned before, fusion takes a lot of energy to bring the atoms into one large atom and it will also release a lot of energy.

There are few advantages and disadvantages for fission and fusion. A few advantages of fission is, free from heating generating CO2, there are harmless radiations to the environment, and the cost is really effective. Since fission is free from the heat generating CO2, it gives the
people a chance to use the energy source that is alternative free from CO2. It is there to produce electrical energy that does not produce CO2. With this being said, it will also not produce other forms of toxins in the air. An example would be, smog. Nuclear fission radiation is harmless to the environment. The nuclear power plant does produce radiations, and they do not lead to hazardous to the nature. Since it does not produce hazardous effect, nuclear fission is safe to use in the nature. The energy source is effective because the fuel that is used in the reaction of fissure is affordable. With that being said, the fission will not take too much of the fossil fuel (Advantages and Disadvantages of Nuclear Fission, 2015). Clearly, it can be seen that fission does have a few useful advantages. It is really helpful when that electrical energy can be produced without giving off too much CO2. The reason why this is advantage is because it will be beneficial for the people’s health. They will be less likely to develop any health issues. It is helpful that power plants do not lead to hazardous because if they did then fission would not be safe to use in the environment. The reason why is because it can cause toxins. The cost is really effective because the fission will not eat up too much of the fuel. Since it does not eat up too much fuel, the fuel will be used longer. Hence, it can be seen how useful fission can be in the environment. It can be used in many different aspects and not do damage to the environment.

Furthermore, there are also few disadvantages that tie with fission. They are high cost for nuclear fission plant, the reaction waste of fissure is hazardous, and the break down is harmful to the people and to the environment. When establishing the plant, it will have high costs, although the application of the energy source is affordable. Therefore, the safety is added on and it is also expensive. When the fissure is in process, the waste that is produced is hazardous. Lastly, when the breakdown is in process, in the plant, that leads to some dangers effect to the environment and to people’s health (Advantages and Disadvantages of Nuclear Fission, 2015). With all of this being said, it can be seen that whenever there is an advantage there is a disadvantage that comes along. Just as it mentioned before, just because the source is affordable, the plant will require a lot expensive to hold the plant and to have the safety features in there so it does not cause hazardous to the environment and to the people. This will keep the environment and the people safe while being able to use. Furthermore, when the fissure is in process and the waste are being separated from it that can cause some hazardous because the waste contains certain toxins. When the fission is being broken down it is dangers for the environment and for the people because it will have some toxins. Overall, it can be seen how fission has its own disadvantages. Although, it has many beneficial advantages, it will have its downside. Just as it before, the energy source is afford, but the downside to that is keeping the plant is highly expensive. Even though it is really important to keep the environment safe and the people, it will still have its own of having some type of hazardous. For instance, as it mention when the fission is being broken down. That can result some hazardous.

Just as fission has its own advantages and disadvantages; fusion also has its advantages and disadvantages. The advantages are cost competitive, high energy, less pollution, and it can be sustainable. As it mentions, holding a power plant is expensive; however, the electric generation is a lot cheaper than the oil, gas, and coal plants. The fuel that is in the plant is a lot smaller than other power generation plants. Another advantage is, it causes less pollution. It mentioned in the article that it would be the best to replace the energy harnessing process with fusion. Last, but not least, fusion energy is not renewable. Therefore, if individuals know how to control fusion, there will be unlimited energy (Connectusfundadmin, 2015). Without
a doubt, it can be seen how fusion has its benefits. It can be seen that fusion is relatively cheap to have so in reality it will not be a problem to hold this in the power plant. Therefore, it is even cheaper than regular common materials as mentioned before, oil, gas, and coal plants. Furthermore, the fuel is not needed as much to produce high energy density and this does mean a lot. The reason why is because if a lot of fuel is needed, then the expensive will be high. Clearly, that would be a problem later more for the budget. Not causing too much pollution does mean a lot in this world. The reason why is because it will cause less health issues for the people. Therefore, if fusion is used a lot, and if it did produce a lot of pollution, it would be predicated that the people would be in danger. Lastly, fusion being sustainable is really beneficial. Knowing how to control the fusion can be really beneficial in the end. The reason why is because that can lead to having unlimited energy. Therefore, having limited energy can be an issue because it would certainly lead some problems in the environment. Hence, a lot of benefits do tie with fusion. Therefore, knowing how to control fusion can help in the long run. For instance, it would help with the pollution by not harming the environment and the people. Therefore, controlling it and knowing how to use it will help by not putting the fusion in a risk to have it limited. Overall, it can be seen that fusion does have its advantages that it helps the environment and the people around. For instance, it does not cause too much pollution, which will help the people in the environment to not breathe in that air to not cause health issue later on. Furthermore, the fact the fusion produces high energy and the fuel that is used is relatively small. Therefore, that shows too much expense will not be given here because if little amount of fuel is used, but large amount of energy is given, it is really beneficial.

There are also disadvantages that go along with fusion. They are producing radioactive waste, causing accidents, extraordinary measures, and the energy returns are unreachable. The power plant emits only certain amount of CO2 into space, but the chain process will produce the radioactive waste. Causing an accident is a big one. The radioactive waste that comes from fusion does have a risk of hazardous to the people and to the environment. The large amount of radiation can cause a lot of health issues to the people, and in the environment. In nature, gravitational forces and high temperature will create the fusion on its own; however, on Earth, it can be difficult to make the fuel hot and the fusion. Last, but not least, to have an artificial fusion, it requires a lot of energy to go in (Connectusfundadmin, 2015). Without a doubt it can be seen how fusion has its own disadvantages. Having radioactive waste is ultimately bad for the environment and for the people. The reason why is because the waste will pollute in the air and the people will breath in the air and later develop health issues. Causing an accident is a big one. The reason why is because the people are in a risk. The individual’s can develop many different type of health issues when they have too much of the harmful radiation of fusion. Therefore, even in the environment it can cause issues. It can case pollution and that will also cause people to be sick. Having extraordinary measure is a issue. The reason why is because it will use up more materials and work. Therefore, not enough energy will be produced in the end. Lastly, too much energy is needed to put in the artificial fusion. Therefore, that means a whole lot more is needed for the regular fusion. This is really a downside for that because too much energy is needed to put in. Hence, it can be seen how fusion can have its own downside. It goes all the way from radioactive waste to causing accident. It needs to be understood that there really cannot be a perfect energy. Therefore, fusion also has its own advantages and disadvantages. Although, it does balance each other out because fusion is still good to use. For
instance, the fusion has to be taken care of because it can cause accidents to the people. Of course, the accidents are known as radioactive. This is big because if people are exposed to too much of this radioactive then it will cause people to develop a lot of health issues later on in their life.

Fusion has a certain way on how it works. Two isotopes will be needed and they are deuterium and tritium. When these two isotopes are combined, they need to be heated and the heat needs to exceed 100 million degree Celsius. Once this occurs, the gas will turn into plasma. When this happens, the electrons, along with neutrons, will be highly energized. With this happening, atoms will be able to fuse with other atoms. Therefore, when deuterium and tritium come together they will be able to form helium and release energy. For the plasma to be heated, different types of actions can be taken. They are known as, lasers, electric currents, and microwaves (Living on Earth / World Media Foundation / Public Radio International, n.d). Without a doubt, it can that it does take a lot of steps on how fusion works. Therefore, two atoms need to come together to make one big atom just as deuterium and tritium came together to make helium. Clearly, heat is needed in this process to be able to produce the given off heat. Overall, it can be seen what steps are taken for the fusion to take place. It is extremely important for the heat to be this high to have the plasma formed. Therefore, this will cause the atoms to be energized and then the energy can be released.

Fusion also has its own particular way on how it works. Within a round figure reaction chamber, there will be a hydrogen gas, which consists deuterium and tritium. Therefore, the degree needs to be 100 million Celsius so the fusion can take place. No additional energy is needed to have the process completed (How Does a Nuclear Fusion Power Plant Work, n.d). Without a doubt, it can be seen that fusion does have less steps than fission. Therefore, the process is clearly easily done. This process is actually better than fission. The reason why this is begin said is because it does not need addition energy to have the process completed. It can be understood that only certain amount of energy is needed to have the fusion successful. Overall, it can be seen that same degree is needed for fission and fusion. Therefore, this is needed to have it made successfully without causing any damages.

A lot of people deal with some type of cancer. The treatment that individual’s usually get is, chemotherapy. It was said that a group of doctors were allowed to use boron neutron capture therapy on patients that are dealing with cancer. It was told that this type of treatment would be less stressful and harmful to the body compared to chemotherapy that is commonly used in todays work for cancer patients. This treatment would use fission instead of chemicals. Therefore, this would target and kill cancer cells. The main idea of using fission is to fight cancer and to cause less harm to the body. However, there is a downside to this particular method that is being used on cancer patient. It was said that this can also damage healthily cells, but it would still kill cancer cells. Although, the patient may feel a bit sick (O’Conner, A, 1999). Clearly, it can be seen that in the medical field fission is taking place. Therefore, it is being helpful. The reason why is because the traditional chemotherapy will be taken off sooner or later because this therapy, with fission, will have more effect on the cancer patient. Although, it does have its own downside to it, it needs to be understood that all therapy will some advantages will have disadvantage. Hence, it can be seen that fission can be helpful in the medial field because it is helping patients with cancer. It will kill the cancer cells. It will be more effective because it will
not cause as much damage to the body and it will not be as stressful as the chemotherapy is the patient’s body.

Conclusion

All in all, it can be seen how fission and fusion are different from one another. Fission breaks up a large atom into two or even more atoms and fusion will combine two or atoms to one large atom. With this being said, Uranium the main fuel that is used in the power plants. Therefore, for fusion, deuterium and tritium is the fuels used in the fusion power plants (Fission and Fusion, 2017). With this being said, it can be a good thing that both have different aspects when they are being looked at. The reason why this is being said is because they will be able to provide somewhat help with the energy, when one is lacking. With this being said, fission and fusion do work differently. For instance, fission occurs in the nature and fusion will be occurring in the stars and sun. Fission does require a little bit of energy when it is breaking up the atoms apart. Therefore, that contrasts with fusion because fusion requires a lot of energy when the atoms are fusing together. Furthermore, it is great that doctors are using fission to kill cancer cells. The reason why this is being said it because it will help the patient’s fight the cancer and it will cause less damage to the patient’s body. Although, it will damage some of the healthy cells, but in the long run it will help the patient because most of the cancer cells will be killed.

What I believe the future holds with fission and fusion is a big change. The first one, that is most important, is, as it mentions in the article, being able to control the release of energy from the fusion reaction. When the energy from this reaction is released slowly, it can produce electricity (Nuclear Fusion: The Hope for Our Energy Futur, n.d). I look forward to this because it will be able to provide unlimited amount of energy and there will no issues revolving this.
This equation shows the fusion equation. 
Hf means heat of fusion. 
Q means heat. 
M means mass.


This equation shows the fission equation. Element U stands for Uranium. Element Ba stands for Barium, and element KR stands for krypton.


This first image shows fission. The large atom is being split into two atoms. The second imagine shows two atoms are being conjoined into one large atom.

References


The Physics of Nerve Impulses

Joel Doolin
4/18/2017

PHY 112: General Physics II
Professor Michael Swingler
Abstract:

The process of nerve impulse, also known as an action potential, generation is discussed at length. This process primarily consists of the movement of Na\(^{+}\) ions across a previously polarized cell membrane, which sets off a chain reaction along the membrane, followed by the repolarization of the membrane through the removal of K\(^{-}\) ions from the cell. Also discussed are the similarities to traditional circuitry and the on/off concept of information transmission utilized in computers. A brief explanation of neurotransmitters and their function is included, since any discussion of the nervous system would be incomplete without covering their vital role. The paper concludes with the realization that humanity’s existence, and especially its consciousness, results from the simple movement of ions across a thin membrane.

Full Text:

Introduction

The nerve impulse is what separates complex organisms from their single celled ancestors. This simple signal grants these organisms the ability to communicate between cells and thus coordinate complex networks for the benefit of the whole. These networks can be truly immense. For example, the human brain alone is estimated to contain over 120 billion neurons\(^2\). In order for this massive number of cells to coordinate with one another, the body has developed a rather simple mechanism for the transmission of messages. Stated simply, a nerve impulse is a cascade of chemical reactions that produce an electrical signal along a specialized cell membrane. The cells containing these special membranes are known as neurons, and they are the building blocks of the body’s nervous system.

A neuron is special because it constantly polarizes its cell membrane. Proteins embedded within the membrane are designed to pump positively charged ions, primarily Na\(^{+}\), out of the cell, which creates a large enough discrepancy that a voltage develops across the membrane. Essentially, the proteins do work to remove the ions from the inside of the cell and this represents a buildup of electrical potential energy(V). This voltage is maintained at -70 mV, but when a change in the environment, a touch or change of heat, for example, occurs, another set of proteins within the membrane open channels that are designed to allow Na\(^{+}\) to flow back into the cell\(^5\). This sudden influx of Na\(^{+}\) causes the membrane to depolarize, even to the point of overcorrecting and shifting all the way up to +30 mV\(^5\). When this threshold is hit, a third set of proteins begin vigorously pumping another ion, K\(^{+}\), out of the cell in order to correct this imbalance. After a brief refractory period, the cell is able to correct the ion levels and restore the membrane back to its resting voltage of -70 mV. This cycle is then repeated all the way down the length of the neuron’s cell membrane. This wave of depolarization is what is known as an action potential or nerve impulse.

When the impulse reaches the end of one neuron it triggers proteins, based on the frequency of the waves, to release a certain type of chemical, known as a neurotransmitter, whose presence causes the same wave to occur in the next neuron. In this way, the signal is passed from one neuron to the next until it reaches its destination. These impulses physically act similarly to other electrical pulses, such as electrons flowing through a metal wire. This allows us to apply many of the same physics equations to this phenomenon that we use for typical electrical phenomenon.
The human nervous system is the communication center of the body. It allows complex systems to be maintained and updated within a fraction of a second, communicating back and forth from muscles to the brain quickly enough to react to the slightest change in our environment. It is also responsible for every thought that we have, our deepest understanding of the world around us. Without a nervous system of some kind, life cannot exist above the simple single cellular. It is only through the incorporation of neurons that complex life exists upon this planet. The nerve impulse, also known as an action potential, is the signal these vital cells produce in order to communicate with one another. Similar to a radio transmission, these impulses can tell the brain what is happening in the environment, what the body’s internal conditions are at that moment, and make decisions based on the information provided in order to sustain life. Quite simply, complex life could not exist without the nerve impulse.

Any discussion of this topic would be incomplete without mentioning the work done by Alan Lloyd Hodgkin and Andrew Fielding Huxley in the early 1950s. They published the definitive paper “A Quantitative Description Of Membrane Current And Its Application To Conduction And Excitation In Nerve” in 1952 and it has formed the foundation of work on the topic ever since. In their paper, they established that nerve impulses result from the fluctuation of voltage along the cell membrane of neurons, and that the electrical impulses act analogously to a typical electrical current (see figure 1) across the membrane. However, they go on to demonstrate that it is not the passing of electrons, but rather the movement of ions which cause the fluctuation in voltage. They measured the electrical conductance of Potassium(K+) in relation to the voltage across the membrane and found that the level of K+ rose when the voltage increased and conversely decreased along with the voltage (see figure 2). Similar experiments were conducted with Sodium(Na+) and a correlated relationship was observed. In these early experiments, the methods available for measuring current across such a small distance were relatively limited, and the measurements have been significantly refined since then, being now confirmed to be a shift from -70 mV to +30 mV.

Within a typical circuit, tracking the movement of positive charges is the convention (thanks Benjamin Franklin), even though it is the movement of electrons which is causing the “movement” of the positive charges. During a nerve impulse, however, it is indeed the movement of positive ions that cause the electrical signal. As Hodgkins and Huxley confirmed, Sodium(Na+) and Potassium(K+) are the two ions passing back and forth across the cell membrane of the neuron. Each of these ions contain only a single positive charge and are therefore largely analogous to one another electrically. The primary difference between the ions lies in the size of the atom, which impacts their chemical properties. These differences allow proteins in the membrane to differentiate between them, being conformationally arranged to only interact with one or the other at given sites within the protein. The specifics of these proteins will not be discussed, as they have no bearing on the physics of the impulse. The element of importance to the physics lies in the fact that protein pumps within the membrane can exchange 4 Na+ ions for 3 K+ ions, which increases the K+ concentration within the cell and decreases the Na+ concentration. This offset of 1 positive charge creates the potential energy difference, forming a voltage along the membrane.

Once this voltage difference is established, the neuron is considered to be in the resting state or at resting potential. The pumps along the membrane will continue to offset any ion.
leakage caused by the concentration gradient until a stimulus causes the initiation of an action potential. Neurons utilize many different sensor proteins to monitor the environment. Some of these proteins detect changes in heat or electrical energy, some detect the presence of certain chemicals, while others react to electromechanical stimuli. Once these proteins detect a given stimulus, they change their conformation through a series of chemical interactions, which in turn creates a cascade of other responses from surrounding proteins. Eventually, these reactions reach a protein that forms an ion channel in the membrane. The protein opens once it has been activated, allowing $Na^+$ to flow into the cell while the stimulus is present. The entrance of ions causes a depolarization along the membrane, which will become an action potential if it reaches a certain level. This level of depolarization is known, logically, as the Action Potential Threshold and occurs at approximately -55 mV. This threshold exists because it is at -55 mV that the voltage-dependent ion channels open within the cell membrane. Any depolarization of less than 15 mV is not strong enough to warrant a full action potential, and the cell will automatically return the membrane to its resting point without generating a nerve impulse.

The voltage-dependent ion channels are the proteins that carry out an action potential. As one area of the membrane reaches -55 mV, one channel opens, causing a flood of $Na^+$ ions. This increase in the number of ions creates a depolarization not only for that specific area, but their quick diffusion means that the entire area around that channel is now depolarized, causing a chain reaction in the nearest voltage-dependent ion channels, as their region of the membrane now hits the -55 mV level. As this wave propagates outward from the source, the electrical impulse is recreated by each subsequent ion channel, refreshing the strength of the signal all along the neuron. This allows the signal to carry as far as is necessary without the signal degrading, unlike traditional electrical wiring, where the impulse strength weakens with distance due to leakage.

Once the voltage-dependent ion channels have been opened, ions are allowed to flow into the cell until the voltage across the membrane reaches +30 mV. As soon as this threshold is reached, the $Na^+$ channels close and another set of voltage-dependent ion channels open allowing $K^+$ ions to exit the cell, flowing down the chemical gradient previously created by the $Na^+ / K^+$ pumps. The outward flow of $K^+$ ions, along with the continuing work of the $Na^+ / K^+$ pumps, quickly act to restore the polarization of the cell membrane. However, once again there is an overcorrection that occurs and the $K^+$ ion channels remain open until the membrane reaches approximately -75 mV, over polarizing (also known as hyperpolarizing) the membrane. Through ion leakage and restriction of the $Na^+ / K^+$ pumps, the membrane quickly corrects to the resting potential of -70 mV. This period of repolarization is known as the refractory period because the neuron is unable to launch another action potential until the membrane has been restored to its resting potential. The refractory period helps to ensure that nerve impulses only travel in one direction, as previously activated sections of the membrane will be unable to respond to the ion wave while it is still recovering. Figure 3 shows a plot of the membrane potential versus time, illustrating all the components of an action potential, including the refractory period.

As figure 3 illustrates, a typical action potential will only last approximately 6 milliseconds, but the speed with which that impulse travels along the neuron can vary greatly. The exact speed of transmission varies with the nerve’s function and which source you choose to use. Reports vary from 0.62 m/s to 119 m/s. Nerves that report pain, for example, find themselves on the slower side of the spectrum, with the general estimate putting them at around 0.62 m/s, which is as slow as any nerve in the body. Nerves responsible for reacting to intense
stimuli, such as a hot stove, however, can transmit at speeds in excess of 100 m/s. The incredible discrepancy between these rates has to do with an important biological substance known as Myelin.

In essence, Myelin simply acts as an insulator for the nerve axon (the portion of the cell that carries the action potential), much like a rubber coating insulates most copper wiring. Myelin is composed of molecules that resist the diffusion of ions, and because the Myelin sheath surrounding many neurons is much thicker than a typical cell membrane\(^1\), ion leakage is tremendously reduced along the insulated sections of axon. However, this insulation also means that the ion channels are not able to refresh the strength of the action potential wherever the sheath is present. This lack of renewal means that the Myelin sheath will only remain effective over relatively short distances. Evolution has provided a solution to this problem by creating gaps between the Myelin sheaths known as Nodes of Ranvier. These are small sections of exposed axon where the ion channels are able to re-initialize the action potential, sending the signal shooting through the next section of Myelin coated axon. Because the action potential is not being constantly renewed, the ions are simply diffused along the Myelin covered portion, which occurs more rapidly than the standard wave of ion channels. This increase in speed is what accounts for the increase of transmission speed from 0.62 m/s to 119 m/s across different kinds of neurons. While 119 m/s may seem to be an incredible rate of transfer (it certainly is fast enough to transverse the human body quickly), it is really quite sluggish when compared to the speed of electricity moving through copper wire, which is approximately 2.80 x 10\(^8\) m/s! That means that a fast nerve impulse traveling at 100 m/s takes about 1/50 of a sec to go from the brain of a 2 meter tall human to their feet, but an electron moving through a copper wire would only take 1/1.4 x 10\(^8\) seconds to cover this same distance. That’s virtually instantaneous in comparison.

So why does the body use neurons with ion channels and not electrons with a copper conduit? The reason that metal wiring would not be appropriate to use within the human body is the need for consistency. An action potential within the body maintains the -70 - +30 mV wave, allowing proteins to be produced that are tailored specifically to this variation in voltage. At every position along the membrane, the precise activation and deactivation of ion channels allows the signal to be consistently refreshed. This creates a constant “on/off” signal, much like a computer’s “1/0.” While the strength of the signal never changes, the frequency of these waves is what carries the message. This phenomenon is remarkably similar to the way radio stations can send out complex songs and speech patterns, despite the fact that the receiver is only tuned to “listen” to one specific spectrum. The flow of electrons through a metal conductor breaks down the further it travels due to resistance within the substance and leakage of the charges into the environment, which means that the signal at the other end may vary if the transmission were covering too great a distance. The use of ion channels ensures that the message remains the same no matter how far it may need to travel.

Distance travelled may not seem like a big deal in a human being, but in a large mammal like a blue whale, it can be incredibly important. It is much easier to understand why there is a need for precision in the brain. Use of electron-generated signals would quickly become confused within biomatter, as it would easily diffuse throughout our bodies, seeing as how the human body is almost entirely made of water. Therefore, precise communication between two neurons amongst billions would become nearly impossible. The use of ions, which are easily controlled by the cell membranes, provides a more manageable scheme for biological systems.
Perhaps if life had developed with Silicon as its base, metal wiring would have been more appropriate, but it is not an efficient solution for Carbon based organisms.

Any discussion of nerve impulses would be incomplete without at least mentioning synapses. There are 2 ways that neurons communicate between one another. Direct electrical communication, meaning the passing of the action potential, like we have already discussed, occurs between dendrites and neurons, usually when a group of nerves need to fire simultaneously. A good example of this need is within the heart. Such an organ would not work as well if all the muscles within the heart fired at random times instead of all at once. Indeed, the system would simply fail to function. As a result, the neurons that control muscle contraction within each chamber of the heart are tied together to electrically communicate with one another, allowing the contraction of muscles to be well coordinated. However, most nerves only want individual communication, as one would truly be in trouble if all the muscles within the body fired simultaneously (although something similar does occur during a seizure). As a result, most neurons within the body interface with one another through synapses. A synapse is a junction gap between 2 neurons across which communicative chemicals, known as neurotransmitters, may be passed.

Once an action potential reaches the end of the pre-synaptic neuron it activates proteins that cause bundles of specific chemicals to be fused with the cell membrane, releasing the chemicals within, which then diffuse across the synapse into receptors on the post-synaptic neuron. Once these chemicals bind to the receptor sites a new action potential is typically generated which matches the one experienced by the pre-synaptic neuron. Most psychoactive drugs interact with this process in one form or another, blocking or falsely activating these chemical signals in order to produce the resultant effect. Specifics of these neurotransmitters and the actions of psychoactive drugs is outside the purview of this paper, as these elements are far more chemically and biologically relevant and have little impact on the physics of the impulses themselves.

As previously stated, an action potential acts in a manner congruent with a typical electrical circuit, and therefore it is possible to make some general statements related to this system. Figure 1, for example, gives a good illustration of a possible circuit diagram one might write up to describe the behavior of these systems. If the cell membrane is 4 nm thick, as many sources suggest, and the resting potential of the membrane is -70 mV, the electric field (E) can be calculated for the membrane at rest as follows:

\[ V = E \cdot d \Rightarrow E = V/d \Rightarrow E = \frac{-7.0 \times 10^{-5} V}{4 \times 10^{-9} m} = -1.75 \times 10^{4} \text{ N/C} \]

However, at the peak of an action potential, the electric field becomes oriented in the opposite direction and the strength becomes:

\[ E = V/d \Rightarrow E = \frac{3.0 \times 10^{-5} V}{4 \times 10^{-9} m} = 7.5 \times 10^{3} \text{ N/C} \]

This means a single positive charge near these fields will experience electrical forces of:

\[ F_{e} = E \cdot q \Rightarrow F_{e} = (-1.75 \times 10^{4} \text{ N/C})(1.602 \times 10^{-19} \text{ C}) = -2.80 \times 10^{-15} \text{ N (at rest)} \]

\[ F_{e} = E \cdot q \Rightarrow F_{e} = (7.5 \times 10^{3} \text{ N/C})(1.602 \times 10^{-19} \text{ C}) = 1.20 \times 10^{-15} \text{ N (during impulse)} \]

This force would result in the following acceleration of a sodium ion if there were no other forces involved:
\[ m_{Na} = (22.99 \text{ kg}/6.022 \times 10^{23}) - 9.11 \times 10^{-31} \text{ kg} = 3.818 \times 10^{-23} \text{ kg} \]

\[ F = m \times a \rightarrow \vec{a}_e = F/e/m = (-2.80 \times 10^{-15} \text{ N})/(3.818 \times 10^{-23} \text{ kg}) = 7.33 \times 10^{7} \text{ m/s}^2 \]

\[ F = m \times a \rightarrow \vec{a}_e = F/e/m = (1.20 \times 10^{-15} \text{ N})/(3.818 \times 10^{-23} \text{ kg}) = 3.14 \times 10^{7} \text{ m/s}^2 \]

Naturally, the electric field is not the only force acting upon the ions, as chemical gradients, hydrostatic pressure, and gravity are all at work as well. However, it does illustrate the importance that the membrane potential plays in the movement of these ions. These calculations also demonstrate why touch screens in electronics were able to be developed to detect the electric field given off by the human body.

Many interesting mathematical relationships have been established with regard to the creation of action potentials, and although the math is beyond the scope of this paper, here are some of those relationships:

An action potential generation can be represented by the following equation\(^1\):

\[ \frac{\partial^2 V}{\partial x'^2} - \frac{\partial V}{\partial t'} - V' = 0 \]

This equation can then be represented as a Fourier integral\(^1\):

\[ V'(x', t') = \int_{-\infty}^{\infty} U(v') \exp\{\alpha(v') x' + i2\pi v' t'\} dv' \]

Also, the relationship between the proportion of ions inside the cell\((P_i)\) versus outside the cell\((P_0)\) have been found to be\(^3\):

\[ \frac{P_i}{P_0} = \exp[(w + zeE)/kT] \]

Where \(w\) represents the work done by the cell to move the molecule from inside to the outside, \(E\) represents the potential difference between in and out, \(e\) is the charge, \(z\) is the molecule’s valency, \(k\) is Boltzmann’s constant, and \(T\) is the absolute temperature in Kelvin. This relationship was one of the ways that Hodgkin and Huxley was able to identify Na and K as the 2 elements involved in the process.
Conclusion

The nervous system is an incredibly unique phenomenon. Biology using both chemical and electrical properties in order to communicate and coordinate vastly complex processes in the effort to maintain life. This system is the reason we humans are self-aware, why a blue whale can be 24 meters long and still manipulate its tail with precision, and how birds can react to wind currents within a fraction of a second, allowing them to maintain the delicate balance of flight even in most turbulent of conditions. The ability of the nervous system to provide constant information, both conscious and unconscious, process that information, render a decision, and then act upon that information is truly exceptional. We owe our existence, our presence of mind, to this vastly complex and yet deceptively simple functioning of neuronal cell membranes.

We have seen that these membranes begin by establishing a chemical and electrical gradient through the use of the Na⁺/K⁺ pumps, building a potential energy difference of -70 mV across the membrane. This resting potential is then put into action once a signal is received, and voltage-dependent Na⁺ ion channels open as soon as those sensor proteins lower the potential by only 15 mV. These channels opening then results in the rapid depolarization of the membrane, setting off a chain reaction in other sensor proteins near the site of excitation. This wave then travels along the membrane, down the axon of the neuron, refreshing itself along the way. Whenever the pulse needs to be sped up, Myelin sheaths are utilized to minimize the amount of ion channel replenishment required, allowing for simple diffusion to occur down relatively long stretches of nerve, which is much faster than the standard wave reaction. Once this electrical signal reaches the end of one neuron, neurotransmitters are released that create a similar signal in the next neuron, passing the message on down the line until it reaches its destination, where other neurons coordinate with one another to interpret and respond to the signal.

The membrane does not remain depolarized, however, but returns quite rapidly to its resting state by the removal of K⁺ by another set of voltage-dependent ion channels. This process resets the neuronal membrane so that it is once again ready to fire off a signal wave, an action potential. This reset period allows the neuron to limit the direction of the signal and standardizes the frequency of production, so that proteins can be developed to respond only to these specifications. The standardization of these pulses is what act as the “1/0” in a computer system, carrying the information being communicated by the cell in a binary fashion that is easily unpackaged at its destination.

As we have seen, the physics of nerve impulses relies heavily upon electrical properties, as the cell membrane can be represented similarly to a traditional electrical circuit, containing capacitors, wiring (the ion channels), resistors, and chemical batteries. All of these components represent subjects of study within physics, not biology or chemistry. And yet, here they are within the human body. In a way, we are electrical entities, even if it is not the movement of electrons that is being tracked but rather the movement of simple ions. Hopefully this fact can be exploited in the future to allow the human race to better interface with computers, enhancing our capabilities as a race and better equipping us to deal with life’s many dilemmas. Who would have thought that our entire self-awareness results from the passing of ions back and forth across a 4 nanometer thin sheet?
Figures:

**Figure 1**

Fig. 1. Electrical circuit representing membrane. $R_{Na} = 1/g_{Na};$ $R_{K} = 1/g_{K};$ $R_{l} = 1/g_{l}.$ $R_{Na}$ and $R_{K}$ vary with time and membrane potential; the other components are constant.

Hodgins, Huxley. 1952. 1

**Figure 2**

Fig. 2. A, rise of potassium conductance associated with depolarization of 25 mV; B, fall of potassium conductance associated with repolarization to the resting potential. Circles: experimental points replotted from Hodgkin & Huxley (1952b, Fig. 13). The last point of A is the same as the first point in B. Axon 18, 21° C in choline sea water. The smooth curve is drawn according to eqn. (11) with the following parameters:

<table>
<thead>
<tr>
<th></th>
<th>Curve A</th>
<th>Curve B</th>
</tr>
</thead>
<tbody>
<tr>
<td>$g_{Ko}$</td>
<td>0.09 m.mho/cm²</td>
<td>7.06 m.mho/cm²</td>
</tr>
<tr>
<td>$g_{Kc}$</td>
<td>7.06 m.mho/cm²</td>
<td>0.09 m.mho/cm²</td>
</tr>
<tr>
<td>$\tau_{n}$</td>
<td>0.75 msec</td>
<td>1.1 msec</td>
</tr>
</tbody>
</table>

Hodgins, Huxley. 1952. 2
Figure 3

Action Potential in a Neuron

- Membrane Potential (mV)
  - 50
  - 0
  - -50
  - -100

- Time (milliseconds)
  - 0
  - 1
  - 2
  - 3
  - 4
  - 5
  - 6
  - 7

- Phases:
  - depolarization
  - threshold potential
  - resting potential
  - action potential
  - refractory period
  - repolarization
  - hyperpolarization

http://www.dummies.com/education/science
References:


Bioelectrics and Living Organisms
Aubrey Downey
Dr. Durandet
Physics 112
November 17, 2016
Abstract:
This research paper covers the subject of bioelectrics in living organisms. It particularly focuses on specialized electric organs that generate electric fields outside of an organism’s body and are only found in five orders of fish. Bioelectrics is a relatively new scientific study, that involves understanding how the physiological mechanisms within living organisms contribute to the electric currents and potentials created by them. This paper begins with a general overview of bioelectrics within all living species. It then goes into detail about the special electric organs that can produce electricity outside of an organism’s body as well as the electric organs that can receive electricity produced from something else. The cellular detail of how these organs work is then explained to show how they are able to do this. The paper finishes with an overview of the five orders of fish that possess these organs and concludes with the future possibilities that the study of bioelectrics contains.

Introduction:
It is surprising that bioelectrics just became valid scientific study in 1973 since bioelectricity is present in every living organism (Marino 1988). The bioelectricity found in a living organism’s body is somewhat comparable to that of the electricity produced by lightning but just varies on the degree of strength (Brown 2012). For comparison, a living organisms bioelectric potential is around one to a few hundred millivolts while that produced by lightning is up to a billion volts (“The Positive” 2011). The electric signals produced within an organism do vary from electrical signals produced by light or sound in their physical properties. These electrical signals have current flow out one end and into the other, like an electrostatic field. This varies from light or sound electrical signals that are electromagnetic waves in nature. Since bioelectric signals are not electromagnetic waves, they are unable to propagate. This is important to remember when the discussion of electric organ discharges is covered further on in this paper (Stoddard 2009). The bioelectric potentials produced by living organisms are also comparable to potentials created by batteries and generators. However, the difference between electricity produced by a battery and bioelectricity lies in the current. Bioelectric current is the flow of ions in and out of cells, instead of a movement of electrons as seen in an electric current (“Bioelectricity” 2007). Both currents involve the movement of charged particles, which is what is able to generate charge, leading to current flow and electricity. Like any generator the produces electricity, there is an electric-potential difference in living cells that results in the production of voltage. In living organisms, this potential difference is created in the individual cell membranes by the separation of various concentrations of charged ions resulting in disequilibrium. Voltage-gated ion channels embedded in the cell membranes contribute to the ion separations and the generation of voltage. In a typical cell, this disequilibrium sets a steady resting membrane potential of around 50 millivolts (Zhou and Uesaka 2008). This resting membrane concentration and any potential difference found in an cell is calculated using the Nernst equation: \( V = \left( \frac{RT}{zF} \right) \times \ln\left( \frac{C_0}{C_i} \right) \), where \( R \) is the universal gas constant in joules per kelvin per mole, \( T \) is the temperature in kelvin, \( z \) is the number of electrons, \( F \) is Faraday’s constant in coulombs per mole, \( C_0 \) is the concentration of ions outside the cell, and \( C_i \) is the concentration of ions inside the cell (“The Electric Eel” 2001).

All cells of an organism’s body utilize these bioelectric potentials to communicate signals to one another, which in turn, keeps the organism alive and functioning. Some of the examples of cells that use bioelectric potentials are: the cells responsible for metabolic processes, nerve cells, muscle cells, receptor cells of sound, light, touch, and hormone cells. The action potentials
produced by nerve cells in an organism are electric pulses carried along nerves that communicate to the rest of the body to signal various processes in the body ("Bioelectricity" 2007). Without these electrical signals, the brain would be unable to communicate to the rest of the organism’s body.

What this paper focuses on is various organisms that inhabit the earth that have specially devised electric discharge organs that take the above concept of an electric-potential difference that all cells use, and concentrate it to the degree that these organs are able to produce electricity on a much larger scale outside of the body. The only class known in the animal kingdom to possess these specialized electric organs is fish (Lissmann 1958). The reason the electric organs have only evolved in water dwelling species is likely because of the physics concept of thermal conductivity. Water’s thermal conductivity is much better for electricity than air. Water’s thermal conductivity at 25 degrees Celsius is .58 watts per meter per kelvin, about 24 times greater than atmospheric air at 25 degrees Celsius, which is .024 watts per meter per kelvin (Serway & Vuille 2015). These electric organs seemed to have evolved six times independently in these unrelated fish families, making them a particularly interesting and puzzling topic to study. They also have developed from varied muscle groups throughout the fish body so they are found in various locations within the electric fishes bodies, adding to the perplexity of these specialized organs (Bennett 1971; Hill et al. 2004; Lissmann 1958; Figure 1; Stoddard 2009).

Discussion:

There are two different types of electric discharge organs, weakly electric and strongly electric discharge organs, depending on the fish’s purpose for their electric organ. As the name suggests, weakly electric organs produce much smaller amounts of electrical voltages while strongly electric organs can generate massive amounts of electricity. There are also two different types of electric organ discharge patterns: wave-like and pulse-like. Again, this is related to the purpose of the electric organ for the fish. The fish with pulse-like discharges emit pulses with intricate phasic structures at unequal gaps and alter the recurrence rate of the discharges depending on the fishes’ activity level, and these pulses can be long or very short. When a fish is active, the electric discharge rates are high but while the fish is resting they tend to be produced at a much lower rate. The electric fishes with wave-like discharges produce constant discharges, regardless of activity level, at a particular personal frequency that yield a sine wave shape (Zakon et al. 2008). This is well demonstrated to the right of each fish in Figure 1 as well as in Figure 3 below. These fish use their electric organs to generate electric organ discharges to create electric fields outside of their bodies that can be used for self-defense, hunting, electrocommunication, and/or electrolocation purposes (Bennett 1971, Hill et al. 2004, Kramer 1996). Electric fishes generate electricity through their electric organs in a similar way as other animals generate electricity within their cells, but they are able to produce such great potential differences on the surface of their body that there is a much larger voltage of electricity produced (Zhou and Uesaka 2006). Weakly electric organs are used for electrolocation and electrocommunication. These organs constantly discharge weak electric currents. Anything that is in the surrounding possessing a different electrical property than that of water will change the patterns of the currents created by the electric organ discharges (Zakon et al. 2008).

A majority of fish species with electric organs contain electoreceptor organs on their body surface that are able to detect the change in frequencies and electric fields in the water that occur close to them. As mentioned previously, these electric fields do no propagate which is why these electric fishes can only detect changes that are near them. These receptor organs are able to
create an electric visual of the surrounding environment from receiving successive electric organ discharges (Zhou and Uesaka 2006; Von 1999; Stoddard 2009). The electroreceptor organs possess specialized electro-receptive sensory hair cells embedded in the skin of the fish that act like voltmeters because they detect changes in voltages in frequencies (Stoddard 2009). There are two types of these electro-receptive sensory cells found in electric organs: ampullary electroreceptors and tuberous electroreceptors. The ampullary electroreceptor cells are common in almost every fish species with electric organs. They are very sensitive to weak electric fields, on the scale of 100 microvolts per centimeter, and respond to low frequency direct currents, on the scale of 1-100 hertz. These are the types of currents and fields that are created by ventilation or muscle actions of small animals and because of this, the ampullary receptors tend to be used for electrolocation of prey (Stoddard 2009). On the contrary, tuberous electroreceptor cells are only found in a few species. These cells are sensitive to strong electric fields at high frequencies, from 80 to 5000 hertz, and lack at low frequency detection (Kramer 1996). These receptors are able to recognize species-specific frequencies (Stoddard 2009).

The electroreceptor electric organs are able to analyze the electrical components of an object or animal that interrupts their discharge current. The receptor organs measure the nearby changes in amplitudes of the currents, which informs the fish of the impedance of that object or animal, and electric fishes can analyze the capacitance and resistance of complicated impedences to determine if the object is living or not (Von 1999). This is because if an object is more resistive than the water, it will create an “electric shadow” on the electroreceptors in the skin. If an object is more conductive than the water, it produces an “electric hotspot” on the surface of the fish, and it is mostly living things that carry the property of capacitance (Stoddard 2009). Through electroreceptor organs, electric fish are able to determine the material, size, shape, and distance of an object that interrupts their electric field. If another species is in the surrounding area, the electric discharges can help the fish identify what the other species is and the sex of that species, as well as allow same species to communicate (Hill et al. 2004). This electrocommunication happens when one fish generates an electrical discharge and another fish is able to receive this through their electroreceptor organs. The electric organ discharge properties, like the amplitude, waveform, and rate, can be adjusted to produce varying electrical signals leading to communication. Electric fishes are able to recognize the sex since male and female electric fish have different discharge rates and electric organ discharge waveforms. This sex waveform variation is under the control of hormones, like any other organism. Species recognition occurs since each species has specific electroreceptors that recognize species-particular electric organ discharges and their specific waveforms (Castello et al. 2009; Stoddard 2009).

Electric organs are able to produce electricity through the use of specialized muscle cells called electrocytes (Zhou and Uesaka 2006). Electrocytes develop from muscle cells called myocytes. During development, these myocytes come together and stop expressing the typical muscle cell characteristics, like sarcomeres and proteins responsible for contraction. From here, they begin to develop more like a neuron cell would as they innervate their membranes with numerous voltage-gated ion channels and transporters that allow for a large amount of ion flux. Mature myocytes, electrocytes, can be as large as a millimeter across and are able to be viewed by the naked eye. This makes sense since they have to be able to hand the large electric and voltage load that they produce (Stoddard 2009). Electrocytes are still comparable to normal skeletal muscles in that they are found at the end of nerve cells, have presynaptic cholergeric motor neurons that innervate them, they are receptive to acetylcholine, and hydrolyze.
acetylcholine with acetylcholinesterase (Hill et al. 2004). They create a concentration gradient with the use of sodium-potassium pumps and cellular unit of energy, ATP, which creates a large potential difference and generates action potentials like a muscle cell would. The number of electrocytes found in an electric organ varies on the species and the purpose for the electric organ. There can be up to 200,000 electrocytes arranged in a single organ. The electrocytes are similar to cells in a battery in that they are shaped like discs and can be aligned the same way as cells in a battery would be (“The electric eel” 2001). Thousands of them can be stacked asymmetrically into columns with each organ containing up to 1000 columns (Hill et al. 2004). The electrocytes are arranged in series to allow charge to flow from the negative end to the positive end, just like in a battery. The columns, on the contrary, are arranged in parallel to allow ions to steadily flow through them creating a constant resting potential and the ability to generate a massive potential difference (“Bioelectricity” 2007, Jie et al. 2016, Zhou and Uesaka 2006). The series circuit allows the generation of a large amount of voltage from the electric organ since voltage changes are additive across the electrocytes and the parallel arrangement of the columns allows for greater current flow (Hill et al. 2004). This is because in series circuits with direct current, which is the case in the electric organ, total voltage follows this equation: $V_{tot} = V_1 + V_2 + \cdots + V_n$, where each “n” number of electrocyte voltages are added together to get the total voltage in volts. This structure is similar to a Marx generator in that a high voltage pulse, as in the “electric shock” produced by an electric eel, is created from a low-voltage direct current supply, as in the individual voltages of the electrocytes (Zhou and Uesaka 2006).

In the medulla of electric fish, there are neurons called pacemaker neurons that trigger the initial neural signal that leads to an electric discharge (Stoddard 2009). The neural signal is sent through the large innervation of spinal electromotor neurons in the electric organ resulting in the synchronization of the activation of each individual electrocyte (Zhou and Uesaka 2006). These signals from the brain cause the release acetylcholine between the synaptic cleft of the electromotor neuron and the electrocyte. The acetylcholine binds to its receptors on the electrocyte and this causes the opening of ion channels and an influx of positive sodium and potassium ions into the electrocyte, causing massive depolarization and the discharge of electricity (“The electric eel” 2001). The impulses from these nerves are correctly timed to activate each individual electrocyte at the exact same time resulting in all cells being activated together (“Bioelectricity” 2007, Kramer). The electrocytes respond to the neural signal by changing their membrane potential producing up to 150 mV of electricity each (Hill et al. 2004). The electric organ is also tightly enclosed by connective tissue, limiting the amount of current that leaks out. This allows the summation of all the individual voltages produced by each electrocyte, which in some species can result in a massive amount of voltage and electricity (“The electric eel” 2001, Kramer 1996). The production of one electric discharge is relatively energetically cost efficient since the positive ions flow with their concentration gradients into the cell. However, reestablishing the disequilibrium after an electric organ discharge does require the use of ATP. Depending on the species and the strength of the electric discharge, this energy cost can consume anywhere from 2-25% of this fishes total energy (Stoddard 2009).

The five orders of fish that contain species that have these electric organs are: gymnotiformes, osteoglossiformes, siluriformes, perciformes, torpediniformes (Kramer 1996). As noted above, these electric organs are located in various parts in the bodies of these five orders of fish. In torpediniformes, the electric organs are located in the front of the head while in siluriformes, the electric organs encase the entire body. In gymnotiformes, the organ runs along the length of its body but in perciformes, the electric organs are in their eyes only. Finally, in
osteoglossiformes they are found in the tail fin and axial muscle (Kramer 1996). This is demonstrated nicely in Figure 1 in the figures section below. Perciformes, more commonly known as stargazers, have one of the eighteen genera called Astroscopus with only three species that possess electric organs. These fish do not have electroreceptive organs so they are unable to use their electric organs for electrocommunication or electrolocation. The main purpose for their electric organs is for prey capture and defense. Each of their electric organs is located in their eyes and there are 150-200 electrocytes per column. These organs produce up to five volts of electricity that is used to stun their prey before capture. Gymnotiformes, or South American Knife Fishes, have about 108 fish species that have electric organs. Within this order, there are the two types of electric organ discharges: wave-like discharging and pulse-like discharging. These fish are able to produce electric organ discharges at a rate of 2000 discharges per second. It is in this order of fish that the electric eel is found. The electric eel is particularly interesting since it has three pairs of electric organs with one pair strongly electric and the other two pairs weakly electric. It is the only electric fish species that is able to produce both low and high voltage discharges (Stoddard 2009). The eel’s electric organs can generate currents up to 1000 volts and the three pairs make up four-fifths of the total length of the eel (“Bioelectricity” 2007). The three organs are called the Main organ, the Sach’s organ, and the Hunter’s organ (Souza et al. 2007; “The Electric Eel” 2001). The Main and Hunters’ organs are the strongly electric organs used for hunting and protection and they produce the highest voltage. In hunting, the eel surrounds its prey in a C-shape and then discharges its high voltage discharge from these two organs. This shape concentrates the voltage and stuns the prey. The Sachs’ organ produces much less voltage and is considered the weakly electric organ type. It is mainly to assist the eel in electrolocation and electrocommunication in the water and only produces about 10 volts per centimeter (“The Electric Eel” 2001; Stoddard 2009). It is also used to locate the stunned prey that has likely floated to the surface after being subject to its high voltage discharge (Stoddard 2009). Figure 2 shows these three electric organs and their location in the eel’s body. It also shows a cross section view of the organs and how the electrocytes in the main organ are aligned (Hill et al. 2004). Siluriformes, otherwise known as catfishes, only have one member of the order with an electric organ: malapterurus electricus. This electric fish, like the stargazers, does not generate electric organ discharges for electrocommunication purposes. They do produce electric organ discharges when there are disturbances, when capturing prey, and when defending themselves. The disturbance discharges are short, producing a max rate of 67 discharges at a very low frequency. However, prey capture and defense discharges are much longer at a very high frequency and produce a rate of up to 562 discharges with 450 volts of electricity. Osteoglossiformes, or elephant fish, have two families with electric organs: Gymnarchidae and the Mormyridae that have about 200 species of electric fish. They use their weakly electric organ discharges and electroreceptor organs strictly for communication and for locomotion. These fish are highly dependent on forming schools and group cohesion so the use of species-specific electrical discharges to communicate and locate each other is key for their survival. Torpediniformes contain the well-known electric ray species. The powerful electric rays produce electric organ discharges only in defense and prey capture. Prey capture involves discharges of a rate up to 340 lasting 24 seconds that produces a voltage of 50 volts and 1 kilowatt of power. This results in immobilization, spinal chord breakage, and blackening of one or both sides of the prey’s body. If the prey is able to escape, it will still die within two days due to the shock of the electrical impulse. There are also weakly electric skates under this order of fish that use their
electric organs strictly for communication. These electric discharges in contrast only produce a voltage of 1-1.5 volts (Kramer 1996).

**Conclusion:**

As noted before, the study of bioelectrics is a relatively new scientific field of study. Electrocommunication and electrolocation is only a few decades old (Kramer). With this being said, there is a limit on the available conclusive evidence out there in regards to this topic and it leaves a lot of room for growth and discovery in this particular scientific topic. These electric fish species I have mentioned above have fascinated humans since the time of the ancient Egyptians, where they documented how these animals could shock them and stun their prey (Marino 1998). However, electric fish are just now being understood as advances in science, such as voltage clamping, allow us to measure and record more accurately the electrical discharges that they produce (Zhou and Uesaka 2006). Although there are only five orders of electric fish known to possess electric organs, there is speculation that is definitely possible that other species with these electric organs have yet to be discovered. It is also still a continual debate on how these specialized organs came to develop in the first place, and why they managed to evolve seemingly independently of each other (Zakon et al. 2008; Lissmann 1958). This gives this topic an exciting edge knowing that there is still a lot left to be discovered and understood.

Studying bioelectrics is important since all organisms and cells have bioelectric properties. Understanding deeper how bioelectrics work can allows us to see why certain mutations occur or why things go wrong in development. It is through bioelectric signals and bioelectricity within cells that all production and actions occur in an organism. Bioelectricity contributes to the creation of new cells and the rebuilding of damaged and old ones. All higher order commands that come from the brain are relayed throughout an organism through bioelectricity. The ability to control and manipulate bioelectric signals opens up a Pandora’s box of scientific and health possibilities such as disease understanding, prevention, and treatment. In electric organs, the large amount of acetylcholine that is found in them alone is a great reason to study these specialized organs. Acetylcholine is an incredibly important transmitter for human brain and muscle function. These electric organs have such a high concentration that they allow scientists to get a better focus on just acetylcholine and get a better understanding of this complex enzyme. This enzyme is one of the most abundant neurotransmitters in the brain and imbalances of acetylcholine is known to contribute to various diseases like Alzheimer’s, schizophrenia, cardiovascular disease, and osteoporosis to name a few (Hill et al. 2004; Stoddard 2009; Kramer 1996; Devitt 2014).

I believe that this topic provides many exciting possibilities for the future. If we are able to understand exactly how these electric organs came about from normal muscle cells, by narrowing down their genetic make up, we have the ability to possibly manipulate muscle cells in other organisms to create electric organs. It also gives us a better understanding of evolution and how species are adapt in such magnificent ways to survive in this changing world. Science has made major strides and advancements in bioelectrics in a little over a few decades. If advancement in this field continues at this rate, it could lead us to be able to artificially create electrocytes that could be used to power bionic devices in humans, such as artificial limbs, pacemakers or for uses that have yet to be discovered (Devitt 2014).
Figure 1: The various locations of electric organs in the five orders of electrogenic fishes. The electric organs are highlighted in black on the fish body with (A) showing strongly electric organs and (B) showing weakly electric organs. On the left side of each fish is a cross section of the electric organ and on the right side shows the various wave-like or pulse-like patterns the electric organs generate. Reprinted from *Electroreception and communication in fishes* (p. 20) by Kramer, 1996: Progress in Zoology.
Figure 2: The three electric organs of the electric eel.

(1) The location of the three electric organs in the eel’s body: the strongly electric main organ and the two weakly electric Hunter’s and Sach’s organs. (2) A cross-sectional view showing the organs and how the electrocytes are stacked in columns in the main organ. Each electrocyte is separated by insulated tissue and are 4 cm long, 1.5 mm wide, and 100 micrometers thick.

Figure 3: The various wave-like and pulse-like electric organ discharges produced from five different electric fish species. (A) A wave-like discharge from the knife fish *Sternopygus macrurus*. (B) A wave-like discharge from the brown ghost fish *Apteronotus*. (C) The pulse-like discharge from the electric eel *Electrophorus electricus*. (D) The pulse-like discharge from the knifefish *Brachyhypopomus pinnicaudatus*. (E) The pulse-like discharge from the elephant nose *Gnathonemus petersii*. 
Reprinted from *Molecular evolution of communication signals in electric fish* by Zakon et al., 2006: Journal of Experimental Biology.
References:


The Physics Behind Gymnastics

Macie DuBiel
Abstract

The purpose of this paper is to demonstrate how certain principles and laws of physics relate to gymnastics. It will give a detailed summary of gymnastics and each event competed including vault, bars, beam and floor. This paper will describe certain skills performed in each event that will be analyzed and dissected to better understand the physics behind it. The concepts of physics being examined include linear kinematics, angular kinematics, linear kinetics and angular kinetics. Both qualitative and quantitative analyses will be used to help better comprehend the physics demonstrated in each skill.

Introduction

Gymnastics is a highly competitive sport that includes four main events. Each event allows for very diverse skill sets ranging from powerful and energetic motion, to soft and balanced movement. These four events include vault, bars, beam and floor. Gymnasts must be flexible, strong, powerful and controlled to be able to pull off the skills they are performing. To be able to obtain all these qualities and improve and learn new skills, gymnasts give up many hours to train for this sport. It is an all year sport with no breaks. Higher level gymnasts usually train about four hours a day, five days a week. Training includes intense workouts to build up strength, drills for new skills being learned, time to stretch to gain flexibility, routine run-throughs, maintaining skills along with many other training techniques to improve the overall performance of gymnasts.

Gymnastics mainly includes trial and error when gymnasts are attempting new skills. Coaches will give feedback on corrections and improvements needed to be able to perfect the skill. The coaches are able to identify small and large errors by understanding the biometrics of movement and observing the gymnasts motion throughout her performance. According to William Sands, coaches identify error by dividing their though process into four areas. These areas include linear kinematics, angular kinematics, linear kinetics and angular kinetics (Sands et al. 2011). By understanding and applying these four concepts, coaches can usually pinpoint the gymnast’s exact mistakes.

Linear kinematics focuses on motion that is moving forwards or backwards in a straight line. Vault, floor and beam all demonstrate linear kinetics. This concept focuses on distance, speed, velocity, displacement and acceleration. The understanding of linear kinematics can be very helpful when observing a gymnast sprinting towards the vault. The variables can help identify how to gain explosive movements which are greatly needed by acceleration or how the speed of the gymnasts run can affect her overall performance of the vault routine.

Angular kinematics is comprised of two different types of motion which include rotational and revolutionary motion (Varley 2010). An example of rotational motion is earth spinning on its axis. This can also be applied to gymnasts twisting while flipping. An example of revolutionary motion is a tire spinning or a gymnast performing giants on a bar. Angular kinematics deals with angular distance, angular velocity and angular acceleration.

Linear kinetics is cause and effect. It looks at the transition of motion. Linear kinetics implies that all objects have inertia meaning objects have the tendency to remain unchanged unless an outside force intervenes. When looking at linear kinetics momentum, mass and force are the main variables being observed. Newton’s Laws of motion apply to linear kinetics. Newton’s Laws include: Newton’s 1st law that states an object at rest will stay at rest unless acted upon by an outside force. Newton’s 2nd law is the law of conservation meaning force is equal to mass times acceleration. Newton’s 3rd law states that forces occur in equal but opposite directions (Nave 2000). These three laws can be seen on every event in gymnastics.
Angular kinetics has similar variables as linear kinetics. The variables include moment of inertia, which can be related to mass, torque which is related to force, angular momentum which is related to momentum and Newton’s Laws of motion. Angular kinetics can be seen mainly on vault and floor. Both vault and floor require much force, torque and inertia to be able to spring off of the floor or spring board and perform any sort of skill. Many of the tricks demonstrated on these two events involve twisting which is where angular momentum comes into play.

In this paper, linear kinematics, angular kinematics, linear kinetics and angular kinetics will be broken down and observed in different skills performed on vault, bars beam and floor along with a detailed description of the skill being examined. Both qualitative and quantitative analyses will be demonstrated to help better understand the full effects of the physics behind gymnastics.

Vault

The vault is an event that is competed in gymnastics. It includes a runway, which is about 25 meters long (Dupuis 2005) followed by a spring board and horse (an apprentice which stands about four feet off the ground and is set perpendicular to the runway and parallel to the floor) along with a landing matt on the other side. Gymnasts begin this event by standing towards the beginning of the run way at a designated spot. The distance at which they start is measured on a tape measure and marked based on the stride length of their run and the number of steps taken upon impact on the spring board. Their start distance can be measured in meters and is consistent throughout each trail. Next, gymnasts can either jump onto the spring board or round-off (similar to a cartwheel, only landing with feet together) onto the spring board depending on the skill they are performing. Then, the gymnasts’ hands quickly touch and spring off the horse allowing for airtime to finish their vault with their chosen dismount (e.g., see Figure 1). Dismounts can range from skills that simply go to straight to their feet, to dismounts that flip and twist even multiple times in the air.

The skill that will be examined for principles and laws of physics on vault is the Yurchenko full. A Yurchenko full is a skill that begins with a gymnast sprinting and round-offing onto a spring board. The gymnast then does a backhand spring (starts from standing position and proceeds to jump back onto hands, finishing back on feet in an upright position) onto the horse where she springs off of her hands to then perform a full twisting layout (a backflip that stays in a straight and hollow position throughout and completes one full rotation). The Yurchenko full demonstrates a wide variety of physics concepts including linear kinematics, angular kinematics, linear kinetics and angular kinetics.

Linear kinematics can be mainly observed when a gymnast is sprinting full speed towards the spring board. Here, speed, velocity, displacement and acceleration can be applied easily. The gymnast begins her vault with an initial velocity of zero. She stands ready for her vault routine in the same spot throughout each vaulting routine. Both of these variables will remain constant. The distance between where she stands and the spring board is known as displacement. Then her initial velocity and final velocity before impacting the spring board along with the time it takes to reach the spring board can determine the acceleration. For example, the equation for acceleration is \( a = \frac{\Delta v}{\Delta t} \). It is known that the initial velocity is 0, and let us say that the final velocity of the gymnast is 5.5m/s. If it took the gymnast 4 seconds to get to the spring board then the equation would be set up as \( a = \frac{(5.5\text{m/s} - 0\text{m/s})}{(4\text{s} - 0\text{s})} \) and her acceleration would equal \( 1.4\text{m/s}^2 \). Average speed of the gymnast’s run can also be calculated by using the distance traveled, divided by the time it took to reach the spring board.
Angular kinematics is another principle that can be applied to the Yurchenko full. Angular kinematics includes rotational and revolutionary motion. As revolutionary motion is not present in this skill, rotational motion is. In the Yurchenko full, as the gymnast leaves the vault table and begins her dismount, she executes a layout full. The full twist in her dismount demonstrates a rotational force around an axis. Gymnasts are able to change their rate of rotation due to several factors including increasing the velocity of the rotation, more airtime and changing the distance of their center of mass (Lewis 2014). By applying these variables, gymnasts can control the rate of rotation and the number of rotations performed.

Linear kinetics can be observed in the Yurchenko full by looking into Newton’s Laws of motion. The first law implies that an object in motion will stay in motion unless an outside force is acting upon it. The most obvious force that can be mentioned is gravity. As the gymnast is performing her dismount, gravity is pushing down on her forcing her to return back to earth. Newton’s 2nd Law is F=ma. By this formula we can assume that the greater the mass and acceleration, the greater the force will be. The Yurchenko requires much force when springing off the spring board and vaulting table. The greater force that is applied, the more power the gymnast will have to complete her routine. The 3rd law of Newton’s Laws of motion is that there is an equal and opposite reaction. For example, when the gymnast punches onto the spring board, the amount of force she applies to the spring board will give her equal and opposite force back, allowing her to get up on top of the horse and finish her routine. Momentum is another important part of vault. The great momentum a gymnast has the more power she has to complete a skill. Momentum is based on mass times the velocity. Similar to force, the greater mass and velocity a gymnast obtains, the greater momentum she will receive.

Last, angular kinetics plays a role in the performance of a Yurchenko full. Focusing on the full twisting layout, angular momentum can be observed. Due to conservation of momentum, a gymnast cannot achieve angular momentum after her hands have left the horse. Angular momentum is achieved by pushing off the vault table at an angle (Lewis 2014). The angle in which the gymnast enters and leaves the table is very important and can make or break the routine.

Bars

The uneven bars is another event that is competed in gymnastics. It includes two bars, a lower bar which is about five and a half feet off the ground and an upper bar which is about eight feet off the ground. Gymnasts perform different skills while circling, swinging and releasing around the bars. The skill that will be focused on is called a giant. A giant begins with a powerful cast to a handstand. Then the gymnast lets her body fully swing down and around the bar ending back into her original handstand shape (e.g. see Figure 2).

Linear kinematics, as stated before, deals with objects moving in a straight forwards or backwards motion. Since a giant is performed by swinging in a circular motion around a stationary bar, linear kinematics does not apply to this skill.

Angular kinematics plays a definite role in the performance of a giant. The gymnast is holding onto a stationary bar with her hands, and swinging her body fully around the bar. In angular kinematics, this motion can be seen as a revolution. The centripetal force which is placed upon the gymnast can be found by plugging a few variables into the centripetal force equation which is Fcentripetal = m(v^2/r). The smaller the mass, the smaller the centripetal force, the smaller the velocity, the less centripetal force there is and the smaller the radius, the more centripetal force there will be (Wagon 2000). Gymnasts must apply enough centripetal force to be able to make it fully around the bar without stopping.
Linear kinetics can also be applied to a giant swing. Newton’s 1st Law can be seen as the gymnast swings around the bar, she will continue to do so until an outside force interrupts the motion. In most cases, giants usually end with a dismount so the gymnast releasing from the bar would be considered the external force leaving gravity to do the rest. Newton’s 2nd Law which is F=ma, can be used in an angular form meaning the greater mass and centripetal acceleration a gymnast has, the greater the centripetal force there will be. Newton’s 3rd Law states that force has equal and opposite affects. A gymnast needs to gain enough centripetal force on the down swing, to allow her to make it on the upswing. If not enough force is applied in the first half of the giant, then the gymnast will not be able to make a full revolution. Momentum is another variable that is necessary for a gymnast to complete a giant. By building up momentum in the giant, the gymnast can create enough potential energy and power to perform her dismount. The more momentum and kinetic energy the gymnast builds up in her swing, the more airtime she will have to complete her dismount (Littleton 2015).

In a giant, an angular kinetics concept that can be applied is torque. Torque is the amount of force that can be produce to create rotation around an axis (Lippert 2011). This is due to eccentric forces which create rotation. Torque in a giant is gained by two different positions. The first position is a hollow position (tight straight body with a slight forward tilt). The gymnast begins her giant in this shape and continues until she is almost at the bottom of her giant. She will then switch shapes to an arched position (hands and toes are slightly behind the body giving the appearance of a “C” shape) and quickly snap back to a hollow shape as she begins her upward swing. The quick shift of positions would be considered the eccentric force.

**Beam**

The balance beam is an event in gymnastics where gymnasts must perform different skills on a beam that is four inches wide and four feet off of the ground. Gymnasts must demonstrate great balance, coordination and control while performing their routines on the beam. The skill that will be looked at on this event is the punch front tuck. A front tuck on the beam starts with a step punch onto two feet before the actual flipping begins. As the gymnast punches onto the beam, her arms rise above her head reaching towards the ceiling. Then she will use the force from the punch to jump into the air and snap her arms forward over her head. This will allow the rotation of the flip to begin. Next, the gymnast will tuck her knees close to her chest and open up for the landing. As the gymnast lands the skill, she will end in a stick it position. In this position, knees and hips are slightly flexed, arms are extended out in front of the body and standing in a strong upright position. This allows for immediate distribution of force making the landing smooth (Gymnastics…2016). This position helps with balance and control.

When performing a front tuck on the beam, linear kinematics can be applied. For a gymnast to pull off this stunt, much acceleration is needed. The gymnast only has about one to two steps and a punch to gain as much acceleration as she can. Acceleration is based on the change in velocity over the change in time. The more acceleration she can gain the more height and power she will have in the air.

As for angular kinematics, there is no revolution happening in this skill, but the gymnast does demonstrate rotation. The axis point of rotation is located on the transverse plane (cutting the body into upper and lower halves) of the gymnast’s waist. Her body will then rotate around the axis before landing back onto the beam. It is important that the gymnast maintains an axis point parallel to the ground to keep her in an optimal position for her landing.

Linear kinetics is another concept that can be observed when a gymnast performs a front tuck on the beam. All three of Newton’s Laws of motion apply. Newton’s 1st Law is
demonstrated when the gymnast is flipping. She will continue flipping until an outside force disrupts the motion. In this case, gravity as well as change in positions is the external forces stopping her. Newton’s 2nd Law is observed during her take off. Force is determined by the mass and acceleration. The more acceleration she can gain from the one to two steps, the greater the force will be. This leads to Newton’s 3rd Law. The greater the force the gymnast can apply the greater the equal and opposite reaction will be. This will give her more time in the air to perform the skill making the landing much easier on the beam.

Angular kinetics can also be found in this skill. Torque and angular momentum can be identified during a punch front tuck on beam. As stated before, torque comes from an eccentric motion giving force to rotate around an axis. In this case, torque can be found the moment before the gymnast takes off. As she is leaving the beam, her arms are snapping in a forward direction giving her rotational acceleration. Angular momentum can also be seen as the gymnast demonstrates a front flip. The use of torque generates and absorbs angular momentum (Lesson 13… [date unknown]). The equation; angular momentum is equal to the moment of inertia times the angular velocity, or \( L=I\dot{\theta} \).

Floor

The last event competed in gymnastics is floor. Floor is comprised of a spring loaded floor that is 12 meters by 12 meters. The main skills performed on floor are tumbling series with some choreographed dancing and leaps in between. One of the most important and basic skill that gymnasts first learn on floor is the backhand spring. This skill can be performed from standing, or be used as a connecting skill in a tumbling pass.

A backhand spring begins with a gymnast standing in a hollow upright position. She then flexes her knees and changes the angle of her body so she begins to fall backwards. Then at just the right moment, the gymnast jumps, throwing her arms and head back towards the ground. As her hands touch the ground, she is now in an arched position. She then snaps her toes over and at the same time pushes against the floor allowing her to return to her original upright hollow position (e.g. see figure 3).

While performing a standing backhand spring, the linear kinematics of the skill can be looked at. To begin, it is known that since this is a standing skill, the initial velocity is zero. By measuring the distance from the gymnast’s initial feet placement, to her final feet placement will give the displacement. Another variable that can be determined is the duration of the back handspring from start to finish.

The angular kinematics of a standing backhand spring includes rotational movement. The center of gravity begins behind the gymnast in the center of her body. As she jumps back to her hands, the center of gravity remains on the backside and as she returns to her feet, the center of gravity moves towards the front, center of the gymnast (Jeffrey [date unknown]).

Linear kinetics of a back handspring demonstrates Newton’s second and third law as well as momentum. Newton’s 2nd Law in a backhand spring can be observed as the gymnast initially jumps back into the skill and as her hands touch the floor. Newton’s second law states that force is the product of mass times the acceleration. The greater the gymnast pushes off of the ground when she jumps, the greater the force is. This also applies as her hands touch the floor. The more the gymnast pushes into the floor with her hands, there greater the force is. This then directly relates to Newton’s 3rd Law. The more force which is applied throughout the skill, an opposite and equal reaction will occur. The product of this reaction is known as a rebound. The more force applied throughout the back handspring, the higher and more powerful the rebound will be. Momentum is another concept that can be seen in a back handspring. Momentum is
equal to the mass times the velocity. The greater the velocity and mass of the gymnast is, the more momentum she will gain resulting in more overall power of the skill.

While watching a standing backhand spring, angular kinetics can be observed. Torque and moment of inertia are both present in this skill. Torque can be found as the gymnast begins the back handspring and throws her arms and head back reaching to the floor. This creates rotational motion. The torque is what allows the gymnast to return to her feet once the rotation of the movement begins. The moment of inertia in a back handspring can be seen as two different equations. The first equation relates to Newton’s second law which is \( T=I\alpha \) and the other equation relates to momentum which is \( L=I\omega \).

Conclusion

Gymnastics is an extreme sport that requires much strength, flexibility, balance and coordination. Not only does gymnastics test the physical boundaries of the body’s limits, but the mental limits as well. Gymnasts need to be able to take corrections given by coaches and apply them to their next practice run. Coaches need to understand the biomechanics and physics of skills to help gymnasts fix mistakes in their routines. This paper relates certain skills performed on vault, bars, beam and floor, and demonstrates the physics behind it. The skills mentioned in this paper include a Yurchenko full performed on the vault table, a giant which is a skill seen on the uneven bars, a front tuck on the beam and a back handspring performed on the floor. For each of these skills, the concepts of linear kinematics, angular kinematics, linear kinetics and angular kinetics are covered. By understanding these concepts while observing different skills, coaches are able to quickly find minor and major mistakes made in routines as well as give input on how to increase power and performance.
**Figure 1** (Vault setup and an example of a vault routine called a Yurchenko)

**Figure 2** (Gymnast performing a giant on the uneven bars)
Figure 3 (Gymnast performing a backhand spring)
Cited References


Available from:

http://www.hunter.cuny.edu/physics/courses/physics110/repository/files/section51/11Kinematics
ofAngularMotionRev.pdf


Available from: http://www.regentsprep.org/regents/physics/phys06/bcentrif/default.htm
Iontophoresis Therapy

Janessa Dunbar

April 20, 2017

CHM 152 11174

Dr. Scott Massey
ABSTRACT

Transdermal drug delivery has been around for quite some time now. There have been several advances in the last few decades. New advancements in technology and research allowed iontophoresis to develop faster and increase popularity. Chemistry and biology play major roles in the success of this method. There is a list of benefits that make iontophoresis a great competitor in medicine. When compared to other ways to deliver drugs, this method stands out from the systems that are typically used. There are many uses for iontophoresis, particularly in dermatology and physical therapy.

INTRODUCTION

Iontophoresis is essentially the transfer of ions. When the word is taken apart that is what the prefix and suffix mean together. Ionto = ions and phoresis = transfer, therefore iontophoresis is ions transferring. More specifically in the medical industry, transferring positively or negatively charged ions into the skin. This process is done with the assistance of an electrical current. The charged ions are in the form of some type of medication. This is called transdermal iontophoresis. There is a list of medicines that can be used in the technique of drug delivery. The use of this method is starting to take off and it is becoming great competition for typical ways to administer medication. There is also a variety of different issues that can be treated with the use of iontophoresis.

HISTORY AND BACKGROUND INFORMATION

Although iontophoresis has been around for a few centuries, it has only been recently that this method of drug delivery has gain popularity. It began in the mid 1800’s with an Italian librarian named Giovanni Francesco Pivati and began to grow from there. His finding did not have to do with medication, or even humans at all. He found that applying an electrical current to an airtight jar of Peruvian balsam, a thick liquid from a tree in Peru with a scent of cinnamon and vanilla, allowed the scent to be apparent in that room (Khan et al. 2011). From about the time of Pivati’s findings to around 1800, there were not many attempts to see what other things could be transmitted with an electrical current. In 1800, the finding of the Voltaic pile started more efforts to see what chemicals could be transferred through membranes. Alessandro Volta found the Voltaic pile, which is a simple process of producing a continuous electrical current (Khan et al. 2001).

Contributions to iontophoresis were also made by Benjamin Ward Richardson and Hermann Munk. These two men explored the use of medicine with this method. Richardson brought this method into the dental industry, and was even given the title of “father of dental iontophoresis”. Most of Munk’s studies were focused on the transport of substances through porous membranes (Khan et al. 2011). Many years later into the 19th century, this system had some major progressions. William James Morton, Stephen Leduc, and Fritz Frankenhauser are the ones noted with this progression during this time period. During their time, metal ions and alkaloids were tested while using this method. The term “iontophoresis” came from Frankenhauser. Before that, it was known as “cataphoresis”. “Electrically assisted transdermal drug delivery” would be one of the terms used for iontophoresis in more recent times (Khan et al. 2011). Initially, it started as just the transfer of positively charged ions, then progressed to the transfer of all types of charged ions.
It wasn’t until about three decades ago that the first transdermal drug delivery system was introduced in the United States (Rawat et al. 2008). Although the technology showed promise, the development of it commercially was very slow. Newer research, technology and products has led to more advances in this area. It is still relatively new in the medical field, but recent research into all the aspects of iontophoresis may help this method of drug delivery.

**PRINCIPLES**

Chemistry and biology are both major parts of iontophoresis. The chemistry aspect plays a large role in how the drugs are delivered. Most of this system is based on how charges of ions work. Charges that are alike repel each other, whereas charges that are unlike each other would be attracted (Dhaval et al. 2010). An iontophoresis machine has an anode and a cathode that are placed on the skin. The charge of the medication involved would be used on the electrode with the same charge. For example, a positively charged medicine would be used on the anode. The ions in the medicine would push away from that charge coming off of that electrode. When the electrodes are placed on the skin, it pushes the ion into the skin, allowing that medication to be delivered.

Along with the charge of the ions, there are many other factors of chemistry that play a role in transdermal drug delivery. These include: the weight, size and polarity of the molecule, the concentration, pH and strength of the solution, and the current, duration and electrode material used for the treatment. When it comes to molecular size and weight, typically the smaller ions are transported faster than larger ions. As for polarity, more hydrophilic ions work best for this method (Dhaval et al. 2010). Drug concentrations are very important for iontophoresis to work well. It is found that an increased concentration of the medicine will lead to a greater drug delivery (Prakash et al. 2014). The influence of pH is one component that should be watched carefully. When using solutions in iontophoresis it is best to keep the pH around 7. If the pH were to be decreased, there would be an increased risk of a vascular reaction. This reaction would be caused by an increase in hydrogen ions in the solution. If needed the pH can be dropped a little, but not any lower than a pH of 5.5 (Kirubakaran et al. 2015). The strength of ions is directly linked with the permeation of drugs. It is suggested that an increase in the ionic strength leads to a decrease in the permeation rate (Khan et al. 2011). When it comes to current, the most commonly used in iontophoresis is a direct current. The other type of current used would be a pulsed current. The two are different in that, a direct current will be a constant administration of drugs, whereas a pulsed current will turn the delivery of the drugs on and off. If the pulse frequency is too high it may influence how well the drugs are transported. When done correctly it can overcome the problem that the direct currents face, which is that after some time there can be a reduction of the movement of ions (Khan et al. 2011). Since iontophoresis is an electrically assisted system, current strength and density are important factors when using this delivery system. It is important to have the right amount of current in the desired area with the right strength of current to be sure there are no harmful effects the patients skin. Typically, the maximum strength used is 0.5 mA/cm² (Khan et al. 2011). Therefore, if the electrode in use is 6 cm², then the milliamperage used should be 3. Since electrodes are the main source of pushing the medication into the skin, it is important that they not cause harm to the body and are flexible enough to be applied very close to the body. Commonly used electrodes for iontophoresis are, aluminum foil, platinum and silver/silver chloride. Silver/silver chloride is typically preferred because this electrode resists changes in pH (Khan et al. 2011). There are several factors of chemistry that are used in iontophoresis. Each factor plays
their own role in how successful the treatment can be and several of them are all connected to one another.

Biological factors are also very important to consider when using iontophoresis. One of these factors is the condition of the skin that is being treated. Testing in 1967 showed that areas with more hair follicles worked better for the diffusion of medicine than areas with less hair follicles (Khan et al. 2011). Regional blood flow is another factor to be considered. Although it has not shown much effect on the penetration of a drug through the epidermis is may determine how well it is absorbed in the tissue below (Dhote et al. 2011). All factors, whether biological or chemical, should be considered when using this for treatment to make sure the patients is not harmed and to get optimal results.

HOW IT WORKS

The basis for iontophoresis drug delivery is how the charges of ions work together, or against each other. The charged ions are put on the electrode that have the same charge. Positively charged medications are placed on the anode and negatively charged medications are placed on the cathode. The two electrodes are placed on the skin, usually with an adhesive patch, and are also connected to some type of power supply. The power supply is what provides the electrodes with the necessary electrical current to repel the ions into the skin. Both electrodes receive the current, but only one is considered active. The active electrode is the one that has the medication used for treatment on it. This electrode would be placed directed on the area that is being treated. When the power supply is turned on it starts to repel the medication into the skin. Then, because opposites attract, the medication starts to flow through the skin towards the other electrode. This is how the medications gets delivered to the area that is being treated. [Figure 1]

There is also another form of iontophoresis that uses a different type of setup. This involves a patient putting their hand or feet in plastic trays that are filled with water. These trays are shallow and have metal electrodes on the bottom, with an acrylic mesh on top of it. The electrodes are then connected to an iontophoresis machine and are provided with an electrical current to push the ions into the sweat glands of the areas being treated. The form of iontophoresis is one major difference than the first one discussed, and that is that it does not use any type of medication. It is often used with just tap water and because of that, it is also a treatment that patients would typically do on their own after buy the machine.

TECHNOLOGIES

With new advances in area of transdermal drug delivery, there are several technologies for this method of treatment. All the different devices are still using electrical currents to drive ions into the skin, but some of the products are very different and are used for different types of treatments. For example, some devices use one single patch that has both electrodes on it and has the power supply that connects right on top of the patch [Figure 2]. Other devices use a separate pad for each electrode and have lead that connect them to a power supply [Figure 3]. Even the power supplies are different. Some come in small, hand-held devices, where others may be larger and have to be placed on a table. Devices should have at least two controls on them. These controls would be for the duration/time of the treatment and for the current that is going to be used. The other type of device that could be used is the tap water iontophoresis machine. The
first two type of devices discussed are mostly used for therapeutic treatments whereas the tap water machine is used specifically for the treatment of hyperhidrosis [Figure 4].

**IONTOPHORESIS vs. OTHER DRUG DELIVERY**

Competition is a large part of the medical industry. With medicine, it is what drugs work the best for certain problems and there is always someone testing and researching to find something new, something better. Similar to the competition of medications, drug delivery methods are also constantly under development and improving. When a person thinks about putting medication into their body, oral medication or some type of injection immediately comes to mind. These are the two most commonly used methods for the administration of medication. When iontophoresis is thrown in the mix, it falls far behind on the popularity scale. Although it has been around for a long time, this method is still steadily growing. Development and research in transdermal drug delivery may be slower than its competitors, but there are several advantages over other delivery methods

When compared to injections, iontophoresis has some benefits. First, and probably one of the most common reasons someone would decide to pick this delivery method over injections is to avoid needles. Iontophoresis is a non-invasive method of drug delivery; therefore, needles are unnecessary. This means that potential problems like tissue trauma that could come from injections is minimized (Prakash et al. 2014). Transdermal drug delivery is also practically painless. Not having to deal with pricks of a needle and the soreness that can come along with it is main advantage of using iontophoresis. Although shots are almost always administered in a doctor’s office or clinic, a benefit of iontophoresis is that many machines are portable and can be used outside of that doctor setting. This allows for occupations like physical or occupational therapist to treat patients easily without having to worry about giving injections (Kirubakaran et al. 2015).

Iontophoresis also has a few advantages over oral medication. When medicine is ingested, it may take some time before it actually reaches the point where treatment is needed. It can also have adverse side effects on the patient’s body. Lastly, medication is typically dissolved in the body when taken orally, and this can lead to a loss of potency and efficiency of the drug (Kirubakaran et al. 2015). All of those downfall of oral medication can be avoided with iontophoresis. Iontophoresis allows the medication to be delivered directly on the affected area. This also means that there are not going to be any other side effects on other parts of the patient’s body, and that the medication will not lose any of its potency.

Although this method gives several benefits, there are also a few downfalls. Iontophoresis is somewhat limited compared to its competitors. There is only a handful of medications that can be used with this method. This restricts the kinds of treatments that are available. Another setback is that it is very possible to cause some type of damage to the skin. If the machine is not set up properly, there is an opportunity for chemical burns or soft tissue damage. Electrical shocks may also be caused by high current density (Khan et al. 2011). Iontophoresis is not the right treatment for all patients as well. Those with pace-makers or defibrillators may want to choose other treatment options because their devices may be extra sensitive to the electrical current being used in this method (Carlson 2008). Similar to any type of drug delivery method, iontophoresis has its advantages and disadvantages. It is only recently becoming more well known, so with time it may make a more significant mark in the medical industry.
TREATMENTS

Although there are limitations on the types of medication that can be used in iontophoresis, there are plenty of issues that can be treated. Several of these treatments have to do with dermatology, but it is also used in dentistry and therapeutics.

Dentistry was one of the first medical settings that iontophoresis was used in. It has been around since about mid to late 1800’s, but only has a few treatment options. Using negatively charged fluoride ions, dentists are able to treat hypersensitive teeth. Oral ulcers and herpes orolabialis lesions can be treated with negatively charged corticosteroids and antiviral drugs. Iontophoresis can also be used to apply local anesthetics to the mouth (Dhote et al. 2011).

Dermatology has many uses for iontophoresis. One of the main treatments in this area is hyperhidrosis. This is the condition of excessively sweating. This is usually an issue on a patient’s underarms, palms and/or the soles of their feet. Tap water iontophoresis would be the treatment of this condition. This treatment is commonly done in the patient’s own home and is also the most successful way to treat hyperhidrosis. Iontophoresis has also been used for treatments of ulcers, fungal infections and viral infections like warts and herpes simplex (Prakash et al. 2014). Iontophoresis can also be used in the diagnosis of cystic fibrosis. This is done by increasing the sweating patterns of the patient, then testing the concentrations of sodium and chloride in the sweat to confirm diagnosis. Iodine can be used in the reduction of scar tissue and iron/titanium can be used for tattoo removal (Kirubakaran et al. 2015).

Physical therapy is one of the few areas of the medical field that has introduced the usage of iontophoresis therapy. The goal of physical therapy is to help patients recover from injuries. In this practice, the treatment of pain is a major part of practice. Iontophoresis drug delivery system works very well in the physical therapy field for pain management. The main treatment in this area is of inflammatory conditions. This could be inflamed tissues, tendons, ligaments, or muscles. It can also be used where there are joint problems. When it comes to inflammation, dexamethasone is one of the most common medications used. Hydrocortisone and lidocaine are also commonly used in similar treatments that dexamethasone can be used for. Lidocaine actually gives an effect of a local anesthetic and the other two are specifically anti-inflammatories (Electrotherapy). Using iontophoresis in physical therapy or in any type of treatment can be very beneficial to the patient being treated. In many instances though, the medication will not be enough to magically cure the problem. Specifically, in physical therapy, it is necessary to have other forms of treatment like massaging, stretching or strengthening. Iontophoresis is one of many tools used to help patients recover from an injury, and the efficacy of this treatment depends on the type of injury, knowledge and skill of the physical therapist.

Another type of treatment that is fairly new in the transdermal drug delivery business is reverse iontophoresis. In this form of iontophoresis, certain molecules are going to be extracted through the skin. The cathode would be attracting positive ions and the anode would be attracting negative ions. The first efforts to use this method involved a non-invasive way of checking insulin levels. For diabetics that do not enjoy pricking themselves several times a day, this seemed like a great solution. It did not work out as expected, but it is still introducing new techniques into the iontophoresis world (Giri et al. 2017).
CONCLUSION

Iontophoresis has made great progress since its start in 1747. In the last few decades it has made even more significant progress. During this time, the first transdermal drug delivery system was introduced in America, and newer products have been hitting the market since then. When administering iontophoresis, it is important to understand how chemistry and biology are involved. It could be very easy to cause harm to someone using the treatment without a clear understanding of it. Compared to other forms of drug delivery, iontophoresis has some valuable benefits that other treatments cannot offer. The advantages of this method clearly outweigh the disadvantages. Transdermal drug delivery is often neglected when treatment options are discussed. It is still a growing area in medicine so the advantages are usually dismissed. In areas such as dermatology and physical therapy transdermal drug delivery has already made a presence in the treatments of several problems. Two of the most common uses for iontophoresis is the reduce excessive sweating to decrease inflammation in areas like joints, tendons, and soft tissues.

The research I conducted provided me with a broad view of iontophoresis. I would have liked to focus my research mostly on physical therapy, but I struggled to find sufficient journals to narrow it down to just that area. While researching I was curious to why there is a lack of growth in the transdermal drug delivery field when it had been around for so long. After looking more closely into the machines and systems of iontophoresis, I realized that this method is not the most cost-efficient. The prices can be steep compared to other methods, and constantly having to replace materials would start getting expensive for therapist and doctors looking to use this as a treatment. Another reason I think it has yet to gain popularity is because there really have not been any major ground-breaking research that grabs peoples’ attentions.

Iontophoresis still is not a common treatment that we hear every day. Hopefully as the growth of new technologies and the advancements in research will allow iontophoresis to expand. I think this method could be a big game changer in the drug delivery field. It has been around for so long, and its only recently getting more attention, but once it hits its stride, it could possibly become the next best thing.
FIGURES

Figure 1: Diagram of how iontophoresis works. Example with positively charged drug being pushed into the skin by the anode and is then pulled towards the cathode underneath the skin.

![Diagram of iontophoresis](image)


Figure 2: Smaller iontophoresis device with both electrodes on one patch.

![Smaller iontophoresis device](image)

FDA, Orthopedic and Rehabilitation Devices Panel. 2014. Classification of Iontophoresis Devices Not Labeled for Use with a Specific Drug. [Internet]. [Cited 19 April 2017].

302
Figure 3: Separate patch for each electrode, and small power supply with leads connecting to each patch

FDA, Orthopedic and Rehabilitation Devices Panel. 2014. Classification of Iontophoresis Devices Not Labeled for Use with a Specific Drug. [Internet]. [Cited 19 April 2017].

Figure 4: Tap water iontophoresis machine. Used for hyperhidrosis.

FDA, Orthopedic and Rehabilitation Devices Panel. 2014. Classification of Iontophoresis Devices Not Labeled for Use with a Specific Drug. [Internet]. [Cited 19 April 2017].
References

Carlson A. Therapy with a charge: iontophoresis offers therapists an effective way to address soft-tissue inflammation. rehabpub.com. 2008 Mar 5 [accessed 2016 Apr 7].


Physics in Nuclear Medicine

Nada Elebrashi

November 17, 2016

PHY112 – 26104

Dr. Casey Durandet
Abstract

This paper covers the research discovered by the author on the topic of physics in nuclear medicine. The reader will learn about the history of the discovery of nuclear medicine in addition to what nuclear medicine actually is. Nuclear medicine is commonly used in medical and health facilities to help diagnose and treat patients with specific illnesses. Physics plays a major role in nuclear medicine but the author will mostly be highlighting the key elements involved such as electromagnetic radiation, radionuclides, radiation dosing, and radiation therapy. The reader will be able to identify the importance of the use of physics in nuclear medicine to accurately diagnose and treat patients. This topic will be helpful for people entering the medical field or are already working in the medical field.

Introduction

“The very nature of nuclear medicine depends on physics since it deals with the interaction of the radiation emitted from within the patient with the detectors used to provide the images as well as with the patient him or herself,”¹. The discovery of nuclear medicine began in 1896-1898 by Henry Becquerel and Marie Curie when they initially discovered radioactivity of several elements. In 1913, Georg de Hevesy was the first to use the tracer approach in a biological system.² Nuclear medicine is used to identify certain diseases or illnesses present within the body with the use of minimal amounts of radioactive matter. Figure 4 shows what a typical nuclear medicine scanner looks like, and figure 5 is an example of nuclear medicine imaging results. It is different than other radiological exams that expose patients to radiation externally, in contrast to nuclear medicine that recognizes radiation coming from inside of the body.³ It is a very helpful technique when diagnosing illnesses, as well as recognizing the severity of them and how to treat them if possible. To get a better understanding of how a nuclear medicine scan works, the Radiological Society of North America (RSNA) best summarizes it as the following:

“Depending on the type of nuclear medicine exam, the radiotracer is either injected into the body, swallowed or inhaled as a gas and eventually accumulates in the organ or area of the body being examined. Radioactive emissions from the radiotracer are detected by a special camera or imaging device that produces pictures and provides molecular information.”⁴

Physics plays a crucial role in nuclear medicine as it applies electromagnetic radiation, radioisotopes, quantifying radiation dosing, and radiation therapy. The author will discuss in further detail how each physics aspect is applied in nuclear medicine studies.

Discussion

Electromagnetic radiation is the fundamental element of nuclear medicine that also involves physics. Simon R. Cherry and others declared two specific forms of radiation: particulate radiation and electromagnetic radiation. Particulate radiation contains atomic and subatomic particles that carry kinetic energy of mass in motion. Electromagnetic radiation carries energy by electrical and magnetic fields that travel through space at the speed of light.² Electromagnetic radiation can be expressed by:

\[ \lambda \times v = c \]

In this equation, \( \lambda \) is equal to the wavelength, \( v \) is the frequency, and \( c \) represents the velocity of light.
Radionuclides are consumed or injected through the body to transform an unstable nucleus into a stable nucleus while also converting mass into energy as they decay. The stability of the nucleus depends on the ratio of protons and neutrons and their arrangement inside the nucleus. The process of the unstable nucleus transforming into a stable nucleus allows for the energy to be emitted in the form of gamma (γ) rays, also known as photons, and be measured by a radiological instrument. Converting mass into energy is expressed in Einstein’s equation:

\[ E = \Delta mc^2 \]

where \( E \) equals the energy (also identified as nuclear binding energy) released, \( m \) is equivalent to the sum of the masses of the atom’s protons, neutrons, and electrons, and \( c \) is equivalent to the speed of light (approx. 3.00×10^8 m/s). \(^5\) Aside from gamma rays, alpha and beta rays may also be emitted in electromagnetic radiation but are not as detectable as gamma rays because they can easily be blocked. Beta emission occurs when a neutron inside the nucleus transforms into a proton and an electron given by the following expression:

\[ n \rightarrow p + e + v + energy \]

Beta particles are undetectable because they are only able to penetrate solids with minimal thickness. The most common type of nuclear medicine that emits gamma rays is called single photon imaging, and the other type of nuclear medicine is positron imaging which decays radionuclides through positron emission. \(^2\) The World Nuclear Association discussed how Positron Emission Tomography (PET) has proven to be the most precise way to non-invasively detect and evaluate most cancers thus far, which is a major development in the medical field. \(^6\)

Radioisotopes are tracers present in the radioactive substances that are either ingested by or injected into patients when a nuclear medicine study is performed. They are only used in small amounts to avoid harming the patients. Christopher Cooper explains in his book *The Basics of Nuclear Physics* that “their movements can easily be traced by detectors that generate a computer image of the body” and the places where the isotope concentrates appear on the image. Cooper also mentions how some radioisotopes can actually be used in treating tumors or other harmful growths. \(^7\) This is done by using strong radiation and focusing it on one area to destroy the harmful cells while trying to avoid the surrounding healthy tissues. It is very important that the radioactive substance being consumed or injected maintains a short half-life so it does not remain in the body and potentially harm the patient. According to Raymond A, Serway and Chris Vuille in their tenth edition of College Physics, “The half-life of a radioactive substance is the time it takes for half of a given number of radioactive nuclei to decay,”. \(^8\) The decay rate is expressed as:

\[ R = \frac{\Delta N}{\Delta t} = \lambda N \]

where \( N \) is equal to the amount of nuclei available, \( \Delta t \) is the time interval, and \( \lambda \) is the decay constant. The expression for half-life is defined as:

\[ N = N_0 \left( \frac{1}{2} \right)^n \]

where \( N \) is the amount of radioactive nuclei that remains, \( N_0 \) is the amount of radioactive nuclei initially present, and \( n \) is equivalent to the number of half-lives and is related to time. For example, if an individual starts off with three radioactive nuclei of a specific element that lasts for two half-lives, and they are looking for how much radioactive nuclei will remain then that can be expressed as:

\[ N = 3 \left( \frac{1}{2} \right)^2 \]
The individual can then conclude that \( N = 0.75 \), therefore, he/she will remain with 0.75 nuclei in his/her body after two half-lives. A typical graph plot of exponential decay of an isotope is shown in Figure 1, which is an example of a plot of radioactive decay of Nitrogen over time. Another expression to know that relates half-life to the decay constant is:

\[
T_{1/2} = \frac{\ln 2}{\lambda} = \frac{0.693}{\lambda}
\]

The units used for this equation are “becquerels”, 1Bq = 1 decay/s. 8

As mentioned before, radioactive decay involves an unstable nucleus transforming into a more stable nucleus by releasing particles and/or photons, while also converting mass into energy. 2 The unstable radionuclide is typically called the parent nuclide while the stable radionuclide is commonly called the daughter nuclide. 9 It is also important to note that these radionuclides enter the blood and interact with different enzymes, proteins, and cells in the body which may cause difficulty obtaining certain results depending on how well or not well the drug and the biological system interact. According to Jurrison and others in their article Coordination Compounds of Nuclear Medicine, “The radiopharmaceutical must be stable sufficiently long to reach its destination, and in some cases it must remain intact during its lifetime in the body.” 10

The radiation that is emitted from these nuclides must be strong enough to breach through the body and allow for an external instrument to measure the energy photons. Jurrison also indicated in his article that these emitted photons must have an energy level greater than 35 KeV. 10 Some of the photons can be absorbed, scattered, or successfully pass through the body. According to the World Nuclear Association, doctors and chemists have discovered that certain chemicals are absorbed by specific organs and, based on this knowledge, “…radiopharmacists are able to attach various radioisotopes to biologically active substances. Once a radioactive form of one of these substances enters the body, it is incorporated into the normal biological processes and excreted in the usual ways.” 6 An interesting fact about radioactive decay is that it is very spontaneous and impossible to predict when a given nucleus would decay so one may have a short lifespan while another may remain in the body for a very long time.

The most imperative part of this nuclear process is quantifying radiation dosing. Giving a patient a higher radiation dosing than their body can hold can potentially harm the patient. According to the RSNA, there are three different types of dosing: absorbed, equivalent, and effective. Absorbed dose is the energy absorbed by human tissue and is used to evaluate biological changes in specific tissues. Absorbed dose is measured in units of (mGy). Equivalent dosing is the amount of damage that is expected to be done to a tissue due to different types of radiation, and this is measured in units of milliSievert (mSv). The last and most important type of dosing is effective dosing which is a calculated value that identifies the absorbed dose in the body, the level of harm of radiation, and how each organ reacts to radiation depending on their sensitivity. Effective dosing is usually taken into consideration when evaluating the potential of long-term risks due to radiation, in contrast to absorbed and equivalent dosing that evaluate short-term risks. This is also measured in units of mSv. 11 Fred A. Mettler and others clarified in an article, “[E]ffective dose is a calculated age- and sex-averaged value that is used as a robust measure to compare detriment from cancer and hereditary effects due to various procedures involving ionizing radiation.” In other words, effective dose is calculated based on a patient’s age, sex, and organ dose. The organ dose is accustomed by the International Commission on Radiological Protection (ICRP) radiation weighting factors, which is then multiplied by the organ’s equivalent dose. In Figure 2, a table is shown from Mettler’s article that displays “Effective Doses for Adults from Various Nuclear Medicine Examinations”. 12
information, it is crucial for a physician to understand effective dosing and how to calculate it in order to avoid causing any harm or long-term effects to the patient.

In College Physics, Serway discussed the damage radiation can cause to the body and classified them into two types: somatic and genetic damage. Somatic damage affects all cells except reproductive cells, while genetic damage affects only reproductive cells. Exposure to radiation is quantified in units of rad (radiation absorbed dose) and Serway states, “One rad is the amount of radiation that deposits $10^{-2}$ J of energy into 1 kg of absorbing material.” He also noted an important factor of radiation to measure which is relative biological effectiveness, or RBE, and this is defined as “the number of rads of x-radiation or gamma radiation that produces the same biological damage as 1 rad of the radiation being used.” In Figure 3, a table displays RBE factors for several types of radiation.

Interestingly enough, nuclear medicine is not only used to identify illnesses, but could also be used to treat them. Some illnesses may be treated via radioactive-iodine therapy or radioimmunotherapy (RIT). Radioactive-iodine therapy is used for thyroid disease and involves swallowing radioactive iodine (I-131) as it is absorbed by the thyroid gland and destroys cancer cells. 13 This kind of therapy is typically used to terminate any thyroid tissue that remained after a surgery, or to treat any type of thyroid cancer that may have spread to other organs in the body. According to the American Cancer Society, radioactive iodine therapy is most effective in patients that have high levels of the thyroid-stimulating hormone in their blood. 13 Radioimmunotherapy is another kind of therapy that combines radiation therapy with immunotherapy to directly attack a tumor with a high dose of radiation. It involves a monoclonal antibody being injected into the bloodstream along with a radioactive substance to travel and bind to the cancer cells before destroying them. This is able to happen because monoclonal antibodies imitate natural antibodies in the body’s immune system that function to attack foreign substances such as bacteria, viruses, or cancer cells; therefore, they lead the radioactive substance directly to the cancer cells. Other types of radiation therapy identified by the National Cancer Institute include but are not limited to: external-beam radiation therapy, 3-dimensional conformal radiation therapy, intensity-modulated radiation therapy, tomotherapy, proton therapy, and stereotactic radiosurgery. 14 The main purpose of all of these different types of therapies is to target the tumor or cancer cells and directly attack them while trying to avoid healthy cells as much as possible. Although, there are some serious side effects that result from these kinds of radiation therapies. One major side effect is the body still giving off radiation after receiving treatment, requiring the patient to remain isolated in the hospital for a few days to prevent exposing radiation to others around them. Radiation therapy could also possibly lead to infertility depending on what area of the body is receiving the therapy. Some short-term side effects include nausea, vomiting, fatigue, and hair-loss. 13

There are many significant pros and cons to nuclear medicine which are important for health care providers and patients to be aware of. Some benefits include being able to diagnose many different diseases and illnesses non-invasively, being able to use radiation to treat patients, and it is proven to be most effective in detecting and treating cancer. Some disadvantages to nuclear medicine include being exposed to high doses of radiation, which can lead to cancer rather than preventing it, and also being costly. It costs a lot for facilities to have the equipment and it costs a lot for the patients to have these scans done. Although there are alternative methods to be used in place of it, nuclear medicine has still proven to be the most effective in diagnosing and treating severe illnesses.
Conclusion

The research completed on physics in nuclear medicine was a very fascinating topic for me as I am planning to go into the medical field. I started my research barely knowing anything about this topic and I am glad I chose it because I learned so many new things about it. For example, I always thought nuclear medicine was only used for diagnosing but I never knew it was also used for treatments and therapy as well. In addition, I knew a little about how a nuclear medicine scan was performed but I never knew about gamma rays being transmitted through the body until I did my research. It is so amazing to me how these studies can detect cancer so efficiently and also provide radiation treatment. Unfortunately, there are many factors present in our environment today that potentially lead to a high risk of cancer such as global warming and consuming so many processed foods. Thankfully, we have these types of nuclear medicine scans that help us detect cancer almost immediately. Although the thought of being injected by a radioactive substance may be a little scary to some people, I feel that it will do more good than harm because we would want to prevent cancer cells from spreading as fast as possible. A few things I hope that can change in the future are making these exams more cost-efficient and also finding safer ways to treat patients with it. I feel that this technology has greatly improved modern medicine, and I predict that nuclear medicine will continue to progress in the future. I am looking forward to seeing how great of an impact nuclear medicine will make on our society in the future.
Figure 1: Example of a plot showing radioactive decay of Nitrogen over a period of time.

**Figure 2: Effective Doses for Adults from Various Nuclear Medicine Examinations**

http://pubs.rsna.org/doi/full/10.1148/radiol.2481071451

<table>
<thead>
<tr>
<th>Examination</th>
<th>Effective Dose (mSv)</th>
<th>Administered Activity (MBq)</th>
<th>Effective Dose (mSv/MBq)</th>
</tr>
</thead>
<tbody>
<tr>
<td>Brain (99mTc-HMPAO-exametazime)</td>
<td>6.9</td>
<td>740</td>
<td>0.0093</td>
</tr>
<tr>
<td>Brain (99mTc-ECD-NeuroLite)</td>
<td>5.7</td>
<td>740</td>
<td>0.0077</td>
</tr>
<tr>
<td>Brain (131I-FDG)</td>
<td>14.1</td>
<td>740</td>
<td>0.019</td>
</tr>
<tr>
<td>Thyroid scan (sodium iodine 123)</td>
<td>1.9</td>
<td>25</td>
<td>0.075 (15% uptake)</td>
</tr>
<tr>
<td>Thyroid scan (131I-percheketinate)</td>
<td>4.8</td>
<td>370</td>
<td>0.013</td>
</tr>
<tr>
<td>Parathyroid scan (99mTc-sestambit)</td>
<td>6.7</td>
<td>740</td>
<td>0.009</td>
</tr>
<tr>
<td>Cardiac stress-rest test (tahium 201 chloride)</td>
<td>40.7</td>
<td>185</td>
<td>0.22</td>
</tr>
<tr>
<td>Cardiac rest-stress test (99mTc-sestambit-1 day protocol)</td>
<td>9.4</td>
<td>1100</td>
<td>0.0085 (0.0079 stress, 0.0090 rest)</td>
</tr>
<tr>
<td>Cardiac rest-stress test (99mTc-sestambit-2 day protocol)</td>
<td>12.8</td>
<td>1500</td>
<td>0.0085 (0.0079 stress, 0.0090 rest)</td>
</tr>
<tr>
<td>Cardiac rest-stress test (t-tetofosmin)</td>
<td>11.4</td>
<td>1500</td>
<td>0.0076</td>
</tr>
<tr>
<td>Cardiac ventriculography (99mTc-labeled red blood cells)</td>
<td>7.8</td>
<td>1110</td>
<td>0.007</td>
</tr>
<tr>
<td>Cardiac (18F-FDG)</td>
<td>14.1</td>
<td>740</td>
<td>0.019</td>
</tr>
<tr>
<td>Lung perfusion (99mTc-MAA)</td>
<td>2.0</td>
<td>185</td>
<td>0.011</td>
</tr>
<tr>
<td>Lung ventilation (xenon 133)</td>
<td>0.5</td>
<td>740</td>
<td>0.00074</td>
</tr>
<tr>
<td>Lung ventilation (99mTc-DTPA)</td>
<td>0.2</td>
<td>1300 (40 actually inhaled)</td>
<td>0.0049</td>
</tr>
<tr>
<td>Liver-spleen (99mTc-sulfur colloid)</td>
<td>2.1</td>
<td>222</td>
<td>0.0094</td>
</tr>
<tr>
<td>Biliary tract (99mTc-disofurin)</td>
<td>3.1</td>
<td>185</td>
<td>0.017</td>
</tr>
<tr>
<td>Gastrointestinal bleeding (99mTc-labeled red blood cells)</td>
<td>7.8</td>
<td>1110</td>
<td>0.007</td>
</tr>
<tr>
<td>Gastrointestinal emptying (99mTc-labeled solids)</td>
<td>0.4</td>
<td>14.8</td>
<td>0.024</td>
</tr>
<tr>
<td>Renal (99mTc-DTPA)</td>
<td>1.8</td>
<td>370</td>
<td>0.0049</td>
</tr>
<tr>
<td>Renal (99mTc-MAG3)</td>
<td>2.6</td>
<td>370</td>
<td>0.007</td>
</tr>
<tr>
<td>Renal (99mTc-DMSA)</td>
<td>3.3</td>
<td>370</td>
<td>0.0088</td>
</tr>
<tr>
<td>Renal (99mTc-glucocorticohionate)</td>
<td>2.0</td>
<td>370</td>
<td>0.0054</td>
</tr>
<tr>
<td>Bone (99mTc-MDP)</td>
<td>6.3</td>
<td>1110</td>
<td>0.0057</td>
</tr>
<tr>
<td>Gallium 67 citrate</td>
<td>15</td>
<td>150</td>
<td>0.010</td>
</tr>
<tr>
<td>Pentetide (111mIn)</td>
<td>12</td>
<td>222</td>
<td>0.054</td>
</tr>
<tr>
<td>White blood cells (99mTc)</td>
<td>8.1</td>
<td>740</td>
<td>0.011</td>
</tr>
<tr>
<td>White blood cells (111mIn)</td>
<td>6.7</td>
<td>18.5</td>
<td>0.360</td>
</tr>
<tr>
<td>Tumor (18F-FDG)</td>
<td>14.1</td>
<td>740</td>
<td>0.019</td>
</tr>
</tbody>
</table>

**Table 29.3 RBE Factors for Several Types of Radiation**

<table>
<thead>
<tr>
<th>Radiation</th>
<th>RBE Factor</th>
</tr>
</thead>
<tbody>
<tr>
<td>X-rays and gamma rays</td>
<td>1.0</td>
</tr>
<tr>
<td>Beta particles</td>
<td>1.0–1.7</td>
</tr>
<tr>
<td>Alpha particles</td>
<td>10–20</td>
</tr>
<tr>
<td>Slow neutrons</td>
<td>4–5</td>
</tr>
<tr>
<td>Fast neutrons and protons</td>
<td>10</td>
</tr>
<tr>
<td>Heavy ions</td>
<td>20</td>
</tr>
</tbody>
</table>

**Figure 3: RBE Factors for Several Types of Radiation**

Figure 4: Image of nuclear medicine equipment
http://www.camisrd.com/services/nuclear-medicine/

Figure 5: Example of nuclear medicine imaging
http://www.svmedicalimaging.com/services/nucmed.html
References
https://www.cancer.gov/about-cancer/treatment/types/radiation-therapy
Abstract:

The main objective of this paper is to convey the physics involved in surfing. This will be achieved by first explaining the physics that occurs in the sea, including topics like wavelength, wave frequency, Navier Stokes equation, and so on. Using those principles as a base, the physics of surfing will be explained. Much of the physics of surfing is comprised of topics such as drag, buoyancy, lift, gravity, and so on. This paper will also show how surfboards are built to maneuver certain types of waves and how each board is affected by the laws of physics. Finally, once understanding and harnessing all these laws of physics, surfing can be carried out properly.

Introduction:

The sport of surfing dates all the way back to the time of the Ancient Hawaiian. In ancient times, long before Europeans ventured through the Pacific, it was referred to as he’e nalu. He’e means to ride, and nalu means the surf, making it’s fitting, literal translation “to ride the surf”. It was a pastime that all people could participate in, no matter which social status people identified with. Men and women of all ages could surf together, even commoners among kings. Physical, athletic ability and prowess was praised in Ancient Hawaiian culture and many would train their bodies in order to be able to keep fit. Surfing was also a spiritual and religious experience that allowed Ancient Hawaiians to connect with the world and be one with the sea. Kahunas, or ancient priests, would pray over the surfboards, thus sanctifying them and declaring them as holy sporting tools. There were also associated prayers and chants, called meles, that were associated with surfing.1

However, as the islands in the Pacific began being visited by Europeans in 1778, traditional cultures began to fade away. The first group of Europeans were led to Hawaii by Captain James Cooke, and from then on, the Hawaiians began to lose economic, social, and religious freedoms that they had previously enjoyed. Many traditions were lost forever as the islands were flooded with missionaries who sought to evangelize the them.2 Although contact with European evangelists made surfing much less popular, the sport was revived in the early 1900’s to bring more tourists to the islands. Sadly, surfing had lost much of it’s prestige and spiritual meaning by that time. It became extremely commercialized during the revival. As surfing resurfaced, the sport was quickly recognized and spread around the globe to places like California and Australia.3

Although surfing has lost much of it’s religious meaning, it is still an amazing activity. Gliding along a rushing wave is by no means an easy feat. It takes practice, skill, and involves a lot of physics. Not only does the ocean play by its own physics rulebook, humans must be able to adjust their surfing methods to the ocean’s dynamics. To do this, humans created a system that followed the laws of physics in order to balance themselves with the sea.

Discussion:

In order to understand surfing, it is necessary to first understand the physics of the ocean, specifically how waves work. So, where do waves come from? All waves come from disturbances. Most waves are a result of storm winds blowing along the surface of the sea,
otherwise referred to as the air-water interface. Often, storms occur out in the middle of the sea, kick up a bunch of water, and waves are created. Other waves come from energy being released in the ocean. Some examples of this are internal waves, caused by bodies of water with different densities converging, splash waves from objects crashing into the ocean, or waves caused by humans. These waves travel miles and miles until they reach the shore. Here, Newton’s 1st Law can be applied. This law states that an object in motion will stay in motion until it is acted on by another force. In this case, a wave will stay in motion until it encounters another force, the shore.\(^4\)

It is important to know that, in waves, it is not actually the water moving through the sea. In fact, waves are a result of kinetic energy surging through the water as a result of the wind or energy put in the water. The actual water particles stay in place as the energy moves through the body of water. There are also multiple types of waves. One type is a progressive wave, which oscillates in a uniform pattern without breaking. This type of wave can be longitudinal, orbital, or transverse. A longitudinal wave follows a motion similar to a spring or coil, being that as the energy surges through it the wave, the water particles compress and expand as a spring would. They are nicknamed push-pull waves and move in the same direction that the energy is flowing. Also, transverse waves are classified as waves that move side to side while the energy moves perpendicular to the water. This type of wave form moves up and down. Ocean waves are a combination of longitudinal and transverse waves. They are known as orbital waves and this wave type transports energy along the atmosphere and ocean interface.\(^4\)

Wave lengths are measured from crest to crest, or from the tip of one wave to the next tip. They can also be measured from trough to trough, or from the lowest point of the wave to the next lowest point. A diagram of this can be seen in Figure 1 in the Figures section. The steepness of a wave can be calculated by solving for the height of a wave by measuring from trough to crest, and then dividing by the length of a wave \(\text{wave height (H)} / \text{wave length (L)}\). For instance, if a wave is 20 ft high and 10 ft long, the steepness of the face would be 2 feet (H/L). Also, a wave period is the amount of time (T) it takes for a full wave, crest to crest, to pass a fixed location. Next, the frequency of a wave can be defined as the number of wave crests that pass by a fixed location per unit of time. Wave frequency can be calculated using the equation \(F(f)=\frac{1}{T}\). Therefore, if waves have a period of 10 seconds, they have a frequency of 0.1 waves per second. This means the waves have a frequency of 6 full waves per minute. Lastly, wave speed (s) can be calculated using the equation \(\frac{\text{wavelength (L)}}{\text{period (T)}}\). For example, if a wave is 20 feet long and has a period of 10 seconds, then it’s speed would be 2ft/sec, or 0.6096 m/s.\(^4\)

Another important equation that can be used to understand the ocean is the Navier Stokes equation. It is used as a fluid dynamic model and is often referred to as Newton’s 2nd Law applied to water. Newton’s 2nd Law of motion can be summed up as force is equal to mass times acceleration (F=ma). Utilizing this equation can lead to an explanation of why breaking waves give much of their energy to the atmosphere and the air, often leading to the formation of clouds and other climate change.\(^5\)

The equation states: \(P \left(\frac{dv}{dt} + \nu \nabla \nu\right) = - \nabla p + \mu \nabla^2 \nu + f\). Here, P is defined as density. Also, \(\frac{dv}{dt}\) represents unsteady acceleration, \(\nu \nabla \nu\) is convective acceleration and together they give the
inertia of a wave. Next, $\nabla p$ shows the pressure gradient, $\mu \nabla^2 v$ is the viscosity, and together they give the divergence of stress. Lastly, $f$ accounts for other body forces acting on the system. This equation overall gives the incompressible flow of waves. It also explains why longer waves travel faster in the ocean. This is because the equation takes into account the boundary conditions of the upper surface of the sea and the boundary conditions at the bottom. It also helps find the dispersion rates of waves as they travel through the ocean.

So, since all waves are basically the same, why are some waves bigger and scarier than others? This can be understood using an east coast versus west coast of the U.S. comparison. On the east coast, waves are typically smaller and gentler. In contrast, on the west coast you get some of the biggest waves on the planet. The first factor that contributes to wave size, again, is the wind. On the west coast, the prevailing winds blow behind the waves, making them more and powerful as energy builds and rolls through. However, on the east coast, the prevailing winds blow against the waves. Here, we can again see Newton’s 1st Law being applied. The waves on the east coast are being met with an opposite force, the wind, and they are being slowed down while their energy decreases. Another factor that changes wave size is the continental shelf. The east coast has a gently sloping continental shelf. As the waves move towards the shore from out in the open ocean, they gradually come into contact with the shelf. This gives the waves a chance to slow down because the water particles are colliding with the continental shelf and losing energy to friction. The friction allows waves to gradually collapse on themselves, making them smaller and gentler. The west coast, on the other hand, has a continental shelf that rises suddenly as the sea comes towards the shore. Here, the waves have barely any time to be slowed by friction. This causes the waves to generally maintain their size and be pushed up rapidly, resulting in larger waves. A diagram of this can be seen in Figure 2. The last factor that affects wave size is fetch, or the distance the wind has to blow across the sea. The longer the distance the wind has to blow over, the longer the waves must travel through the sea, and the more time their energy has to build. All this time allows waves to grow in size and length. On the west coast, storms far out in the Pacific ocean cause most of the waves. These waves have far to go before they come in to contact with the coast. This results in larger waves than the east coast, which often does not have storms in the Atlantic to affect the Pacific.

With all those factors in mind, here are some of the biggest, scariest waves on the planet. To start off, there is Pe’ahi, otherwise known as Jaws. Pe’ahi is nicknamed Jaws for a reason. Not only did the classic movie, *Jaws*, recently grace the screens around the time Pe’ahi was starting to be surfed, but the waves were vastly unpredictable, just like a shark. People warn that if surfers are not careful while they surfed Jaws, big waves will come out of the blue and catch them. Jaws is a surf spot on the north shore of Maui, Hawaii with waves that range from 25-50ft high during north shore season. It is the place that premiered tow-in surfing due to the waves being so powerful. In order for surfers to be able surf the wave, they have to be towed in by a jet ski. This is so they can reach the rapid acceleration of the fast pumping, powerful wave and actually catch the wave.

Another place with an extremely powerful wave is Teahupo’o on the island of Tahiti. This wave makes it’s debut over a sharp, shallow reef that rises up abruptly from the sea floor. Tahiti is a volcanic island that has formed due to volcanic activity and growing reefs, also known as an archipelago. But, this is a special place because the reefs grow vertically up from the sea bed. Also, the island is surrounded by deep sea channels, allowing any waves that break on shore to be quickly carried back out to the ocean. The waves at Teahupo’o are considered surging
waves, due to the fact that they come from deep water to such shallow water at a rapid speed. This rapid change results in fierce, hollow waves with a 20 ft face on average. These waves can grow up to 30 ft tall and are as thick and heavy as a building. They have such power that, often times, the trough of the wave occurs below sea level as the water is being sucked up from the reef and into the wave.9

With this new understanding of how the ocean works, surfing can be better comprehended. This is another area that is jam-packed with physics concepts. If these concepts are utilized correctly, surfing can be a piece of cake. The main thing needed to surf successfully is a surfboard. There are many types of boards, but the different, general components of boards must be understood before any specific board can be utilized. The first thing surfers can look at is the overall plan shape of the board. This is like the template of the board and shows the details of the top of the deck of the surf board. The more area there is on the plan, the greater the buoyancy the board will have in the water. Next is the outlining curve of the board. This can be found on the sides of the board and it is said that the rounder the outline curve, the easier it is to turn. Also, a shorter outline allows easier turning as well. Turning or slowing can be accomplished if the surfer applies torque at various parts of the board. For instance, torque is applied while the surfer leans back on the rail of the board and moves the board laterally, causing the board to rotate. The next aspect of the surfboard is length. The longer the board, the more room there is for increased foam volume in the board and also greater buoyancy. This allows for easier paddling due to more buoyancy forces pushing the surfer out of the water. This buoyancy also helps the surfer move across the surface of the water rather than through it, resulting in less friction and drag. However, with more length comes a higher risk of nose diving. Balance, center of gravity, and weight distribution are crucial here. Also, with more length, it becomes harder to turn the board. This is because on long boards, where the surfer typically stands at the back of the board, there is more swing mass at the nose. Turning can become stiff due to the amount of torque and power needed to turn a large board. An additional component of the surfboard is width. The wider the board, the looser turns are at slow speeds. This is because wider boards sit higher on the water, instead of sitting in or below the surface. This results in less resistance, drag, and friction between the board and the water particles. Therefore, it is easier to turn. However, the wider the board, the more drag there will be when paddling for a wave. This is due to an increase in surface area contact between the water and the board. Also, as boards increase in width, there is more stability while the board is lying flat on a wave. Load capacity and stiffness increase while board width increases, as well.10

Now, look at the ends of the board. At the top is the nose, and at the bottom is the tail. With a round nose comes more life and buoyancy due to increased surface area. On the other hand, a pointed nose has less buoyancy and lift, but give way to a more seamless, holding rail turns. The final aspect of the surfboard is the tail. The types of tails vary from pin shapes, to rounded, to square, and so on. A diagram of different tails can be seen in Figure 3. The more narrow the tail, the more hold it will have in the water. In contrast, the more area the tail has, the less hold.10

Different types of boards are used for different types of waves. One common type of board is a shortboard. They are usually around 7 ft long and are made for aggressive surfing and maneuverability. The short length and thin rails make them perfect for doing tricks on waves from waist high to 10 ft in height. Another type of board is the longboard. These classic, 8 to 12 ft boards were often used by the Ancient Hawaiians to glide along the waves. They are typically
2.5 inches thick and 20 inches wide, making them buoyant and good for beginners to learn on. However, they are hard to maneuver due to their large size. A final board example is a gun board. They usually range from 6’6 to 10 ft in length and have good curve outlines for maneuverability. These boards are used for big waves with big drops and high speeds.

Now that all the foundation has been set on the laws of the sea and surfboard mechanics, the sport of surfing can be analyzed using physics. There are multiple forces that act on a person while surfing. First is buoyant force, or the weight of water displacement due to the board and the surfer. This occurs when a less dense object than water sits on the surface of the ocean and is pushed upward. The force of buoyancy is also equal to the displacement of fluid. The more buoyant the board and surfer, the easier it is to paddle. The next force that acts on a surfer is gravity, which amounts to the combined weight of the board and the surfer. While surfing, gravity pulls the surfer to the trough of the wave, down the face, and towards the shore. The less weight in the system, the higher the board will sit on the surface, and the easier it will be to paddle. Also, the more vertical, or perpendicular, the weight is to the wave, the more speed the surfer will be able to pick up. Therefore, if the wave is relatively steep and the weight is perpendicular to the wave, the more the forces align and allow for a drop, resulting in the surfer moving at higher speeds. It is also important to keep the center of gravity of the surfer and board aligned with the force of gravity. This results in more balance in the system overall. The next force acting on a surfer is drag, or the water-surf interactions that result in the slowing of the board. Drag moves opposite to the direction that the surfer moves in. In surfing, there are two types of drag. The first type is friction drag, which results from interactions between the bottom of the surfboard and the water flowing along it. At high velocities, the drag force increases. The other type of drag is form drag. This force occurs when water is pushed aside as the surfer paddles through it. Again, the more speed a surfer picks up, the more form drag. Lastly, there are various forms of lift forces that can apply while surfing. Lift is described as a force that drives a surfboard up towards the surface of the water. If a person is surfing at high velocity, there will be more planing lift. This is similar to form drag, as the board will be pushing water out of the way and travel up higher in the water at high speeds, therefore lifting the board. Another form of lift occurs when water is flowing up the wave face. Here, the orbital wave shape becomes more apparent and water particles start pushing up against the bottom of the surfboard.

The first step to actually surfing is to grab an appropriate board and paddle out. It is important to paddle out in a channel where the water is flowing back out to the sea. This way, surfers will not have to fight the much more powerful forces of waves coming towards them and can save energy. The next step is to choose a good wave. Waves that are still flat are not ideal, along with waves that are already breaking. Waves that are half-built are the best. They give the chance of a clean entry and can be caught with proper paddling. They also can help the surfer create momentum and velocity that allow for a good, long ride.

Once a wave has been chosen, the surfer must paddle accordingly to catch it, utilizing the laws of physics. The surfer must paddle and get themselves at an equal acceleration to the wave. If are paddling too fast or too slow, they will not catch the wave. Also, the larger the wave, the faster and harder the surfer must paddle. Once the wave has been caught, the surfer needs to vertically pop up on the board, keeping the center of gravity in order to avoid the system coming out of balance. When all these steps have been accomplished, the person is officially surfing! The whole system is in balance, with buoyancy/lift keeping the board above surface and gravity pulls down. Drag pulls the board, while the waves keeps pulsing the surfer forward. A diagram
of this can be viewed in Figure 4. Also, Newton’s 3rd Law is in full affect here, as every action is counteracted and the system stays balanced.6

With the innovations of the 21st century affecting nearly every aspect of life, it would be impossible to think that new technology is not being created for the surfing world. Flex has been a hot topic in recent the surfing world and companies that specialize in surfboard technology are hopping on board to utilize it. In surfboards, flex can be described as projection. It is what happens when a surfboard builds up energy while surfing various parts of waves. For instance, if a surfer takes a drop down the face and to the trough of the wave, all the forces of the system are working together and building. This allows the speed and energy of the system to increase greatly, propelling the surfer forward. This flex energy can then be harnessed by a surfer if they turn the board and project themselves into a new maneuver. A company called Firewire is creating new surfboard flex technology utilizing Direct Drive. Firewire has created surfboards with the idea of flex in mind. Direct Drive technology has been incorporated into the boards in the form of interior carbon rods. The rods have been put into the frame of the board in a parabolic, suspended configuration, which creates a stiffer and snappier board. The rods are also connected to high density nodes in the interior foam of the surfboard. The configurations can be customized to each individual surfer if they desire to change anything. Firewire claims that this new design will give surfers more control over the rate of flex experienced. The technology helps to create more flex, as well as allow for the building of momentum and spring. Although these boards could be revolutionary, they are currently too expensive to manufacture and sell. Even though they are not feasible now, in the future they could be a game changer.9

Out of all the surfers in the world, there are a few who stand out as some of the best. First, of course, is the world-renowned Kelly Slater. He had won the Championship tour 11 times over the course of his career and is one of the best competitors out there. Another outrageously successful surfer is Bethany Hamilton. Although she may not be as competitive as she used to be, she has worked past being attacked by a shark and losing her arm to become a leading female figure in the surfing community. Lastly, the best surfer in the world as of 2016 is John John Florence. Florence was born on the north shore of Oahu and grew up surfing Banzai Pipeline, arguably the most dangerous, famous wave on the planet. He is currently leading the Men’s Championship Tour with 56,400 points. To gain the Championship title, surfers participate in 10 tour stop competitions around the world from March to December. These stops are put together by the World Surf League. The better the surfers place in these competitions, the more points they accumulate and the higher their rankings. At the end of the year, whoever is number 1 wins the Championship Tour.11

Conclusion:

Surfing is a pastime that goes much further back than any typical sport of modern days, such as baseball, football, or basketball. It is rooted deep in rich traditions that date back to the times of Ancient Hawaiian kings and priests. It has undergone many trials and tribulations, almost causing the tradition to disappear permanently. However, it has stood the test of time.

Not only is surfing culturally rich, but it is strewn with different concepts of physics. The ocean itself is full of laws. For instance, Newton’s 1st Law of motion is observed every time a wave moves from the deep sea and is slowed and stopped by the shore. Navier Stokes equation
can also be applied, alongside Newton’s 2nd Law of motion. Waves can be measured by speed, steepness, period, and even frequency. To me, the most interesting idea about the ocean, however, is the fact that the waves are not a result of the water particles moving. Instead, waves are a representation of kinetic energy rolling through the sea.

More specific to the actual sport of surfing, there are many other concepts of physics that are involved. Certain boards are specifically designed to take on certain breaks. Some boards are made to glide along the surface gracefully, while others are designed to shred. Also, so many forces affect a surfer and their board. Drag, lift, buoyancy, acceleration, gravity, and so on are all factors that act on a person while they surf. If a person can achieve balance between all these forces, they can surf. In regards to the future of surfing, the opportunities are limitless. I doubt that the surfing innovations will stop at flex technology. People will want more. With the ocean being explored more and humans gaining more knowledge about it, in combination with more innovations, I think surfing will grow. Surf culture is already such a big part of coastal cities. With an increased amount of technology and interest, I believe surfing will become more appealing and accessible to others.
Figures:

Figure 1: The figure above illustrates wavelengths being measured from crest/crest and trough/trough. (Picture by Sydney Esposito)

Figure 2: The picture above shows how wind and continental shelves change how waves form and wave size. (Picture by HowStuffWorks)
Figure 3: Figure 3 above shows three different types of surfboard tails. (Picture by Sydney Esposito)

Figure 4: Figure 4 illustrates all the forces involved with surfing. (Picture from PHD Comics)
References:


6. The Physics of Surfing. Piled Higher and Deeper (PHD Comics); 2012 [accessed 2016 Nov 17]. https://www.youtube.com/watch?v=s8u7ba6yrpe


Abstract:

The purpose of this research is to identify the methods by which mammals in the orders cetacea and microchiroptera utilize echolocation for hunting and navigation. This paper first describes the basic physics of sound waves. A brief history of the discovery of echolocation is provided. Next there is a section about the process by which microchiroptera use different types of pulses to detect and then locate their prey. Lastly, a description of cetacean echolocation is presented and discussed in how it varies from microchiropteran echolocation.

There are two major taxonomic orders of mammalia that use biosonar as a means of communication and hunting. These two orders are microchiroptera and cetacea. The order microchiroptera includes all bats whereas the order cetacea includes dolphins, porpoises, and some whales (Hale WG, Margham JP, 1988). Cetacean is defined by The Harper Collins Dictionary of Biology as follows, “any aquatic Eutherian mammal of the order Cetacea, comprising the porpoises, dolphins, and whales. . . Cetaceans have an extensive outer layer of fat associated with their marine habitat, have a large brain, and are capable of complex communication,” (Hale WG, Margham JP, 1991). The fact that these mammals rely so heavily on sound waves is simply astonishing. From the creation of an outgoing signal, to the reception of the resultant sound wave, the process is very complex and relays very important information to the mammal that originally emitted the echo. The goal of this essay is to explore the physics behind echolocation performed by cetacea and microchiroptera. The complex processes of emission, reception, and interpretation of the echo will be explained in great detail by using figures, equations, and simply explaining with words.

Before grasping the concept of echolocation it is pertinent to have some knowledge on the physical properties of sound waves. Frequency, amplitude, wavelength, and Doppler shift are just a few of the concepts that one should be well versed in before attempting to understand how echolocation or biosonar works. Wavelength and amplitude are the simplest concepts pertaining to wave motion. Wavelength is the distance between two peaks or two troughs of a wave. Amplitude is the displacement from the midline of the wave to the peak of the wave. The distance from the midline to the peak is the same as the distance from the midline to the trough of the wave. There are two commonly known waves that exist: transverse and longitudinal. According to the physics text, College Physics, “Sound waves are longitudinal waves traveling through a medium, such as air,” (Serway RA, Vuille C, 2015). An example of what a longitudinal wave looks like can be seen in Figure 1. For microchiroptera, which will be referred to as bats for the remainder of this article for the sake of simplicity, the medium through which these sound waves travel is air. In contrast, water is the medium that longitudinal sound waves travel through for cetacean echolocation. As a general rule it can be noted that sound waves travel much faster through water than they do through air. To be exact, sound travels four times faster in water according to John Montgomery, a professor at the University of Waikato (Montgomery J, 2011). This difference in speed will be an important factor to keep in mind for the remainder of this research paper. The question begs to be asked, “So what accounts for the difference in speed of a sound wave in water versus air?” One of the main differences between water and air is how densely packed their molecules are. Water is more densely packed than air which means that it takes more energy to produce a sound wave in water than it does in air (Montgomery J, 2011). However, once that wave has started propagating, it will move at a much faster rate in water than it would in air. Sound waves occur by particles bumping into one another and transferring their vibrational energy (Montgomery J, 2011). Because the water
molecules are so densely packed, they have less distance between them before they make contact with one another. They are able to collide with one another faster and transmit the sound wave in a short span of time. In summary, sound waves in water move much faster but require more energy to create and sound waves in air move at about one fourth of the speed of a sound wave in water but they do not require as much energy to make them. To be precise, the speed of sound in water is 1,376 meters per second and the speed of sound in air is 344 meters per second (Hughes HC, 1999). The next factor that is important when discussing sound waves is frequency. Frequency can be thought of as the number of cycles, or oscillations, per second. Frequency is measured in Hertz which is abbreviated as Hz. Frequency is a measure of how often something occurs in a given time frame. The frequency of the outgoing echo created by bats and cetacea is of the utmost importance and will be explained in greater detail later in this essay. The next factor that will be paramount in understanding echolocation is the Doppler Effect. The Doppler Effect is the phenomena that most everyone has observed at least once in their life. An example of it is when a person in a car and an ambulance with its siren on are travelling towards each other. As they approach each other, the observer in the car will notice that the frequency of the siren seems to increase as the ambulance comes closer. As the ambulance passes the observer and moves farther away, the observed frequency of the siren will gradually begin to decrease. There are several different factors at play which include a moving source of frequency, a stationary source of frequency, a stationary observer, and a motile observer. The Doppler Effect can be calculated by using the following equation provided by *College Physics*.

\[
\nu = \frac{\nu_0}{\sqrt{1 - \frac{v}{c}}}
\]

(Serway RA, Vuille C, 2015). The observed frequency is equal to the source frequency multiplied by the velocity of the observer divided by the velocity of the source. These fundamentals will prove to be a great foundation for which to build upon the concept of echolocation.

Bats have lived on earth for millions of years but it wasn’t until recently in the scope of time that research was conducted on them to determine how they navigate at night. The first experiments done on bats were very crude and barbaric. Early scientists began experimenting on bats by taking away their senses one at a time. An Italian scientist by the name of Lazzaro Spallanzani began experimenting on bats in the year 1794 by sticking a hot wire in their eyes in order to blind them (Hughes HC, 1999). Being the thorough man that he was, he eventually decided to remove the eyes altogether in order to be sure of the results of his experiments. Indeed, the bats that he experimented with were able to fly just fine regardless of missing their eyeballs. Other scientists were not convinced by Spallanzani’s findings. Several unsuccessful experiments followed which resulted in some very strange and incorrect theories. Almost a century and a half later there was a breakthrough and echolocation was proven by Donald Griffin and George W. Pierce in 1938 (Hughes HC, 1999). The experiment performed by Griffin and Pierce was very simple and thankfully had no ill effects on the bats that were involved. Griffin was a student at Harvard who had the idea to conduct the experiment and Pierce was a professor that had the machinery that Griffin needed. Using a recorder and a high frequency oscillator, they were able to record the sounds of the bats and calculate the frequency of the bat’s pulses (Hughes HC, 1999). Donald Griffin is credited with identifying the sixth sense, echolocation, that allows bats to navigate and find their prey in total darkness.

When a person observes a bat flying they may think to themselves that the bat looks very erratic. Humans are unable to adequately assess the complex flight pattern that bats exhibit as they are hunting for prey in the night sky. However, when caught on video, scientists are able to
slow down footage in order to analyze the flight of many different species of bats. The reason that bats appear to be flying erratically is due to the fact that they are able to change their “flight path” at a moment’s notice. While the bat is hunting it will send out a series of chirps in order to locate its prey which can be seen in Figure 2. The signals that are sent out will come in contact with the prey and then rebound and the bat will interpret the incoming signal. These signals are sent out and received rather quickly and the bat will change its direction in accordance with the signal that it has just received. Another interesting fact about the way that bats fly is the fact that their wings are able to move asymmetrically. This makes them fantastic fliers and it is this feature that enables them to make changes so quickly. This is different in comparison to the way that birds fly. Birds are only able to fly symmetrically which means that their wings are always at the same location (think of it as a mirror, the right side will always match the left and vice versa). The nature of bats’ wings allows them to be very dynamic fliers. It is a quality that is studied and would be a major breakthrough if it was emulated with technology. The next anatomical structure that is important for microchiropteran echolocation is a bat’s ears. The size of the ears of bats varies between species. For instance, bats that emit weaker acoustic pulses have much larger ears which helps them to receive the incoming signal (Hughes HC, 1999). The size of a bats ears is a direct result of the types of frequencies that they use for echolocation. The following paragraph will explain the complex physical processes that echolocation entails.

There are three different echolocation strategies that bats employ and they are carried out by brains that weigh a modest half of a gram (Denny M, 2004). Constant frequency, frequency modulated, and a combination of the two aforementioned strategies are the three different strategies that bats use for echolocation signals (Suga N, 1990). These strategies will be abbreviated as CF, FM, and CF-FM. These three strategies each have their own unique sound. Constant frequency pulses, as one can see from the name, maintain a constant tonal frequency for the duration of each individual pulse. CF has its limitations and is not the best at getting the finer details of a target or discerning its exact location, however, long CF pulses can be used for measuring Doppler shifts and detecting objects that are longer than the wavelength of the signal (Suga N, 1990). FM pulses on the other hand are very useful for identifying fine details such as the distance of prey by computing specific echo delays after they have rebounded off of the target (Suga N, 1990). The echoes that hit the target are compounded and return with a frequency that is much higher than the original pulse (Suga N, 1990). The bat will then interpret the Doppler shift and will be able to predict the location of their target or prey. Bats mostly use echolocation for the purpose of hunting insects. There is a very specific sequence of echoes that a bat uses while it is hunting and happens upon its prey. The attack sequence of a bat is comprised of three distinct stages: detection, approach, and terminal (Denny M, 2004). The following description of an attack sequence is specifically for bats that utilize both CF and FM strategies. During the detection phase of an attack sequence the pulses that are emitted are long CF pulses that are repeated at low pulse repetition frequency (Denny M, 2004). As the bat enters the approach stage of the attack sequence the CF pulses are shorter and are transmitted at a high pulse repetition frequency (Denny M, 2004). In the terminal stage of the attack sequence the pulse repetition frequency increases even more as the bat emits short FM pulses of high bandwidth (Denny M, 2004). These pulses are so close together that it sounds like a constant buzz which is why this phase is also known as the feeding buzz (Denny M, 2004). Bats are able to manipulate many factors of the pulses that they emit. These factors are the frequency of the pulse, the pulse repetition frequency, as well as the type of acoustic signal (CF, FM, or CF-FM).
Much like the history of bats, the first studies done on bottlenose dolphins were also somewhat morbid. Bottlenose dolphins are the most studied species out of the species included in the order cetacea. In his book, *Sensory Exotica*, Howard C. Hughes mentions how the first experiments on dolphins entailed studying their severed heads (Hughes HC, 1999). Scientists would force compressed air through the larynx in order to simulate a “bark” or “whistle” (Hughes HC, 1999). However the scientists soon realized that the results from their experimentation would not be accurate due to the fact that dolphins have more control over their musculature in order to shape the pitch of the sound. Unlike bats, dolphins may have not suffered the same amount of unnecessary, inconclusive experimentation. There was great debate for some time about whether or not the acoustic impulses were made by forcing air through the larynx.

Cetacea is a taxonomic order that includes dolphins, whales, and porpoises which uses echolocation to navigate and hunt in the darkest depths of their aquatic environments. They use echolocation for social purposes as well especially since dolphins and whales travel together in pods that can reach up to 1,000 members at any given time. A vast majority of cetacea survive and thrive in the ocean and a select few live in freshwater rivers. Living in an aquatic environment rather than land has an inevitable impact on the way that they use echolocation. There are many similarities in the ways that bats and cetacea utilize echolocation. The frequency of the pulses they use are both so high that they are outside of the range of what humans are able to observe without the aid of technology. Some of the more obvious similarities between cetacean and microchiropteran echolocation is the hunting and navigation purposes. One of the main differences is that dolphins and killer whales are highly intelligent creatures that use echolocation as a means to communicate with their pods. Another distinction of cetacean echolocation is that their prey is not airborne like it is for a bat. This section will mainly focus on aspects of echolocation that are unique to cetacea and how it has affected their anatomical morphology.

To begin, one of the differences between cetacea and bats is that cetacea do not have external ears. The reason for this is having external structures, such as the pinna, would decrease the speed at which the mammal would be able to travel (Fulton J, 2015). External ears would be a disadvantage in situations where the dolphin or whale was racing to catch their prey and would significantly hinder their pursuit. Instead, dolphins and whales have developed a different mechanism for “collecting” incoming frequencies that allows them to remain hydrodynamic in their environment rather than causing a hindrance. In fact, the method for which they propagate outgoing signals differs from bats as well. Figure 3 is diagram of the anatomy of a dolphin’s echolocating system which depicts some of the most important features that allow them to produce, collect, and interpret acoustic pulses: the mandible, melon, and nasal sacs (Zinner, 2009). The outgoing signal propagated by cetacea is dispelled from nasal sacs located just below the blowhole (Zinner, 2009). Air is forced through the nasal sacs and is then focused with a structure in their forehead that is called the melon (Zinner, 2009). According to Harper et al, “The melon is a lipid-rich structure located in the forehead of odontocetes that functions to propagate echolocation sounds into the surrounding aquatic environment” (Harper CJ, 2008). The melon looks and acts very similarly to the way that the lens works in an eye. When light passes through the lens of an eye, the light is refracted so that it focuses on the macula of the retina. When air is squeezed through the nasal sacs it is then focused by the melon, which can be thought of as an acoustic lens, and a pulse is transmitted in the desired direction (Zinner, 2009).
Therefore, the first difference is that the pulse signal is transmitted via the melon rather than the mouth of the dolphin. Once the impulse rebounds off the target, it will return to the dolphin or whale and rather than being received by ears like it would for a bat, the signal will be received through the jaw. Nerve impulses from the vibrations will then be transmitted to the brain via neurons and an image will be perceived by the whale or dolphin. Dolphins have been known to reach sonar calls that are as much as 220 dB in open water (Hughes HC, 1999). In closed tanks they emit signals that are 170 dB so as not to cause reverberation from much stronger signals (Hughes HC, 1999). This is an example of how the dolphin can easily adjust the intensity of its signals in order to adapt to its environment and create an optimal echo. Unfortunately, the research available for cetacean echolocation is not as extensive as that of microchiroptera. As an order, dolphins, whales, and porpoises are much, much larger than bats. This makes it significantly more challenging to research them and collect data about their echolocating process. It is far easier to keep bats which can fit in the palm of your hand compared to aquatic mammals that weigh thousands of pounds. This combined with the fact that echolocation was identified in 1938 just means that this field of research is in its infancy. There is much more to be learned about echolocation that will inevitably be revealed as more research is conducted.

Conclusion:

One of the truly spectacular things about the research of biosonar is that this area of research is relatively new, paving the way for more discoveries to be made. Microchiroptera and cetacea have used biosonar for millions of years and as humans we have only been aware of its existence for less than a century. We have so much more to learn from these adaptive creatures that inhabit the sea and sky. The greatest inventions are the ones that mimic nature and sonar is no exception to that rule. Biosonar and echolocation will undoubtedly continue to be researched and that is for two main reasons. The first reason is that the sonar systems of these mammals are so advanced that the ability to mimic them would be of great importance to engineers who specialize in developing sonar systems. The implications of improved sonar would certainly affect military operations, etc. Sonar is used by submarines for underwater navigation, detection of other objects, and communication. Besides its militaristic benefits, sonar is used in applications such as ultrasound which is invaluable to medical professionals and the patients that benefit from its use. The second reason that biosonar will continue to be researched is human curiosity. A scientist can spend the entire span of their career researching the most miniscule detail in their specific field. There are so many factors involved in biosonar that there are surely more discoveries to be made. The biosonar techniques of microchiroptera have been researched a great deal more than those of cetacea. It will be exciting to see what new information can be found from further research on cetacean biosonar techniques. Thankfully, there is already a great deal of information that has been gleaned from years of research. I have provided a summary below of the information that is available on echolocation that was discussed in this paper.

The main points of this paper have been to investigate the basic physical properties of sound, the history of echolocation, and the similarities and differences of how cetacea and microchiroptera utilize echolocation. The journey to understand how bats “see at night” began in 1794 with the grotesque experiments performed by the Italian scientist Spallanzani (Hughes HC, 1999). Only 144 years have passed since Donald Griffin proved that bats had a sixth sense that he named echolocation (Hughes HC, 1999). Since then, Donald Griffin and many other scientists have devoted their lives to studying the echolocation techniques of microchiroptera and cetacea and have graciously shared their knowledge with the world. This paper offers just a glimpse of
the complex processes that are involved in echolocation. From my research I have learned that microchiroptera and cetacea each use high frequency pulses that are above the range of what humans can hear. This research has shown me that bats and cetacea have different methods of emitting and receiving the outgoing and incoming acoustic pulses. But in the end, they both use echolocation for hunting and navigational purposes.
Figures

Figure 1
A longitudinal wave. This is a visual representation of how a longitudinal wave propagates. The areas of compression travel in the positive x direction when the acoustic pulse is initially sent out. The returning echo moves in the negative x direction after it has bounced off prey. (GCSE, 2015)

Figure 2
A simple representation of a bat sending out a pulse and the incoming signal as it rebounds off of a moth. (KAC, 2015)

Figure 3
This a representation of the anatomy of a dolphin’s head. The figure shows a hypothetical situation in which a dolphin has emitted a pulse and is receiving the pulse after it has bounced off of its target. The returning echo is received through the dolphin’s mandible. The horizontal dashed lines are indicative of a longitudinal wave. (Zinner, 2009)
References:


Figures References:


Aluminum Air Batteries in Times of Emergency

Jessica Flores
November 17, 2016

PHY112 General Physics II
Dr. Casey Durandet
ABSTRACT
This research aimed to test if the aluminum salt water battery experiment could be used to power a cellular device. Most portable devices require 3.7 to 5.0 volts to recharge the internal battery. In this experiment, 9 battery packs were connected in series, with voltages varying from 0.68 V to 0.78 V, to create a total voltage of 5.8 V. The voltage obtained from the system proved enough to power a small speaker (minimum voltage of 3.7 V) but was not successful in powering a cellular device needing a minimum voltage of 3.85 V. Therefore, it was concluded that a minimum current must also be achieved to charge a cellular device.

RESEARCH
In times of crisis, when power may not be immediately available, building a power supply can provide electricity and communication to the outside world. Typical forms of off grid power include mechanical generators, windmills, solar panels, and batteries (U.S. Dept. of Energy). If a household does not contain any of these off-grid sources, a basic aluminum air salt water battery can be made using common household supplies (Tamez and Yu 2007). Sources that have replicated the aluminum air salt water battery experiment have gone as far to suggest that this form of battery would be sufficient to charge a mobile device (Norton and Hand 2013). This research aimed to determine if a mobile phone could not only receive power through a series of batteries, but also receive charge for long enough to be a feasible emergency battery.

To replicate the experiment for creating the aluminum air salt water battery, all the same supplies were used, as seen in the video, “DIY Salt Battery Urine Can Charge Your Smart Phone!” by Patrick Norton and Michael Hand (Norton and Hand 2013). This included Reynolds Wrap aluminum foil, API activated filter carbon, salt water made with iodized salt, paper towels, copper wire, and alligator clips. A diagram and photo of the battery pack composition can be seen in Figures 1 and 2a. At the aluminum anode, pockets of oxygen contained in the activated carbon oxidize the aluminum foil. Copper acts as the reducing agent, or cathode, supplying electrons to reduce the oxygen, forming hydroxide. A salt-water solution, contained in the dampened paper towel, was used as the electrolyte to transport ions between the copper and aluminum. The reactions (from Tamez and Yu 2007) are as follows:

\[ \text{Anode: } Al(s) + 3OH^-(aq) \rightarrow Al(OH)_3(s) + 3e^- \]

\[ \text{Cathode: } O_2(g) + 2H_2O(l) + 4e^- \rightarrow 4OH^-(aq) \]

\[ \text{Overall: } 4Al(s) + 3O_2(g) + 6H_2O(l) \rightarrow 4Al(OH)_3(s) \]

At the end of the video, 7 battery packs were connected in series to provide 4.3-4.9 volts, which meets the threshold voltage requirement of most Android smartphones. Using a spliced USB converter cord, a dated HTC smart phone is connected to the series of batteries and the “charging” icon on the phone appears. This concluded the experiment for the video without any indication of the current achieved through the cells or how long the current could be sustained.

In this research, three devices are used to determine the feasibility of the battery packs. The first device is a portable iHome speaker, the second is an LG Stylo 2, and the third is an iPhone 4. Two variables need to be considered when creating a battery to charge an internal rechargeable
battery; voltage and ampere hours. The voltage listed on the rechargeable battery gives a suggestion of the maximum input voltage that can be supplied so that the battery does not burn out. For the iHome speaker, LG Stylo 2, and iPhone 4, the voltages were given as 3.70 V, 3.85 V, and 3.7 V respectively. Although each salt water battery pack is limited to around 1 V per pack, multiple packs can be connected in series to achieve a higher voltage. The total voltage can be calculated as the summation of all voltages, while connected in series:

$$V_{total} = \sum V_i$$

If connected in parallel, the total voltage could be calculated as the sum of the inverses of all voltages:

$$V_{total} = \left(\sum \frac{1}{V_i}\right)^{-1} = \left(\frac{1}{V_1} + \frac{1}{V_2} + \frac{1}{V_3} + \cdots\right)^{-1}$$

Connecting two identical batteries in parallel will result in a voltage that is half of one of the batteries voltage measurements. However, this type of connection does result in a higher current which leads to the next critical variable: ampere-hours. Ampere-hours describes how many amps the device can consume within an hour. To find the time it would take to charge a device, the rating is divided by the input current, which is measured in amperes.

$$\text{Charge Time} = \frac{\text{Ampere - Hours}}{\text{Input Amperes}}$$

Therefore, if a device is rated at 1000mAh and the input current is 1000mA, it would take 1 hour to charge the device. One advantage of testing the iHome speaker is the low rate of 500mAh compared to the 3000mAh rate for the LG phone’s lithium ion battery and the 1420mAh for the iPhone’s lithium polymer battery. If the homemade battery produces an ideal output current, close to the 1000mA current seen in most wall chargers, it would be a sufficient source of power in the case of an emergency. In this research, the voltage and current are measured over many hours to test the capacity of the battery and the endurance of the reactions.

I. Procedures

Reynolds Wrap aluminum foil was used for the anode and placed on the bottom of the pack. A paper towel was dampened with salt-water solution, which acted as an electrolyte to allow ions to move freely. To make the salt-water solution, one and a half cups of water was microwaved to 120 degrees Fahrenheit before dissolving two tablespoons of salt. The 28.5 cm x 15.5 cm dampened paper towel was placed on top of the 30 cm x 20 cm aluminum foil and provides insulation from the copper wire placed on top. A fourth of a cup of activated carbon is placed on top of the dampened paper towel and copper wire as shown in Figure 2. The activated carbon was moistened by pouring an eighth of a cup of the salt-water solution on top. After achieving the stack shown in Figures 1 and 2a, the paper towel was folded over the carbon and copper wire to cover completely from the aluminum, which was folded over all the materials. The finished battery packs are shown in Figure 2b.
Alligator clip leads were used to connect the circuit together, one side clipped to the copper wire and the other to the aluminum foil. Battery packs were connected in series by attaching aluminum ends to copper ends (Figure 2b). Each battery pack was individually measured for voltage and current using a multimeter before being connected in series (Table 1). All measurements were repeated after a period and then averaged to create trend lines relating voltage and current to time. These results are shown in Figures 4 a-c.

To test if the series of batteries could charge a phone, a USB converter cable was stripped to reveal four colored wires. The red and white wires represented the positive ends and the green and black wires represented the negative ends (Chris Porter 2016). Each wire set was intertwined (Figure 3) to connect to the alligator clips. On one side, the positive ends connected to the copper cathode and on the other, the negative ends connected to the aluminum anode. A typical USB to micro USB cord was used to connect the speaker and mobile phones to the circuit to determine if the home-made battery carried enough voltage and current to charge the device.

II. Results
During the first trial, measurements for voltage on the new battery packs were all within a standard deviation of 0.028 V with an average of 0.73 V. This was significantly lower than the anticipated 1 V value although the current was well over the 100 mA expectation, averaging a value of 132 mA. The standard deviation can be calculated as

$$SD = \sqrt{\frac{\sum(x - \bar{x})^2}{n}}$$

where x is a data point, \(\bar{x}\) is the average, and n is the number of data points (Khan). For current, this deviation was as high as 22.2mA. In this experiment, current was seen to fluctuate more violently than voltage. Current is directly dependent on the transport of electrons throughout the system and can be modeled by the equation

$$I = \frac{\Delta Q}{\Delta t}$$

where Q is the total charge of all electrons flowing through the system and t is the time interval in which the current is measured (Serway and Vuille 2015). When measuring current on the multimeter, this equation becomes apparent with a continuously increasing or decreasing value whereas voltage levels out almost immediately to the measured value. For this reason, a large margin of error in recorded current values existed over the range of possible currents, sometimes as large as 20 mA. This rate of changing current can explain the speed of the reaction in short term and the degradation of the reactant materials in the long term. Over a small period, the amount of oxygen provided by the activated carbon limits the rate of reaction and therefore, the current is also limited (S. Chasteen, D, Chasteen, and Doherty 2008). Possible solutions for maximizing the current include crushing the activated carbon and coiling the copper wire to allow for more surface area, thereby increasing the absorption sites at which the reactions can occur (PEN wiki 2012).
Long term current was also a key variable in this experiment. Although the reaction is limited by the oxygen available and can be maximized by increasing the surface area for reaction sites, the original surface area of the metals can become polluted with byproducts over time, creating a barrier for the reaction to occur (S. Chasteen, D. Chasteen, and Doherty 2008). Observing the current averages in Table 2 confirms that current does decrease over long periods of time. In Figure 4b, these results have been graphed with a best-fit correlation. This trend reveals that the surface area of the reactions sites is most likely experiencing exponential decay, with the function given by

\[ I = I_0 e^{-rt} = 105.83e^{-0.002t} \]

where \( I_0 \) is the original current when the battery was first created and \( r \) is the rate of decay.

Figure 4a shows that the voltage measurements could not be clearly defined by a function. In general, voltage measurements should not be varying over time since the value is dependent on the metals (S. Chasteen, D. Chasteen, and Doherty 2008). For the reaction observed in this experiment, the electronegativity of aluminum being oxidized is \( E^0 = 2.30V \) (Park 1985).

When connecting the 9 battery packs in series, similar trends were measured over time for both voltage and current (Figures 5a and 5b). As described in the introduction, the total voltage can be calculated by adding the individual battery voltages together. For the first trial, the total voltage should have been 6.54 V, rather than the measured 5.8 V. One possibility for this error could arise from measurement errors throughout the 9 battery packs but a more probable cause would be the internal resistance provided by the 10 sets of alligator clips. All trials measured voltage drops ranging from 0.28 to 0.74 V.

This research aimed to create a series of batteries that could charge a cell phone. The measurements for the first trial suggested that there was enough voltage (needed ~3.7-3.85 V), and enough current (153mA) to charge the iPhone within 9 hours and the LG Stylo within 19.6 hours. Neither phone showed any indication that a battery source was connected. When connecting the portable iHome speaker (rated at 500mAh), the device reacted immediately. Using all 9 battery packs was too intense for the speaker. Therefore, 3 trials were conducted to see how the device would react. In the first trial, Figure 6a, 5 batteries with a total voltage of 3.83 V were connected in series to the speaker resulting in the normal intensity of blue light, indicating that the device is charging. For the second and third trial (Figures 6b and 6c), the series circuit was connected to the speaker to determine if it would still operate at the high voltages of 5 V and 5.6 V or if there would be an internal protection mechanism to prevent the speaker from burning out. In both trials the speaker did indicate that it was charging but the intensity of the LED indicator increased dramatically as a bright blue at 5 V and a bright purple at 5.6 V. Both tests were conducted for less than 30 seconds to avoid burning out the speaker’s battery.

After viewing the results of the speaker test, the iPhone and LG Stylo were both tested again at voltages near their internal battery voltage to determine if they source was being blocked to prevent an internal battery malfunction. For all tests, the iPhone and the LG Stylo would not charge. The most plausible causes for this situation could be the current is too low for the phone.
to recognize or there could be an internal switch mechanism for compatible chargers to prevent unsafe power sources from damaging the phone.

III. Conclusion
After obtaining negative results for charging a cell phone, it is hypothesized that the gentlemen in the video replicating the aluminum air salt water battery experiment either a) took an additional step in their procedures to increase the current without noting that required step in the video, b) tested a phone that was made before power protection mechanisms were installed, or c) forged their results to get a more entertaining outcome. Possible ways they could have achieved a higher include using a thicker copper wire, adding more loops to their copper wire, or using a copper mesh to provide maximum surface area on the copper. Grinding the activated carbon to a finer grain rather than the rocky substance from the box would also increase surface area for a higher current. However, none of these steps were noted in the video which leads to an issue of experiment reproducibility. Since this experiment is not directly reproducible, the results cannot be verified leading to the conclusion that a cell phone cannot be charged connecting multiple aluminum air salt water battery packs in series.

Another question that arises if the proper current can be achieved, either through increased chemical reactants or by connecting multiple battery packs in parallel and series, is if it would be a feasible energy source over a long period. As seen in Table 3, or roughly calculated using the exponential decay function, the current in the series battery circuit drops to half after about 3 hours and has maintained only a fourth of the current after 12 hours. Therefore, the time it would take for a 250mA battery to charge a phone rated at 3000mAh could turn from 12 hours to almost a month for a full charge, after the decay of the reactants. To maintain an adequate, the reactants would have to be constantly replenished which may not be possible given the limited resources of an emergency.

This experiment serves as a great resource for learning about the properties of batteries and connecting physics and chemistry concepts together. However, in case of an emergency, taking the time and resources to create an aluminum air salt water battery is not a feasible solution for charging a mobile device.
Figure 1: Diagram of battery pack components. Copper serves as the cathode and the aluminum foil is the anode.

Figure 2a (left) Aluminum foil, napkin, activated carbon and copper wire. Figure 2b (right) Battery packs connected in series.

Figure 3
Top Image: USB-A Wiring Diagram.
http://wiringdiagram.dvlgservices.com/usb-a-wiring-diagram/

Left Image: Picture of spliced USB-A cord.
### Table 1: Voltage and current of all battery packs measured over 5 trials.

<table>
<thead>
<tr>
<th>Pack #</th>
<th>Trial 1</th>
<th>Trial 2</th>
<th>Trial 3</th>
<th>Trial 4</th>
<th>Trial 5</th>
</tr>
</thead>
<tbody>
<tr>
<td></td>
<td>time (min)</td>
<td>Voltage (V)</td>
<td>Current (mA)</td>
<td>time (min)</td>
<td>Voltage (V)</td>
</tr>
<tr>
<td>1</td>
<td>0</td>
<td>0.74</td>
<td>145</td>
<td>86</td>
<td>0.7</td>
</tr>
<tr>
<td>2</td>
<td>0</td>
<td>0.72</td>
<td>138</td>
<td>78</td>
<td>0.65</td>
</tr>
<tr>
<td>3</td>
<td>0</td>
<td>0.72</td>
<td>120</td>
<td>95</td>
<td>0.5</td>
</tr>
<tr>
<td>4</td>
<td>0</td>
<td>0.71</td>
<td>148</td>
<td>82</td>
<td>0.65</td>
</tr>
<tr>
<td>5</td>
<td>0</td>
<td>0.73</td>
<td>139</td>
<td>69</td>
<td>0.5</td>
</tr>
<tr>
<td>6</td>
<td>0</td>
<td>0.75</td>
<td>138</td>
<td>78</td>
<td>0.65</td>
</tr>
<tr>
<td>7</td>
<td>0</td>
<td>0.74</td>
<td>143</td>
<td>75</td>
<td>0.5</td>
</tr>
<tr>
<td>8</td>
<td>0</td>
<td>0.72</td>
<td>154</td>
<td>71</td>
<td>0.65</td>
</tr>
<tr>
<td>9</td>
<td>0</td>
<td>0.75</td>
<td>139</td>
<td>69</td>
<td>0.5</td>
</tr>
</tbody>
</table>

Averages: 0.73 V, 132.11 mA, 83.56 V, 68.00 mA, 169.22 V, 521.11 mA, 735.22 V, 34.11 mA, 720.72 V, 35.72 mA

### Table 2: Average measured voltage and current for battery packs with calculated resistances for selected times

<table>
<thead>
<tr>
<th>Time (s)</th>
<th>Voltage (V)</th>
<th>Current (mA)</th>
<th>Resistance (ohms)</th>
</tr>
</thead>
<tbody>
<tr>
<td>0</td>
<td>0.73</td>
<td>132</td>
<td>5.53</td>
</tr>
<tr>
<td>84</td>
<td>0.67</td>
<td>84</td>
<td>7.98</td>
</tr>
<tr>
<td>169</td>
<td>0.53</td>
<td>55</td>
<td>9.64</td>
</tr>
<tr>
<td>521</td>
<td>0.53</td>
<td>34</td>
<td>15.59</td>
</tr>
<tr>
<td>735</td>
<td>0.60</td>
<td>20</td>
<td>30.00</td>
</tr>
</tbody>
</table>

### Figure 4 a) Voltage; b) Current; c) Resistance
Table 3: Measured voltage and current for battery packs connected in series with calculated resistances for selected times

<table>
<thead>
<tr>
<th>Time (s)</th>
<th>Voltage (V)</th>
<th>Current (mA)</th>
<th>Resistance (ohms)</th>
</tr>
</thead>
<tbody>
<tr>
<td>0</td>
<td>5.80</td>
<td>153</td>
<td>37.9</td>
</tr>
<tr>
<td>84</td>
<td>5.29</td>
<td>130</td>
<td>40.7</td>
</tr>
<tr>
<td>169</td>
<td>4.51</td>
<td>77</td>
<td>58.6</td>
</tr>
<tr>
<td>521</td>
<td>4.26</td>
<td>43</td>
<td>99.1</td>
</tr>
<tr>
<td>735</td>
<td>5.03</td>
<td>35</td>
<td>143.7</td>
</tr>
</tbody>
</table>

Figure 5 a) Voltage; b) Current; c) Resistance
Figure 6
a) Top Left: 5 batteries in series, soft blue light
b) Top Right: 7 batteries in series, bright blue light
c) Bottom Right: 8 batteries in series, bright purple light, maximum “safe” voltage
REFERENCES

(2) S.V. Chasteen, D. Chasteen, P. Doherty; “The Salty Science of the Aluminum-Air Battery”; The Exploratorium; San Francisco, CA; Vol. 46; December 2008

(3) Serway, R.A. Vuille, C.; College Physics; Tenth Edition; Cengage Learning; 2015


(9) S.M. Park; “Boron, Aluminum and Scandium”; Standard Potentials in Aqueous Solutions; IUPAC; Marcel Dekker Inc.; New York; 1985
The Relationship Between Exercise and Diabetes

Logan Franklin
BIO181
Dr. Darra Browning
Abstract

Exercise is an important piece in maintaining a healthy life, however, for many people, it is not as simple to exercise to stay healthy. Those who have diabetes or suffer from hypoglycemia, must always be vigilant when it comes to their blood sugar levels. Exercise reduces blood glucose levels, which creates a predicament for many diabetics, how are they supposed to exercise safely? In this study, methods of safe exercise for diabetics will be discussed as well as the details of why exercise causes hypoglycemia.

Introduction

As everyone knows, exercise is an important part of maintaining a healthy lifestyle, however, what happens when exercise conflicts with a disease? For many people diagnosed with diabetes, it can be difficult to try and balance good blood glucose numbers and proper amounts of exercise. Exercise typically lowers blood glucose numbers and in some cases the drop in that can cause a bout of hyperglycemia or low blood sugar, which, if not handled correctly, can lead to problems such as fainting or feeling very weak. The question that is commonly asked is, how much exercise is too much and how do I exercise while maintaining a good blood glucose?

I. What Happens When a Diabetic Exercises?

One of the key steps in understanding how exercise affects diabetics is knowing how physical activity lowers blood sugar levels and why. According to the American Diabetes Association, exercise decreases blood glucose levels because when muscles contract, they also use up glucose for all of the processes involved and the glucose is acquired from the blood. ADA also acknowledges that exercise heightens insulin sensitivity thus making the insulin more proficient at taking up glucose. What all of that means is, if you’re consistent with the exercise, the ability to use up glucose is increased. In regards to diabetics, this is the main issue revolving around exercise, while the above process would not present a problem for non-diabetics, it does especially for those with Type 1 diabetes. Type 1 diabetes entails that the pancreas was damaged to the point that it does not produce a sufficient amount of insulin to function properly, thus rendering the aforementioned results of exercise a threat. If the body of a diabetic can not produce enough insulin to make use of the glucose, then when muscles contract during exercise and require glucose to be taken from the blood, more of the glucose is used up inefficiently thus lowering the blood glucose levels of diabetics and causing a danger during exercise. Due to the rapid way in which blood sugar levels can decrease, it is pertinent to maintain a watchful eye on the blood glucose levels through out the exercises.

II. What is Hypoglycemia and why is it dangerous?

There is not a certain way to measure how well a diabetic is handling themselves on a daily basis, however, a 3 month average, commonly known as an A1c number is the standard measurement that is currently used. For adults, the desired A1c numbers are 6-7. The lower the A1c number the tighter the line they walk between good blood glucose levels.
numbers and low blood sugar levels. Due to the fact that there is a very small area of forgiveness in regards to excellent blood glucose levels and low ones, it is even more difficult to walk that line thus making the aspect of exercising as well, even more frightening. For these reasons, there has been a recent push for more knowledge of hypoglycemia and more practice with different methods of managing it. On a patient level, endocrinologists and nutritionists have began not only lecturing on diabetes related things, but more specifically on managing low blood sugars. Diabetics often relate hypoglycemic numbers with feeling weak or irritable and simply treat it with a quick bite to eat or drink of juice to bring their numbers up and then be done with it, but that is not the proper way to treat it. Doctors are now advising to slowly raise the numbers with smaller increments rather than a quick heavy dose of carbohydrates, this is due to the fact that while the sudden rush of carbohydrates will increase the blood glucose levels, it may not have a lasting effect and the numbers can drop back down later on. Snacks such as a slice of bread or anything simple will have a much more filling effect and thus not have the risk of a quick spike and sharp decline in blood glucose levels. That method works for less serious situations, in scenarios where the blood glucose is extremely low, i.e anything less than 60, a glass of juice or candy is more often the way to go(Kreider, Padilla, Pereira, 2017).

**How this relates-** Low blood sugar levels are the main obstacle for diabetics when it comes to exercising and therefore it is important to know how to handle near hypoglycemic numbers and preferably treating them before they occur. Knowing when your blood glucose is dropping and understanding the feeling will help prevent major hypoglycemic event while exercising.

### III. Why doesn’t a Diabetic’s body combat hypoglycemia?

For most people exercising isn’t always a cause for concern and that it because their body’s regulate their blood sugars at all times however, this is not the case for diabetics. Blood glucose levels are controlled by the amount of insulin that is present but during exercise if blood glucose levels dip too low, our body will usually inject some substances to artificially raise our blood glucose to bring us back to homeostasis. One of the common substances used by the body to raise blood sugars is glucagon. One of the reasons that hypoglycemia can be so difficult to control especially when involving exercise was discussed in a study conducted by Vanessa Jones Briscoe, Donna Bowman, Tate and Stephen Neil Davis. In the study, they found that measures used by the body to raise blood sugar, such as the use of glucagon, are hampered by hypoglycemia and more specifically, antecedent hypoglycemia. What that means is that if a diabetic’s blood glucose level is even slightly below the norm thus beginning the onset of hypoglycemia, that it will hinder the body’s efforts to raise its’ own blood sugar level. As someone exercises their blood sugar will typically drop some, but with a diabetic that is possibly hypoglycemic even if slightly, it creates a compounding effect thus making it more difficult to raise their blood glucose level.

**How this relates-** The major risk with diabetics and exercising is hypoglycemia and the risks it presents. While the simple solution would be to grab some food and then be on your way, it is not so simple for diabetics. As stated above, without the proper
regulation that a properly functioning pancreas would provide, the body of a diabetic essentially works against itself thus further complicating an already difficult process of regulating their blood sugar levels.

IV. Prevention of Hypoglycemia

In an attempt to discover a way to prevent the onset of hypoglycemia a study was undertaken at a two week sports camp involving twenty five subjects ranging in age from 8 years old, to 17 years old. The goal of this study was to test a product called a continuous glucose monitor (Riddell, Milliken 2011). The GCM has a programmed algorithm that worked to find a correct way to manage carbohydrates with the exercise in order to prevent hypoglycemia. While the subjects had the GCM with them, they also had to consume smaller amounts of carbohydrates in junction with the GCM. With this method, hypoglycemic incidents occurred about 38% of the time (Riddell, Milliken 2011). With that stat in mind, the architects of this experiment were able to claim that their algorithm along with the combination of a small intake of carbohydrates through out the exercise, was an efficient way of protecting against hypoglycemic events.

How this relates- This study was an important piece of information because not only did it find a successful method for reducing the frequency and severity of hypoglycemic attacks, it provided a footnote for many diabetics to follow themselves. Another thing that this study proved was that the diabetics themselves must always remain vigilant of their own blood glucose numbers in order for any hypoglycemic prevention to work properly.

V. What types of exercises are recommended for diabetics?

Despite all of the risks involved with exercising, it is actually one of the most beneficial routines a diabetic can undertake, however, as with everything, there needs to be a degree of restraint and caution in which exercises to do, and how intensely to do them. According to an article written by Sue Cotey and Andrea Harris, both registered nurses, the best exercises for diabetics are- walking, tai chi, yoga, dancing and swimming. They go on to list the reasons why these exercises are particularly good for diabetics; walking is one of the best exercises because it can be done easily and anywhere, it is also a simple way to slowly increase your exercise without too much strain. Harris and Cotey state that the benefits of tai chi are shown by a study done in Florida where 62 Korean women were split into a control group and a group that practiced tai chi. The group that took the tai chi lessons reported significant improvements in blood sugar control. This is due to the fact that tai chi is not an intense workout but steady body movements. In the same article, the reason yoga is healthy for diabetics is stated as, “It lowers stress and improves nerve function, which leads to an increased state of mental health and wellness” (Cotey, Harris, 2017). Dancing is another simple yet effective method for improving blood glucose control and nerve pain, not only is it fun and simple but the amount of calories burned, which helps with any disease, is the contributing reason as to why dancing is helpful exercise. The final exercise in the list is swimming, swimming is a great exercise because the stress on the
muscles is not as focused due to being in the water. In addition to the low stress, but lowering cholesterol numbers as well as blood sugar numbers and the physical workout required contribute to the overall health of anyone, but especially diabetics.

**How this relates** Knowing which exercises are the most beneficial is critical to combating hypoglycemia while maintaining a healthy lifestyle. Also understanding why they’re helpful is key into constructing a workout regime all together.

VI. **What to eat before and after exercising**
Consuming some carbs before and after exercising is recommended for everyone, however for diabetics it is almost critical to maintaining a safe blood sugar level. Constructing a proper snack or small meal may seem difficult but in reality there are only a few components necessary, according to an article by Diabetes Forecast, the essentials for this pre workout meal include whole grain snack bars, apples, peaches, dates, figs, milk, and yogurt. The article goes on to detail that a small yet efficient mixture of complex and simple carbohydrates along with a small dose of proteins should be enough to sustain a workout. Before beginning the workout it is also important to understand how your own body reacts to certain foods and workouts, for example, a simple walk around the neighborhood would not require a meal of any sort and eating too much could also cause a high blood sugar level which would defeat the purpose of exercising(Moore 2015).

**How this relates** Being smart about preparing for exercise is the first step in exercising smartly. Eating anything may provide a temporary boost to blood glucose levels but not be the most efficient way of going about it. Efficiency in diet will allow for more exercise and less worries about having a hypoglycemic event.

VII. **Aerobic vs. Anaerobic**
Both Aerobic and anaerobic exercise are important for diabetics but the main question is, “how much of each method should I do?”. Luckily, the experts have devised a general guideline for which to follow, “Aiming for 30 minutes of moderate-to-vigorous intensity aerobic exercise at least 5 days a week or a total of 150 minutes per week. Spread your activity out over at least 3 days during the week and try not to go more than 2 days in a row without exercising”(American Diabetes Association). While that is the optimal amount of aerobic exercise, it is very likely that most diabetics wont be able to reach those numbers so getting as close as possible is what is asked. As for anaerobic exercise, the ADA recommends, “doing some type of strength training at least 2 times per week in addition to aerobic activity”(American Diabetes Association). Finding the time to perform these exercises is a key component to living healthily. In contrast to popular beliefs, simply running every now and then is not quite enough exercise for most but in particular diabetics.

**How this relates**
Understanding how to break up the types of exercise can save a lot of hassle and useless struggles. It will allow for a more efficient workout regime and a steady guideline of which to follow. Having a set pattern of working out will allow for an accompanying
set diet catered to the individual. A set diet will allow for easy adjustments to be made and a steady lifestyle, which is the goal for any diabetic.

VIII. An Increasing problem?
While it is evident that Diabetes is a major disease, it is not always clear how vast its’ reach is. While Diabetes is not typically a direct cause of death, the effects that can be suffered from hypoglycemia and hyperglycemia (high blood sugar) directly link towards other diseases that are more commonly linked with death. In a way, Diabetes is somewhat of an unsung disease, due to the fact that it is not generally counted amongst the most dangerous. To truly understand the full scope of diabetes it is necessary to zoom out. Diabetes has always been present but it is becoming more and more common, as the CDC notes “from 1980 through 2014, the number of Americans with diagnosed diabetes has increased fourfold (from 5.5 million to 22.0 million)”. The amount of increased cases has thrust both Type 1 and Type 2 diabetes into the limelight.

How this relates
With Diabetes now getting more and more acknowledgement for its dangers, nutrition and exercise are becoming daily factors in the health of diabetics. Now, the focuses aren’t only on prevention of diabetes, but also on how to control it properly and narrow its side effects. It isn’t just type 1 Diabetes that is on the rise, but also type 2. Typically, type 2 Diabetes occurs due to poor health and poor diet, thus studies on relations between controlling blood sugars and diet and exercise are becoming more common.

IX. Conclusion
Living life as a diabetic can seem like a complicated issue to deal with, especially on a day to day basis but the key is knowledge. Understanding how each part of exercise, varying types of food and how they affect the body will allow for a less stressful life. Exercising might seem like a chore but the benefits it brings, especially to diabetics outweigh any negatives and while hypoglycemia appears and feels dangerous and can be very alarming, there are ways of which it can be somewhat controlled through certain methods derived from research and understanding. The idea of exercising and hypoglycemia coexisting is only possible through vigilance as diabetics are always walking a fine line between being healthy and having a hypoglycemic event.
Figures

Figure for paragraph 1
(MedMovie.com)

Figure for paragraph 2

Figure for paragraph 5
(Deadspin.com)

Figure for Paragraph 6
(Muscle and Strength.com)
Figure for paragraph 8
(CDC)

Figure for paragraph 7
(San Diego Health and Wellness)
References

5. 5 Best Exercises for People with Diabetes [Internet]; c2014. Available from: https://health.clevelandclinic.org/2014/06/5-best-exercises-for-people-with-diabetes/.
7. Number(s in Millions) of Civilian, Non-Institutionalized Persons with Diagnosed Diabetes, United States, 1980-2014 [Internet]; c2015 [cited 2017 April 10th.]
The Confocal Microscopy and Wide-field Microscopy

Shima Golshan
11/17/2016
PHY112 – 26104/26171 (Lecture/Lab)
Dr. Casey Durandet
**Abstract:**
All the scientists need to have knowledge about the microscopes that would help them in their fields of their studies. This paper would acknowledge them with very basic information about the Confocal Microscopy and Wide-field microscopy. First, a brief introduction about the confocal microscopy, critical aspects of confocal Microscopy, and the basic parts of the confocal microscope. This would proceed with more information about the imaging model for both Confocal and Wide-field microscopy. In addition to that, it will be more specifically explained more about how the electronic imaging has been done with the presses. It will follow by more information about fluorescence microscopy which is used for study fixed and living cells because of its versatility, specificity, and high sensitivity. And the usage of each type of microscopes. This can be for studies, teaching, treatments, and experimental use. Also, the most important part of experimenting these types of microscopes the orientation of pictures and other software that could be very helpful in clarification of the picture. The following information can be helpful for scientific students, researchers, and those who like to know more about Confocal and Wide-field microscopy. The explanations and visual aids are being applied with pictures, figures, and formulas.

**Introduction:**
The Confocal Microscope is a type of microscope that is being developed for obtaining high-resolution images and 3D reconstruction. The images from the confocal; the new advantage of creating an electronic 3D image that can be rotated and selectively sectioned to better understand underlying structural and physiological mechanisms in the preparation. Confocal microscopy has good impact values in visualizing the objects in life science; one of the top values is to show the fine image possible in 3D. Confocal microscopy has few advantages in different ways compared to Wide-field optical microscopy. One of these advantages is to show the depth of the field elimination or reduction of background information away from the focal plane, and the capacity to collect serial optical sections from thick specimens [1]. The most significant value about the confocal microscope is the spatial filtering techniques that terminate the glare and focus the substance into the view. One of the other advantages of Confocal microscopy and Wide-field microscopy; the confocal type of microscopy has the easiest way to rich the extremely high-quality images. This is based on the recent years' popularity of using confocal microscopes in the science field for growing cells in biology that rely on imaging both fixed and living cells and tissues. However, each type of the microscopes is valuable to be used in different types of research experiments. The high and most significant value about Confocal is the 3D imaging in fluorescence microscopy. Due to experimental errors; it is often very difficult to eliminate artifacts in specimen preparation, to control the experimental variables, and to minima configuration problems with the microscope. In an article that Dr. James B. Pawley has written about the most common extraneous factors that often serve to obscure results collected in fluorescence wide-field and confocal microscopy, he talked about a variety of topics from immersion oil to quantum efficiency, and the specimen embedding medium. The confocal lasers focused on the specimen and causing tracer molecules to fluoresce. This enables the investigator to selectively label a target cell or structure and precisely define the label in a 3D image. These lasers are highly intense variable monochromatic light sources. They are being used for a variety of techniques including optical trapping, lifetime imaging studies, photobleaching recovery, and total internal reflection fluorescence. By changing the wavelength of the illuminating laser, different fluorescent dyes can be excited to produce multi colored, multi labeled images.
The lasers and scanning system of the Confocal microscopy is an outcome collection of few parts of the microscope. The first part is the collection process from the light of the spatially filtered specimen points. The second part is the electronic signal processing and the last part of the collection is on the visual display as corresponding image points. This collection would require a mechanism; the mechanism would start from the scanning more specifically from the focused illuminating beam through the specimen volume under observation. The principle for scanning processes is basically from stage and beam scanning or from both with the array of light points through apertures in a spinning Nipkow disk [1]. Each technique has its own uniqueness in the performance of specific confocal applications.

The Confocal Microscopy helps to see the images in 3D. The general techniques of this device are to produce the optical sectioning and 3D reconstructions. The results of the 3-dimensional structure of it will be a generate movie that would show the structure. The conventional microscope has helped the science field to see the objects and shapes from different angles; therefore, they have been given the name wide-field microscopy or wide-field epifluorescence microscopy to it. The pictures that are produced are both in-focus and out-of-focus light. Therefore, the confocal images of the same specimen will be clear and the out-of-focus parts will be gone. Because each layer illuminated by the laser is extremely thin and therefore in focus and then multiple in focus layers are assembled to form the electronic image.

The techniques that make this possible are in two parts. One is the optical sectioning part, and the second is confocal. The optical sectioning part can take those in-focus images and take a slice through a sample. This will be repeated over and over until the picture is completed. The in-focus images, the ideal of Microscopy, excited fluorescence inside that sample goes through the objective, comes up through a tube lens on the top of the microscope. The another aspect of the confocal Microscope is the pinhole that is the same focal plane to take the in focus lights of the image out. This experiment would bring the attention, what should be done to have the better image instead of scanning each picture, that might miss a spot. The rest scanning, which will take an image each spot in the sample at a time and measure how bright that one spot is. This would repeat for each spot of the image. This process will take the intensities of each part of the picture and creates the new picture of the sample. For this procedure either, the laser or illumination system would be needed. However, for the confocal microscopy, the lasers would be better because its high powered can be focused more on the spots and rejects other lights that are coming to the sample.

**Parts of the Microscope**

Introducing laser systems for Optical microscopy and Confocal microscope scanning system; the laser system has very high-intensity monochromatic light sources. They are very useful in many ways in the health system. These qualities and usage will be verified in more specific details. They can be used in optical trapping, lifetime imaging studies, photobleaching recovery, and total internal reflection fluorescence. Figure 1 shows one of the examples of pre-bleach and post-bleach.

The Confocal Laser Scanning Microscope, CLSM, setup is based on a lamp and a laser beam that is being focused on the object. In figure 2 the diagram shows label parts of the microscope for additional information. One of the mirrors adjusts the beam direction in X and Y directions. Therefore, the mirror brings the beam to the back focal plane of the objective lens; this will bring it to focus on the sample; this needs to be noted if the subject is the fluorescent part of the light that will pass back into the objective lens. The light then passes through and reflects the detection system. This would show only small parts of the sample; however, if this light is a different color from the laser light, emission filters are used to separate it from the laser light that has been reflected from the sample. If reflected light is being examined, it will be passed through a polarizer.
that will allow only the laser light with a different polarization angle from the initial laser light to pass [2]. The photomultiplier is being used in the confocal microscope to detect and amplify the light signal. The output will be an electrical signal and initial light signal; then they will be converted to digital numbers, and by connecting it to the computer, those digital numbers and waves will become a complete 3D picture of the sample. The overall introduction of the parts of the microscope: Laser, Beam splitter, Scanner, Objective lens, Z-control, Pinhole, and Photomultiplier tube. They all mostly have the same characteristics as in regular microscopes.

The confocal microscope scanning systems are the collection of multiple pictures that are being taken in different moments. In figure 3, it shows the systematic approach of the process. This has point by point required mechanism to proceed; these pictures will be placed in the computer system to develop the overall image. There are three parts of the microscope that are very important in this process; stage scanning, beam scanning, and nipknow disk. Each part has their own specific value to produce a better picture in confocal applications. In order to obtain sharp images, the object must receive as much deflected light as possible; this can be done with the wide angular opening. The process of the objective’s light capacity can be captured mathematically by the following formula: \( \text{N.A.} = n \cdot \sin \alpha \) In figure 4, there is an example of this procedure. To place this formula in an example, per se, if N.A is equaled to one, this can happen by using the oil immersing, then the \( \alpha = 90^\circ \). Placing this in practice, N.A values is less than 0.95, only with immersing oil, it can be a higher number. The airy disc is where an objective and light beam path, projected as a bright central spot. It was George Biddell Airy who first theoretically described this phenomenon in 1835, although others had previously observed it experimentally[3]. This would follow by having the smallest distance between the resolution. There is another formula which is very similar to the crevice one. This has been shown in figure 3 in part C. The formula is shown in figure 5.

**Imaging Mode**

The fluorescence microscope has the higher power; it is a greater tool to be used for specific cellular and molecular imaging. Modern fluorescence microscopes can maximize the collection of emitted fluorescent light while minimizing the collection of the incident excitation light [1]. The quality of the image is dependent on the ability of the microscope to pass the fluorescent light to the detector. This is occurring while the excitation light is being blocked. Mostly everything can be seen by this; even single molecules can be visualized in the fluorescence microscope. There are many ways to combine a better picture. This process would have an ability to produce high-resolution images through sequences; this process is within thick sections. In addition, the data is being collected within the different wavelengths; single, double, triple, or multiple wavelengths. The collected data and images will be illuminated and labeled strategized and registered with each other [1].

In the health and science field, the imaging of live cells can be time overlaps. Many of the fluorescent dyes used are non toxic and hence the progression of label distribution can be followed in living cells over time. This process has to go through the processing methods; these methods are being applied to sequences of images. There are two series that will occur before they go through the 3D data processes; these two series are z-series and three-dimensional representation of specimens. This is being followed by 3D data processes where it has a four-dimensional imaging. Usually reflected light imaging mode is being used in confocal imaging but any other transmitted lights that commonly are being used in the laser scanning confocal microscope can be used.
The laser scanning confocal microscope (LSCM) has usually been used to produce the digital three-color imaging of confocal microscopy. The three colors that are being used are red, green, and blue; which are called RGB color. Each color will represent a description in labeling the cells. A simplified version computer software that is being used to produce three-color confocal images is named Adobe Photoshop. A more complex type of software is the Nikon MicroscopyU Confocal Image Gallery. Nikon software will produce a digital image sequence; PCM-2000 confocal microscope scanning system. The system is called Eclipse, E-600 upright microscope. The Acousto-optic tunable filters (AOTFs) is another type of tool that has the ability to transfer the illumination wavelengths on a pixel-by-pixel basis while maintaining a high scan rate.

In the confocal microscopy, there is a possibility of the overlapping of digital images in labeling. This effect is known as Colocalization. This might occur at the moment of labeling the molecules; molecules bind to targets that lie in very close or identical spatial positions. In addition; if a single cell contained two different molecules that were labeled by different fluorescent molecules, for example one molecule that was labeled with a red fluorescent dye and the other molecule that was labeled with a green fluorescent dye the resulting color observed would be the mixture perceived as yellow.

**Electronic Light Detectors**

In this modern field of microscopy, the fluorescence, and confocal microscopy, the laser scanning has a very important responsibility. The laser scanning will collect and measure the secondary emission. They are being gathered by the objective lens or solid-state charge-coupled devices (CCDs). The light is highly affective in imaging. This can be done by a process to making sure that the object has enough light. The fluorescence emission light is located near the image plane and is through a pinhole. However, the light is far away from the objective plane; therefore, it would be needed to reduce the light more. Using the highly sensitive photon detector is very necessary; it has a very high effect in responding with a high level of sensitivity to a continuous flux of varying light intensity. Extremely weak signals can be averaged and electronically amplified to improve detection and resolution.

**Confocal and Wide-field Fluorescence Microscope in the Health Field**

The Confocal and the Wide-field are both very useful in the field of medicine; however, the confocal has more advantage over the Wide-field Fluorescence microscope. One of the highest and most valuable significant is the control of depth of the field of the object and the image. Another value that can be named is the elimination of reduction of background information away from the focal plane of the object in the image. The least but not last significant is the capability to collect serial optical sections from thick specimens. These are being collected by the several wavelengths and from the image of the object. In addition, the other greater values that are very significant and valuable in Confocal microscopy is the ability to use the spatial filtering techniques to eliminate out-of-focus light or glare in specimens. This will allow seeing the thickness and the dimensions of the focal plane. The Confocal microscopy is being used in many different field of medicine. One of the field of medicine that can be named is the biomedical science. The confocal microscopy is being used for imaging either fixed or living cells and tissues. Each part of the image is being labeled with one or more fluorescent probes. There is a large number of them that are being used for education, learning, and teaching system. For example; labeled nuclei, the Golgi apparatus, the endoplasmic reticulum, mitochondria, and also dyes such as fluorescently labeled phalloidins that target polymerized actin in cells.
This needs to be noted, in this part, the confocal wide field fluorescence is not being comprised for purchase advantages. It is being comprised for the usage of the research field and betterment of the research results.

In comparing the Wide-field versus the Confocal, their images are different. The more light, the better the image will be. In figure 6, it shows two pictures of the samples from Confocal and Wide-field. Sample A is an example of Confocal and sample B is an example of wide field image. In the Wide-field microscopy, the images have more specimen planes outside of the focal plane and have more information of interest from the focal plane.

One of the other fields of medicine that has been used for is research method and developing a better treatment plan for the patients. The recent study that was done, it was for the HIV-1 treatment. It was done in 2008 and the topic is HIV-1 Nef Induces a Rab11-Dependent Routing of Endocytosed Immune Costimulatory Proteins CD80 and CD86 to the Golgi. The Confocal and Wide-field fluorescence were very helpful in this research experiment to show the cells and tissues. The cell surfaces are the parts that are being more incubated for various times as indicated. Then after the cells are being identified and absorbed, they are being imaged by either high-resolution wide-field fluorescence microscopy or confocal microscopy [6].

To add more about the Fluorescence microscope. The Fluorescence microscope is a common tool that is being used in the biology field. To give more information about the Fluorescence microscope, the instrument can name the beam formation. The beam formation of the Fluorescence microscope is parallel beam light simultaneously illuminates the whole specimen. The wavelength in Fluorescence microscope is being obtained from the excitation light. Also, the excitation light is being formed from the high-pressure mercury or xenon. The pictures of the Fluorescence microscope can be viewed by the eye when captured electronically. This needs to be noted if the object is out-of-focus or the optic is very limited the image will result in low contrast and spatial resolution.

Overall the Wide-field microscopy has the greater values. The images of the Wide-field microscopy illuminated simultaneously, allowing for faster imaging. However; this is all based on the cameras abilities to perform better and higher picture. It is very common to use the Wide-field microscopes image to observe the object in real-time by eye. This is the best benefit for the user. Choosing the Fluorescent types of the microscope can obtain better image but it requires greater specimen preparation, more technical expertise, and extensive electronic image manipulation in software. The Wide-field microscopy is being mostly used in researchers because it has the low cost. They are other advantages that the Wide-field microscopy has; they can be named as the simplicity and easy use of the tool this would show the flexibility of the system in a different field of the science. [8]

**Orientation of Pictures:**

In the orientation of pictures, of an object, the wide picture of the smaller pieces is the main point. One of the most basic examples that can be given is a picture of collagen fiber. In figure 7, the collagen fiber can be visualized with a confocal microscope and it has been labeled with the fluorescent marker CNA35-OG488. The collagen fiber is very small; the main point is to find a wide picture of the object. The slide of the sample is being made with a very small amount of collagen spread out by a pin on a silicon Petri dish filled with PBS. The position of the sample is being placed upward. The image analysis of the example is given in figure 8. Measurement of the collagen is being done by using IMARIS software. IMARIS software is bitplane’s core scientific software model that delivers necessary functionality for data management, visualization, analysis, segmentation and interpretation of 3D and 4D microscopy datasets. However; there might be a
moment that the object is being stacked on top of each other and not being very clear in the image. In this case, the collagen is being stacked over each other. This software is still useful and can make the image clear to be seen. The example in figure 8 is in 3D. The orientation of fiber in 3D could be described by the following formula:

\[
\delta (\text{between } -\pi/2 \text{ and } \pi/2) \text{ and the latitude (radial) angle } \varphi (\text{between } 0 \text{ and } \pi) \quad [7]
\]

Furthermore; this will not be the same or easy to be separated from all other objects. For example; the experiment with the mitochondrial targeting sequence, is not easy as the collagen fiber. The rapid mobility of the ions in the cytoplasm helps in increasing the image be more clarify. The imaging ions example can be; Ca2+ or H+. This process needs to be undergoing with the cytoplasm; an example for that can be named as following: Fluo-4 or SNARF.

**Conclusion:**

Understanding the structural of the microscopes and the best usage of each is highly important. In my opinion, all the scientist should have enough knowledge in this field. The acknowledgment of different types of microscopes will help the scientists financially and improve their time efficiency. Financially to know what type of microscope is better to be purchased for what type of research and studies. And time efficiency to comply their task in a shorter period of time and with the best quality.

In this research, I have learned about significant values of each part of the microscope and how to obtain the pictures. The Physics part of this research that it was very interesting to me it was about the digital pictures from microscopes. The digital pictures from the wavelengths that were created from the lights and was sent it to the computer information system by digital numbers electronically. The combination of all the numbers and wavelengths was the final picture. The final picture was basically the combination of wavelength every single spot of the object. Furthermore; this was followed by understanding more about the electronic light detectors, protons, neutrons, and identifying different types of wavelengths. Which it was more specifically explained later on in the paper about the orientation of the picture. And then I have learned how to prepare the slide to visualize the picture. Before understanding more about the orientation of pictures; the two types of microscopy, Confocal, and Wide-filed, are being explained more in details. In this part of the research, I learned more about the significance of each type of microscope. It was very interesting to me, the side studies that I did, the types of the researches that were done with each one of the microscopes and the treatments that were done. One of the most interesting ones to me was the HIVE treatment. More about the Orientation of picture, I also learned if the object is too thick to eliminate the picture for labeling and naming, there are software that is very helpful to process the coding system into the computer information.

Overall I have learned how important it is to know more about different types of microscopes and the research studies that are being done with each type of the microscopes. It was very helpful research to experiment and absorb more about the orientation and characteristics of each part of the formulas. Also how the formulas are being applied into the experiments.
Figure 1:

Pre-bleach and post-bleach; as one of the examples of using the confocal in health system. [4]

Figure 2:

Introducing parts of the confocal microscope: [2]
Figure 3:
Scanning image on confocal microscopy. [5]
Figure 5:

\[ d = \frac{\lambda}{2 \text{N.A.}} \]
Figure 6:
[4]

Figure 7:
[7]
Figure 8:
[7]
References:
Genetic Engineering Will Change Forever Through CRISPR

Gage Gommels

March 23, 2017

BIO 181

Dr. Browning
Abstract

Two-thirds of the 150,000 people that die today will die through age-related causes. But what if a scientist has cracked the code to slow down or even reverse aging. A new developing tool in genetic engineering is called CRISPR, and it has revolutionized the way humans can modify our DNA. The revolution started when scientists figured out that CRISPR is programmable. Just like how bacterium fight off viruses we have found a way to store these viruses in our very own cells. This won’t just cure diseases, but it makes DNA-programmable. Soon we will have genetically modified people walking among us. CRISPR could even offer the cure to aging! This could be done by borrowing genetic material from other animals like the Turritopsis (a type of jellyfish ages reversely), and we could apply it to our gene strains. You could say if the old techniques of genetic manipulation were like a map, CRISPR is like a GPS system. Being precise, and cheap CRISPR offers the capability to edit living cells, turn genes on and off, and study DNA sequences. CRISPR also works for every type of cell. This includes bacteria, florals, animals, and or humans. Through CRISPR humans could finally achieve curing all genetic diseases on the planet!

CRISPR

Back in the 1980’s we were always told that technology is going to increase exponentially. We were told robots and getting to other planets were going to be possible. Most people looked at companies like Apple, Nasa and Microsoft like they were crazy. But 30 years later our science fiction became our reality. We now have computers in our pockets that are 1000% smaller and 1000% faster and more efficient than the ones that we used to get to the moon and most of the time we don’t even think about it; it's just become a standard of living. In the science world we are roughly in the same situation with genetic modification; but where did it come from and what plans do we have with this new breakthrough? Us humans have been modifying genes for thousands of years. This was done through selective breeding, and we were able to strengthen the traits that were useful in plants and animals. Take a banana for instance. Some bananas back then had a thick skin layer and big round seeds in the middle which made it hard to eat. But now our modern day banana has a thin outside skin and seeds that we don’t notice when eating them. Our ancestors became excellent at this, but we never understood how this process worked until we discovered DNA. DNA, in summary, is the code of life and is a complex molecule that will guide the construction and growth of living organisms on this planet. In the DNA strand, four nucleotides have a set of instructions for the cells to carry out. If you change the instructions, you modify the animal taking it. When humans discovered DNA we naturally wanted to know what we could do with it, so the experiments began. In the 60’s, scientists shot plants with large amounts of radiation to see if they could get a plant with an unusual or useful mutation, and to my surprise, this sometimes worked out. In the 70’s we inserted DNA into bacteria to study them for research and
medicine. In 1974 the first genetically modified animal was born, which was a mouse. This made mice a standard for testing genetic modification, on which as we find today, saved millions of lives because of the trials and research we found from them. In the 1980s the first commercial genetic modification was available. The first patent for a bacteria that was changed to absorb oil was given. Today we produce a lot of medicine like insulin through the process of genetic modification – all things that we used to gather from organs from animals before. In the 1990’s the first engineered baby was born. To treat maternal infertility, children carried genetic information from 3 humans. Today, we have created featherless chickens, fast-growing salmon, muscle pigs and even see-through frogs. We have even made zebrafish glow, and you can now find them in the pet’s store for a low price of $10. This already is impressive, but until recently genetic engineering has been very expensive, complicated, and took copious amounts of time to process. This changed with a revolutionary modern technology called CRISPR. This new tool is going to revolutionize genetic engineering. Overnight the costs of DNA manipulation went down 99% and on top of that CRISPR only takes a fraction of the time.

Another beautiful thing that CRISPR has to offer is the ability to edit live organisms. Before we had to grow and see the effects through new generations, but CRISPR will affect the body while it’s alive. A good analogy is that if the old way of genetic engineering was a map, then CRISPR is a GPS system and has unlimited potential for a scientist to play around with. CRISPR just appeared out of nowhere. Let’s talk about where it came from, what we plan to do with it and what might be planned for the future of the human race.

The system CRISPR has been around for millions of years and it originated from the way bacteria defends themselves. For the longest time life has been on earth bacteria has been hunted down by bacteriophages. Every day 40% of the bacteria in the ocean is killed by so-called bacteriophages; this is done by the phages inserting their genetic code into the bacteria to make them into their personal factories. Most of the time the bacteria fails to defend itself, but sometimes we have a lucky survivor. This is when the bacteria’s secret defense system kicks in. If the bacteria does survive, then it will save a strand of the DNA from the virus and put it into a DNA archive called CRISPR. So, next time the virus attacks the bacteria then the bacteria will make an RNA copy and load it into a protein called CAS-9. From here CAS-9 will be able to go down the archive and scan for a replica of the RNA copy. When it finds a copy, it will cut it off rendering it useless. Humans have known about this system for a long time, but everything changed when we found out that the system is programmable. You can just give it a copy of DNA you want to modify and put the system into a living cell. This makes the process very easy and cheap to do and possible in almost any lab. CRISPR also allows scientists to experiment on the living cell which was impossible to do before. CRISPR also allows scientists to study a particular types of DNA sequences. Not only that, CRISPR will work in a living organism; this includes plants, microorganisms, animals, and humans.
CRISPR has the potential to change the world and we just need to invest the money and time to perfect the process. But what can we expect from a revolutionary discovery like CRISPR? Well from what I have read, it looks to be that CRISPR has almost no limits. So far scientists have predicted three primary uses for the new tool CRISPR. These uses are, ending disease, designer babies and possibly extending the life of humans by a vast amount.

So far CRISPR was only discovered a decade ago, and we are just now starting to realize the potential CRISPR has for the future of the human race. When scientists first found out what CRISPR can do, we decided the first thing we wanted to test is that can CRISPR cure disease. So of course, we use mice as the first test subjects for a living mammal. One of the first experiments was to try to use CRISPR to correct the wrong letter in the DNA in a group of mice with muscular dystrophy. There were three groups reported that had CRISPR to snip out the defective gene in mice with Duchenne muscular dystrophy, allowing the mice to make a vital muscle protein (Kaminski, 2016). As it looks, CRISPR can be very accurate, and that’s what makes CRISPR so unique. It can effectively target a single nucleotide and make the cut to be able to fix the wrong letter. We can also add DNA to the cells to help them to become better disease fighters. This has already been experimented on mice and rats infected with the HIV. The trial was proceeded by having 99% of the rats and mice’s cells living with the HIV (Recent research…., 2016). Then Kamel Khalili, the director of the Comprehensive NeuroAIDS Center at Temple University would begin to program CRISPR to cut out only the viral cells in the rodent’s body (Recent research…., 2016). Keep in mind that the rodents brain, liver, skin, and other organs were infected with the HIV. Khalili studied CRISPR’s ability to eliminate HIV from both mice and rat subjects and found that it was advantageous in separating out the virus in more than 50% of the cells of each type (Kaminski, 2016). This is a breakthrough and a massive step forward to help cure HIV. A statement from Kamel Khalili “proving that it works in a living animal is a huge step forward to developing the technique as a possible treatment, or even cure, for HIV-AIDS.” (HIV…) Kemal also states that with the combination of HIV pills and the CRISPR we could start to formulate a treatment or start to sterilize people from HIV altogether (Kaminski, 2016). It’s very exciting to see diseases like HIV to have a possible cure and just reading up on the article does sound like they have promising results for a possible cure for the future.

CRISPR could also get rid of one of our worst enemies that have been around for millions of years, cancer. Cancer occurs when cells refuse to die inside your body and then conceals themselves from the immune system. This is what makes cancer so hard to fight because your body along with other medicine cannot find it in your body. What CRISPR has to offer is the means to edit your immune cells and help them to be better cancer fighters. Finding a cure for cancer could mean just getting a couple of injections of a few thousand of your cells that have been engineered in the lab to heal the patient (Reardon, 2016). The first clinical trial
for the use of CRISPR on human cancer patients was approved on 21, June 2016. Not even a 
month later China said they were going to start CRISPR therapy in patients with lung cancer. 
China began to experiment with CRISPR in August 2016 (Cyranoski, 2016). The experiments 
are still going on, and there is no documentation online about the results of the cancer patients, 
but soon hopefully we will see what happened to cancer inside those victims. It is predicted 
that in a couple of decades’ cancer along with other life threatening diseases will be cured due to 
the new tool CRISPR.

CRISPR is not only a means to help cure disorders and diseases; it’s going to be a 
possible instrument for the future of humans that want to have designer babies. For those who 
don’t know designer babies is a baby whose genetic makeup has been selected to eradicate a 
particular defect or to ensure that a particular gene is present (Designer Babies). This means 
that there would be small and irreversible changes to human genetics. The will to edit the 
genetics of a human embryo already exists, but it’s in its baby stages. To our knowledge, it has 
already been attempted twice. In early 2016, the Chinese experimented with human embryos 
and was somewhat successful on their second attempt (Ledford H, 2016). This shows the 
enormous challenges we face when editing the genes of a human, but it also represents that 
scientists are dedicated to solving them. These modified humans will change the genetics of 
the human race forever. This is because the engineered traits of the genetically modified 
human will be passed down to their offspring and this will gradually change the gene pool in 
humans forever (Wu Y, 2014). This process will start slowly; the first designer babies will 
most likely be edited to fix a deadly genetic illness running in the family. As CRISPR 
improves and progresses over time more and more people will begin to argue that not using 
gene editing is unethical because it would sentence the child years of suffering and overall it 
denies them the cure. Many people today will claim that genetic engineering is wrong, but 
when the first genetically modified kid is born, a door has been opened and will never be 
closed again. When this happens, the physical traits of a human will most likely be left alone, 
but when time goes on, and humans get more and more comfortable with genetic engineering, 
then we might start changing the vanity traits to people as the parents wish.

There are many more uses that CRISPR has to offer, and one that people plan to start 
very soon is to edit the genes in mosquitoes to prevent them from spreading disease from 
human to human. An illness that we have in mind is malaria. By using CRISPR to edit the 
gene to where the mosquito cannot spread it could save millions of lives and dollars. Over 600 
million cases have been reported each year, and over 1 million people die each year of this 
disease over in 3rd world countries. Malaria is cured, but the trouble is to get the cure to the 
people. But now instead we have the option to eliminate the source and stop the spread of a 
very well-known disease that has killed billions of people over the past millennia years. To 
cure malaria, we would need to genetically modify a whole insect population. A scientist has
proven that this will work, it’s just up to humanity to decide whether or not we would release these mosquitoes into the wild. Yes, these mosquitoes exist today, and we were able to achieve this by adding an antibody gene into the male mosquitoes DNA, making them immune to being able to carry malaria (Hammond, 2015). Unfortunately, just changing the genetic information was not enough because there was only a 50% chance that the gene would carry down to an offspring. So in numbers, it would not make a difference. But scientist was able to get around this. An engineering method called the gene drive fixes this problem (Hammond, 2015). It makes the new gene to become dominant in the following generations overriding the old gene completely. Thanks to this method, 99.99% of the offspring will carry the gene that prevents malaria from entering their body (Hammond, 2015). Scientists predict that if we release enough of these mosquitoes into the wild, they will mate with others and spread the anti-parasite very quickly. So quick that malaria won’t have time to adapt, this would virtually wipe malaria from the planet in just a few years. The barrier that these mosquitoes face is the ethical problems that we have in our government. We are still unsure of what would happen if we did release these mosquitoes into the wild. But if you take into account that by the time you have read this paper ten children have died of malaria, it seems to be a good option to go ahead and try to get rid of the parasite altogether. This could also be a doorway into a world with no deadly disease spread by insects. We still have mosquitoes with Zika virus and Dengue fever, ticks that spread Lyme disease, fly transfer sleeping sickness, and fleas transmit the page. Humanity has the chance with CRISPR to save millions of lives and stop the suffering in 3rd world countries on an unimaginable scale. CRISPR is technology and must be used with caution, but at some point, we have to ask the question is it unethical not to use this technology, when every day 1000 children die. CRISPR is a very powerful tool that can allow humans to do unbelievable things, and one feat that is being looked at is the cure to aging. (see figure 1 for more details)

66% of the 150000 people that die today will be because of age-related issues. Today we think age is cause by gathered damage to our cells; like DNA breaks and the system is responsible for fixing this problem over time, except that it starts to wear out. There are also gene related issues that directly affect aging as well. But with the use of CRISPR and other types of therapy could stop aging or it might even be able to reverse it (Rodero S, 2011). We know in nature that some animals are immune to aging, like the lobster, Turritopisis Nutricula, and Planarian. We could even possibly take a few of those genes and apply them to ourselves. Some scientists believe that it is feasible to help prevent aging and slow it down drastically. So instead of dying in hospitals at the age of 80 to 90, we would be able to spend a few hundred years with our loved ones. The research into this is in its baby form, and it could take hundreds of years to perfect. The challenges are enormous, and the end of aging could be unachievable. But the people that walk this earth today will be the ones that will benefit from this the most, all we need is a smart billionaire to make this their next problem that they should try to solve.
CRISPR is a revolutionary discovery and there is still a lot to the system that we still don’t know about. It is barely four years old, and yet we have already proven that some of these feats might be achievable. Having a genetically modified immune system with a whole list of diseases could wipe out some of the most terrifying and horrible conditions of the face of the planet. Looking even further into the future, we could engineer humans, so they are a better fit for space travel and other planets in our hostile universe. But still, there are some huge challenges that await for CRISPR, some physical and some ethical. One of the problems that comes to mind is that the public fears that we will create a new standard for humans rejecting the no perfect ones making a divide between natural born and genetically edited; But we already live in this world. Today, pregnant women regularly get an ultrasound of their babies to check for genetic diseases and to make sure that their baby is normal. Unfortunately, when there is a small sign of a genetic defect in the baby the pregnancy is often terminated. If we look at Down syndrome alone, in Europe 92% of the pregnancies that detect Down syndrome are terminated. This is really up to the parents carrying the child and is highly personal, but in reality, we have to acknowledge the fact that we are choosing humans based off of their medical conditions. So we have to act respectfully and carefully as scientists slowly make the changes to the human gene pool. CRISPR despite its power, and yes it’s very powerful, still has many bugs that scientist needs to work out. For example, CRISPR sometimes changes the wrong gene, and we have no idea what might happen to a human if we are editing their genetics in a live human. We might be able to cure a disease, but we could also possibly make some unknown error that could go unnoticed, cause even more trouble for us in the future. Working on the accuracy and monitoring methods is something that scientist are working on as the first human experiments begin. The basic proof of concept for CRISPR is real and exists today; the technology is that powerful. This might lead to some reasons why we should ban genetic engineering, but that would be a mistake because that would mean the technology would go somewhere without any rules or regulations. Only by staying involved can we guarantee that the advancement of CRISPR will be used for active purposes. Through CRISPR we could end many diseases, increase our life expectancy by centuries and we could even voyage out into our solar system. There is no thinking small when it comes to a revolutionary advancement like CRISPR. Whatever your opinion is on genetic engineering, the future is approaching not matter what you think. Our science fiction will soon become our reality, a reality full of opportunities and great challenges.
FIGURES

Figure 1

This figure represents the cas-9 protein showing where to cut and insert new DNA sequences.

Figure 2

This is a visual representation that shows what is the process to manipulate the gene pool of a mosquito. Image was taken from (Hammond, 2015)
References


4. Use of CRISPR Cas9 gene editing therapeutic shown to permanently inactivate HIV-1 in patient's blood for first time. Biotech Week 2016 04/;171.

5. Recent research from case western reserve university highlight findings in HIV/AIDS and genetics (elimination of HIV-1 genomes from human T-lymphoid cells by CRISPR/Cas9 gene editing). AIDS Weekly 2016 03/;64.


7. Sherkow JS. Law, history and lessons in the CRISPR patent conflict: Predicting the outcome of the ongoing patent disputes surrounding genome-editing technology is equal parts patent analysis and history. Nature Biotechnology 2015 03/;33(3):256.


