21st ANNUAL MANCINI SCIENCE SYMPOSIUM

May 14, 2015
Volume I
Foreword

Paradise Valley Community College is proud to present this 2-volume set of the 21st Annual Mancini Science Symposium. This symposium was held on May 14, 2015 in the Center for Performing Arts (CPA).

Students enrolled in Astronomy, Chemistry, and Physics classes from PVCC participated in the event. Each contributor was responsible for selecting and researching his/her topic and preparing a paper. This 2-volume set contains all 85 research papers (28 in Astronomy, 3 in Chemistry, and 54 in Physics). A few students gave oral presentations of their project to their peers. Students chose the oral presentation topics.

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

Casey Durandet, PhD
Scott Massey, PhD
Jennifer Weitz, MS
## Table of Contents – Volume I

<table>
<thead>
<tr>
<th>Title</th>
<th>Author</th>
</tr>
</thead>
<tbody>
<tr>
<td>Skydiving with and without a Wingsuit: What is it, and what makes it Possible?</td>
<td>by Hassan Abbas</td>
</tr>
<tr>
<td>Electricity and its Physical Interaction with the Human Brain</td>
<td>by Hassan Abbas</td>
</tr>
<tr>
<td>The Physics of Electric and Magnetic Phenomena</td>
<td>by Hassan Abbas</td>
</tr>
<tr>
<td>X-Ray Diffraction: A Milestone in Scientific Technology</td>
<td>by Jamileh Al-Asmer</td>
</tr>
<tr>
<td>The War of Currents: The History, Benefits, and Future of Direct and Alternating Current</td>
<td>by Michael Alexander</td>
</tr>
<tr>
<td>Regenerating Brakes and their Potential</td>
<td>by Greg Appenzeller</td>
</tr>
<tr>
<td>The Physics of Horseback Riding</td>
<td>by Ameena Arekat</td>
</tr>
<tr>
<td>Retiring the Drill: The New Age of Dentistry</td>
<td>by Brad Babits</td>
</tr>
<tr>
<td>Age and Distribution of Stars within the Milky Way</td>
<td>by Taylor Bader</td>
</tr>
<tr>
<td>Mass and Radius of Stars</td>
<td>by Jaimie Ball</td>
</tr>
<tr>
<td>Your Pharmacy is Radioactive</td>
<td>by Alanda Barash</td>
</tr>
<tr>
<td>The Development of Wafers from a Silicon Ingot</td>
<td>by Tyler Becker</td>
</tr>
<tr>
<td>Hubble’s Law and the Hubble Constant</td>
<td>by Luke Beckham</td>
</tr>
<tr>
<td>Ptolemaic Astronomy</td>
<td>by Annie Behrendt</td>
</tr>
<tr>
<td>Analysis of Carbon Nanotubes, Polymers, and Applications</td>
<td>by Cody Boeckholt</td>
</tr>
</tbody>
</table>
Major Planets vs. Dwarf Planets
   by Devin Cameron

The Physics of Artificial Joints for Orthopedic Applications
   by Abby Duval

The Physics of Applied Scanning Electron Microscopy
   by Victoria Engle

Space Exploration: Complications and New Possibilities of Exploring the Cosmos
   by Evan Entze

Blasters and Battleships: A Brief Overview of Laser Technology and Modern Laser Weaponry
   by Austin Fendler

The Life and Death of a Low Mass Star
   by Victoria Garcia

The Physics Behind Prosthetic Limbs
   by Anthony Gervasi

Motion in the Spectral Analysis
   by Lucas Godfrey

Exoplanets
   by Leah Goldberg

Telescopes in Orbit
   by Cory Gomes

Deep Brain Stimulation: Mechanism, Use, and Treatment
   by Emily Hanka

Seeing the Light: The History of Lasers and Their Impact in Dermatology
   by Tiffany Hanze

Illuminating Sports
   by Rachel Hauser

Fluid Dynamics and the Body
   by Katie Hawk

Sustainable Batteries
   by Robert Heil
Chemistry, Cavities, and Crowns: Studying the Chemical Processes and Materials in Oral Health
    by Matthew Jones

An Analysis of Different Light Energies
    by Tamara Juarez

You Can’t Touch Anything
    by Matthew Kassir

Drifting in the Concepts of Physics
    by Kelly Koeplin

Behind the Blood: The Truth Revealed through Blood Spatter Analysis
    by Megan Konves

Communicating with Other Worlds
    by Alee Kopp

Asteroids and Comets
    by Bridget Kraiss

Pacemakers
    by Jessica Krehbiel

Drainage Design and Hydraulics vs. Hydrology
    by Jonathan Krukow

Moons of Planets
    by Kate Larsen

Roof Gardening and Energy Consumption
    by Marjan Lavasani
Skydiving with and without a Wingsuit: What Is It and What Makes It Possible?

By Hassan Abbas
11/26/14
PHY 111 (Section 33260)
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Abstract: The purpose of this paper is to understand the mechanics behind Skydiving and Wingsuit diving. The paper provides historical and factual information about each sport. After the definition of each sport is clearly defined, the mechanics of each sport are explained to clarify how the sport of freefalling safely is possible.

Since the dawn of time, human beings have been fascinated with the bird’s capability of defying gravity. Freefalling is the closest experience humans have achieved to grasp the sensations caused by flight (neglecting sensations caused by aircraft experience). According to Cambridge Dictionaries Online, “the sport of jumping from an aircraft and falling through the air before opening a parachute” is referred to as Skydiving. The thirst of flight did not stop at falling vertically through the air. Skydivers who use the aid of a Wingsuit are able to free-fall vertically and horizontally. As far as known, the concept of safely freefalling at high altitudes began before airplanes were invented.

Skydiving has a remarkable history. First of all, the farthest record of “humans falling through the air with some sort of aerial support dates back to China” (Lerner, 2007). Though not much is known about the composition of the parachutes used, it is confirmed that in the 12th century, the Chinese used “umbrella like parachutes to jump from high places and float on the ground” (Lerner, 2007). Progress in safely freefalling did not begin until the late 18th century. A French physicist named Louis Lenormand “jumped from a tree while holding two parasols” (Lerner, 2007). Two years later, “in 1785, J.P. Blanchard used silk to take the first parachute that was not held by a rigid frame” (Lerner, 2007). Less than two decades later, in 1802, the very first person who safely jump off a hot air balloon at an altitude of 8,000ft was Andre Garnerin (Lerner,
2007). Garnerin’s jump set the stage for future skydivers because skydivers now had the full confidence in freefalling at high altitudes. While the thirst for flight continued, it was only till 1912, “Captain Albert Berry of the U.S. Army accomplished the first parachute jump from a moving airplane” (Lerner, 2007). Skydiving did not stop simply as a sport for stuntmen. The U.S. military used parachutes to their advantage during World War 1. On 1922, the “first emergency bailout from an airplane occurred” (Skydiving History, 2014). Moreover, paratroopers in World War 2 “turned the tide of the war against the Axis powers” (Skydiving History, 2014). Parachuting began to look like more of a sport when the “first parachuting competition was held in Yugoslavia” (Lerner, 2007). In summary, when neglecting technological advancements, skydiving has not changed significantly over the course of its history. Skydiving mainly consisted of falling through the air with some sort of device attached to the person. It is truly remarkable how long human beings have been trying to skydive. Nonetheless, it is crucial to understand not only where skydiving originated, but also the different components of skydiving.

Skydiving is not limited by its simple definition because there are many methods of skydiving and a different type of skydiving. Firstly, a skydiver is “a person falling without his parachute” (Long and Weiss, 2014). Broadly as the definition sounds, it is vital to make the distinction between the two phases of the fall. As mentioned previously, the freefall is synonymous to skydiving and when the parachute is opened, the person is gliding. Furthermore, there are three different methods to skydiving. The three major skydiving methods are Accelerated Freefall, Static Line and Tandem jump. Accelerated Freefall is “when a certified instructor accompanies the student in freefall by holding onto the student’s harness” This provides the skydiver with a level of independence
because he is wearing his own parachute. The second method of skydiving is called Static Line. In this method, “the instructor initiates deployment of the student’s parachute when the student jumps from the plane” (Instructor-Assisted Deployment or Static-Line, 2014). The method is called Static Line because the line is what initiates the parachute. A Static Line is “a line of cable where one end of which is fastened to the parachute, the other to some part of the aircraft, [which] is used to activate and deploy the parachute as the student falls away from the aircraft” (Instructor-Assisted Deployment or Static-Line, 2014). In this jump, the skydiver jumps without an instructor. Lastly, the third method of skydiving is called a Tandem Freefall. This method is most popular because of its minimum training time. The training time is shorter compared to the other methods because “the instructor is the pilot in command of the shared parachute” (Tandem, 2014). This means that the instructor is responsible for maintaining a safe landing rather than the student himself. Wesley Moraes, a professional skydiving instructor, had to go through 500 jumps in order to become an instructor. Even after the 500 jumps, he had to go through an instructor course. Moraes’ experience is an example of why the students are willing to be attached to a instructor. Moreover, the skydiver and instructor are attached by a harness, “the webbing of a parachute system that surrounds and retains a jumper” (Skydiving Glossary, 2014). With the harness kept in mind, suspension lines is what connects the canopy to the harness. As for as the skydiving zones, Wesley Moraes reported that in SkydiveAZ located at Eloy, Arizona, the highest altitude skydivers are allowed to jump is 22,000 ft. The reason is because skydivers are required a oxygen tank when jumping over 15,000 ft. Furthermore, there is actually another type of skydiving rather than method. Group formation skydiving is a type of jump where a large number of
participants seek to create different shaped formations in the air as they fall” (Lerner, 2007). The skydivers actually build different patterns with their bodies in freefall. In conclusion, there are three methods of skydiving and one different type of skydiving. The three methods of skydiving include Accelerated Freefall, Static Line and Tandem jump. The other type of skydiving is called a Group Formation skydive. Similar to every other sport, skydiving too has a competitive aspect to it.

There are individual skydivers who compete with each other and there are also skydivers who compete in groups. Group skydivers compete with each other in several different ways. The major group competition is Formation Skydiving. Formation skydiving consists of “teams of 4,8,10 and 16 jumpers who race against the clock to form prescribed geometric formations in freefall before opening their parachutes” (USPA National Championships, 2014). Below is a picture illustrating just one pattern in a group formations:

![Figure 1](image_url): A group formation freefall containing roughly 40 skydivers. (USPA)

The skydivers are not only freefalling over a hundred miles per hour, but also manipulating their bodies into a geometric pattern. Sport competitions usually are made
up of teams, but how could individual skydivers possibly compete with each other? Just as four year olds compete with each other with who can jump the highest from the ground, skydivers settle with who can jump the highest from space. Majority of skydivers jump the “traditional 13,000ft which is about halfway up Mount Everest” (Zaretsky, 2014). 13,000ft is merely the normal height a skydiver falls, and so the current world record of highest freefall was done by a man named Felix Baumgartner. It was “on October 14, 2012, Felix Baumgartner” (Muller, 2014), “jumped from an altitude of 38,969 km” (Barbero, Colino and Tapiador, 2014). He also reached the greatest speed of “1357.6km/h [which] is greater than the speed of sound by 25%” (Barbero, Colino and Tapiador, 2014). Baumgartner holds the world record of the highest altitude a human has jumped from, and also the highest speed a human has attained without the use of an aircraft. It was Baumgartner’s pressure suit that allowed him to travel as such high altitude and velocity. Baumgartner’s suit “kept the pressure around his body something like what a normal human would experience at 11km” (Foy, 2014). His pressure suit prevented him from the risk of ebullism which is when “gas bubbles form in the body’s fluids [which can] can result in unconsciousness or even death” (Foy, 2014). In short, skydivers compete with each other in groups and also by whom can jump from the highest altitude. Baumgartner survived an extremely dangerous jump, but there are always risk factors for even traditional skydivers.

Though the common assertion of risk factors in skydiving is commonly that the parachute fails to open, the real risk falls in the responsibility of the skydiver rather than the technology. Experienced skydivers will say “99 percent of the time it is the people who fail rather than the equipment” (Celsi, Leigh and Rose, 1993). The reason being is
that “instability in free fall is a fatal factor” (Vetenskapsrådet (The Swedish Research Council), 2014). Stability is solely in the hands of the skydiver. The skydiver must jump off the airplane correctly and he must maintain correct posture in freefall. If the skydiver doesn’t maintain stability, the instability of his body “can lead to an unstable parachute activation” (Vetenskapsrådet (The Swedish Research Council), 2014). When there is a probability of an unstable or late parachute activation, Earth’s gravitational force, the mutual force of attraction between any two objects (Serway and Vuille, 2014), would dictate the skydiver’s injuries. Because of how uncommon parachute failure is (considering there is an emergency parachute as well), it is rather the skydiver’s “carelessness or lack of skill in controlling the body or parachute through the air” (Vetenskapsrådet (The Swedish Research Council), 2014). On the bright side, “in 2013, USPA recorded 24 fatal skydiving accidents in the U.S. out of roughly 3.2 million jumps [which is] the lowest amount of deaths per year in skydiving history” (Skydiving Safety, 2014). By inference, there are fewer careless skydivers. Though skydivers are largely responsible for themselves, technology still has the possibility of failure. Wesley Moraes reported that he used his emergency parachute 7 times over the 20 years of his skydiving experience. Technology does have the possibility of failing, but the probability is extremely low considering Moraes jumped over 10,000 times and only used his emergency parachute 7 times. In conclusion, the risk involved in skydiving is largely held in the hands of the skydiver. What exactly would make a person want to skydive if he or she knows that their life is on the line?

People are willing to skydive because of their surrounding environment. Statistics show that “not only did the number of participants in risk-taking activities increase, but
also a wider age range and an increasing number of female participants” (Celsi, Leigh and Rose 1993). The environmental reason for the rise in popularity may be found in the media and technology. By the repetition of dramatic content, the “media may be responsible for the behavior of high risk activities” (Celsi, Leigh and Rose, 1993). For example, the media’s content can provoke emotions pertaining to thrill. The professional skydiver Wesley Moraes decided to skydive the first time after watching the movie Pointbreak. Secondly, technology “enhancement has made it more attractive to jump [off a plane because] equipment is less awkward and more reliable” (Celsi, Leigh and Rose, 1993). Not only does technology comfort the physical body when jumping off a plane, but it also forwards trust towards the success a skydiver has. In summary, people are willing to risk their lives because of their environment. There is an increase of skydivers because of the technological advancement taken place in the course of history.

Technological advancement in skydiving ranges from improvements in reaching high level altitudes and the parachute system. As mentioned previously, hot air balloons were originally used to reach high altitudes. The invention of the airplane made it easier for individuals to reach high altitude levels because skydivers are able quickly attain high altitudes. Furthermore, parachute systems have advanced greatly. The very first parachutes served its purpose to land safely, but did not do so in a stable manner. It is reported that the same Garnerin that jumped off a hot air balloon “became airsick [because] during the descent, the canopy oscillated so wildly [which caused him to experience] painful vomiting for several hours” (Hall, 2000). This uncomfortable descent led the “French scientist Joseph Lelandes [to] introduce the apex vent” (Hall, 2000). The apex vent is a “circular hole in the center of the canopy which eliminates the troublesome
oscillations” (Hall, 2000) because it allows air flow. The parachute system no longer was only a survival tool, but a technological equipment that provided an ease of freefall. Secondly, the material used to make up a parachute advanced over the course of history. The parachute canopies were actually “first made of canvas” (Hall, 2000). Because canvas is a heavy duty fabric, it wasn’t practical to use it anymore, so innovators began to use silk. Silk was more “practical because it was thin, lightweight, strong, ease to pack, fire resistant and springy” (Hall, 2000). Silk proved itself to be superior to canvas, but parachute manufactures had a problem “importing it from Japan during World War 2” (Hall, 2000). Since silk suddenly became scarce, “parachute manufactures began using nylon fabric” (Hall, 2000), and this change was actually for the betterment because nylon “turned out to be more superior to silk” (Hall, 2000). Nylon fabric is “ more elastic, more resistant to meldew, and less expensive” (Hall, 2000). To this very day, nylon fabric is popularly used for manufacturing parachutes though other fabrics such as “Dacron and Kevlar” (Hall, 2000) have been introduced into the parachuting community. Lastly, technological improvements in the parachute such as adding control lines to the system helped make it easier for parachuting because a skydiver is no longer confined to one area of the freefall. With the control lines, skydivers are able direct their area of landing. Moreover, the safety in parachuting was further assured with the Automatic Activation Device. An automatic activaton device is “a self-contained mechanical device that is attached to the interior of the reserve parachute container, which automatically initiates parachute deployment of the reserve parachute at a pre-set altitude, time, percentage of terminal velocity, or combination thereof” (USPA Glossary, 2014). In conclusion, technological advancements in reaching high altitudes and the parachute system include
the invention of airplanes, the design of a parachute, the material of a parachute, and additional equipment added to the parachute. Though there were major improvements in the history of skydiving, technological advancements did not stop at simply achieving a safe downward freefall. Wingsuits became another step towards the evolution of falling through the air.

The existence of wingsuits has forever changed the concept of freefalling. First and foremost, a wingsuit flight “is an extension of freefall skydiving that [has an] increased time in the air and/or allows longer distances to be travelled” (Haussmann, 2014). The longer flight and displacement, the change in distance (the length covered in two points), are due to the wingsuit’s unique capability of manipulating air drag. The skydiver manipulates air drag by taking advantage of the “increased surface area of the suit” (Haussmann, 2014). Below is a view of a wingsuit:

![Figure 2](image)

Though the mechanics behind wingsuits are much more complicated, the basic purpose of a wingsuit can be concluded as slowing down the descent. The existence of wingsuits
changed the concept of freefalling because wingsuits proved that falling down both vertically and horizontally is possible. Accordingly, the evolution of wingsuits is substantial to understanding what wingsuits are.

The history of wingsuits is just as remarkable as the history of skydiving. First and foremost, as skydiving has its ancient history, the concept of a wingsuit started “in the 11th century A.D., Eilmer of Malmesbury leapt from a monastery tower with a winged apparatus and flew more than 200 meters before falling to the earth and breaking both of his legs” (Brasfield 2008). As amateur as it started, it was the beginning of wingsuits. The idea of flying like a bird further went into the 20th century when “Clem Sohn designed, built, and successfully flew a wing suit” (Brasfield 2008). Sohn’s suit resembled a bat’s wing, and “in his new outfit from an airplane at 12,000ft, the first bird man made several loops and at 2,000ft, he folded his wings and deployed his parachute” (Brasfield, 2008). Sohn’s successful dive attracted many future daredevils. Sohn’s legacy was in 1935, and only 60 years later, “the modern wingsuit was first designed and flown by Patrick de Gayardon” (Brasfield, 2008). Patrick de Gayardon may have set the stage for the first person who developed and successfully flew a wingsuit, but it in the near future, a daredevil set a different record. It was Jhonathan Florez who beat four world records. Florez was able to get the longest total distance traveled in freefall, greatest distance flown in a wingsuit, longest freefall time and the highest altitude wingsuit jump all at the same time. For the longest total distance traveled, Florez “achieved 17.52 miles” (Park, 2014). In the greatest distance flown with a wingsuit, he achieved 16.31 miles. Thirdly, “Florez flew for 9 minutes 6 seconds” (Park, 2014). Lastly, Florez “achieved 37, 265ft” (Park, 2014) as his highest altitude jump. In summary, like skydiving, the concept of a
wingsuit began in ancient times and was successfully accomplished in the 20th century, and Jhonathan Florez took the concept of flying one step further by setting a world record of longest total distance, greatest distance, longest time and highest altitude. It was through technological advancements of wingsuits that helped Florez overcome the risks involved in his jump.

In the course of history, technological advancements of wingsuits progressed but the risk factors remained stagnant. To begin with the technological advancements of wingsuits, the modern wingsuit design evolved from a mono-surface wingsuit to a wingsuit that is “multi-layered with cells” (Brasfield, 2008). This principle worked because it made the wingsuit like a ram air parachute, and so gliding in the air became approachable. Though the designs of current wingsuits vary, the principle of having “three wings between the arms and legs [constructed in a way] to form an airfoil, generating lift like an airplane wing” (Haussmann, 2014). Secondly, unlike skydiving, wingsuit divers have much more responsibility hence a greater number of risks. To clarify, many wingsuit divers perform base jumping which is basically jumping off a fixed structure. Below is a photo of a base jumper using a wingsuit:

Figure 3: Wingsuit divers whom jumped off a mountain. (Memmott)
With further ado, it is reported that an estimation of “6 to 7 percent of the 3,000 base jumpers die each year” (Memmott, 2014). Considering the amount of deaths for skydiving, the percent of wingsuit base jumpers is remotely high. The risks involved in wingsuit diving include hitting earth’s surface while in flight (since many wingsuit divers jump from high grounds rather than airplanes), “no airflow to stabilize body position, less aerodynamic control, loss of body position and mis calculation of drop zone” (Mei-Dan, 2014). Though wingsuit technology greatly improved, the risk factors of wingsuit jumping are the same because these factors were taken account of from the time of Gayardon. To sum it up, wingsuits advanced over the course of years because the wingsuit material became multi-layered, and the risks involved in base jumping are mainly a loss of control during flight. Taking to account of the risk factors of skydiving and wingsuit diving, what exactly allows a person to safely land?

The basic physical mechanics applied to a skydiver’s gear and body position is what allows the person to safely land. Beginning with a skydiver performing a jump solely, he looks at the ground 13,000ft away from where he is standing with his parachute, the fabric device used to slow down the fall, inside a bag attached to his back. As he leaps towards the ground, he jumps in a eagle position to increase the cross-sectional area of his body. To rephrase it, the cross-sectional area in this case is the collision between his body and the air resistance (the opposition of the atmosphere to forward movement (Serway and Vuille, 2014)). Furthermore, he dives in a eagle position because this position creates more area for air resistance and so more opposing force, the push or pull of an object, towards gravity. Gravity in this case is the field force. A field force is “the push or pull of objects without physical contact” (Serway and Vuille, 2014).
Additionally, his body position also allows the skydiver to have dynamic stability. For this instance, dynamic stability is the skydiver’s tendency to keep his body position constant, and it is critical for him to do so because there is a possibility of his body tumbling into an uncontrollable manner. Though cross-sectional area and dynamic stability demonstrates how the skydiver manipulates his body posture to achieve solidity, there are still natural forces being applied in his freefall. The traditional equation of \( v_{yf} = v_{yi} - gt \) cannot be used to find the velocity of the skydiver because there is air resistance in the skydiver’s fall. Newton’s third law states that for every action there is an equal and opposite reaction, and so this law presents why the upward air resistance is what counter balances the downward acceleration. That is to say, when the skydiver first went in freefall, gravity was greater than air resistance, but as the skydiver accelerates, air resistance increases as well. This phenomenon continues until air resistance is equal to the downward velocity. In like manner, when air resistance is equal to downward velocity, there is no more acceleration and the skydiver reached terminal velocity (the greatest speed at which a body falls through the atmosphere (Skydiving Glossary, 2014)).

When he reaches terminal velocity, he remains at that speed (average of 125mph in eagle position) because an object in motion stays in motion unless an outside force acts upon it (Newton’s 1st law). To further the complexity of his freefall, the moment when the skydiver’s downward acceleration is greater than air resistance, he unclipped a pin attached to his back to release a pilot chute. A pilot chute is a “small parachute used to initiate deployment of the main parachute” (Skydiving Glossary, 2014). Now as the skydiver is blissfully falling closer to earth, he checks his pressure altimeter, a device that measures atmospheric pressure to determine the altitude (Guo, 2014), to see whether or
not he should activate his parachute. He notices that the altimeter reads that he is about to hit 2,500ft. 2,500ft is the standard altitude skydivers deploy their parachute because it is the safest altitude to anticipate a parachute failure. Once when he reaches 2,500ft, he decides to release his parachute which comes out of his deployment bag. Recalling that a parachute is “used as a deceleration device [which] must decelerate the payload to a survivable velocity before ground impact” (Accorsi, Benney and Kalro, 1998). Considering the skydiver was just falling over a hundred miles per hour, after the parachute is deployed, the skydiver will feel a break in his momentum, the force or speed of movement. This break is often referred to as a jerk because the skydiver momentarily goes upward. The reason he goes upward is because the parachute’s net force is upward, and so the skydiver accelerates upward (Newton’s 2nd law). Furthermore, the “inflation of the canopy begins as soon as the parachute is pulled free from the deployment bag” (Peterson, 2014). Air is what causes the canopy to inflate and this captured “air causes pressure inside the canopy to increase above the pressure outside of the canopy” (Peterson, 2014). The upward drag becomes greater than the downward gravity, and thus the skydiver slows down. Drag is what slowed down the skydiver, but air is not the only factor in drag. The drag “force depends on many factors including the density and viscosity of the fluid (in this case air), and the geometry, surface material, surface regularity, and velocity of the body.” (Long and Weiss, 2014). He is in a relatively low velocity, but to ensure that there is no damage to his knees, the skydiver bends his knees and points his foot upwards for a safe landing. In conclusion, a skydiver manipulates his body to ensure his safety, and the fundamental laws of physics explain why the skydiver’s parachute provides a safe landing. Though the mechanics behind skydiving is
similar to when a person is wearing a wingsuit, it is more complex to successfully complete a base jump.

The basic physical mechanics applied to a base jumper’s gear and body position is what allows the person to safely travel. Picture a skydiver wearing a wingsuit on top of a cliff 12,000ft away from the ground. She can vaguely see her landing base. She takes the brave jump with her arms and legs spread out, so the wingsuit can have a greater amount of surface area. The equation of terminal velocity can mathematically prove why she started her jump with her arms and legs spread out. Below is the equation for terminal velocity. (Villari 1)

\[ v_t = \frac{2mg}{\sqrt{C_pA}} \]  \hspace{1cm} [1]

where \( v \)=terminal velocity, \( mg \)=weight, \( C \)=drag coefficient, \( \rho \)=air density, and \( A \)=surface area

As shown in the equation 1, the surface area is in the bottom of the fraction. This means that a greater surface area would yield a lower terminal velocity. Conversely, a lower surface area would yield a higher terminal velocity. The base jumper begins with a higher surface area because an increased surface area slows down the fall. Moreover, it is when “the air on the top of the wing suit flows faster than the airflow on the bottom of the flyer, lift is produced (Brasfield, 2008).” Lift is responsible for her ability to glide. Additionally, “drag is responsible for movement” (Andrews, 2014). Drag is what slows her forward motion; “It is the frictional force created when air moves through the surface of the suit” (Andrews, 2014). The way drag works is that when there is an “increase in drag, there comes less horizontal distance covered” (Brasfield, 2008). Where as when there is an “increase in lift, [there] comes a lengthened time aloft” (Brasfield, 2008). The relationship between the base jumper’s drag and lift is often expressed as the Glide Ratio.
A Glide Ratio is “the ratio of the forward movement compared to the downward movement in the air; the ratio of lift to drag” (Leblanc, 2014). Furthermore, she is “using the aerodynamic force of drag [so she can] control [her] flight direction and distance” (Villari, 2014). As she flies in the air, depending on the surrounding environment, she manipulates her body to either gain speed or lose speed (average rate of 50 to 60 mph). As she approaches her base, she spreads her body to slow down, and deploys her parachute to safely land. In summary, base jumpers manipulate their body by either increasing the surface area of the suit or decreasing it, and the gear (wingsuit) allows the person to safely travel because the wingsuit provides lift and drag for the base jumper.

To sum it up, both skydiving and wingsuit diving have a unique history, and both sports involve similar concepts to make them possible. Skydiving started in ancient China and a wingsuit was developed as early as the 11th century. Furthermore, the 20th century served as the prime time for innovations for both sports because it was in the 1900's, parachuting became more reliable, attaining high altitudes became easier and wingsuits transformed into genuine material for human flight. These innovations have also led to different methods of skydiving and type. Skydivers can do a Accelerated Freefall, Static Line, Tandem Jump, and lastly a group formation skydive. A group formation skydive is normally competitive. Moreover, with the competition of sports comes Felix Baumgartner, the man whom holds world record of the highest altitude skydive, and Jhonathan Florez, the man whom holds the world record of greatest distance covered using a wingsuit. Furthermore, to make each sport possible, equipment provides a safe landing and body manipulation further incorporates a safe freefall. In skydiving, a parachute is used to decelerate the skydiver by increasing drag. Similarly, wingsuit divers
not only use a parachute to safely land, but use their wingsuit to provide a lift to increase the flight time length. Moreover, skydivers manipulate their body to increase the cross-sectional area, and wingsuit divers manipulate their body and wingsuit to increase their surface area. In both cases, when the area is increased, the velocity is decreased. Though the physical concepts prevail a safe landing, the risk factors for each sport usually involve a human error. With the risk factors kept in mind, people are willing to still freefall at high altitudes because of their surrounding environment. In my opinion, the feeling of skydiving is just as phenomenal as its history and physics. I believe that wingsuits are a fascinating idea, but there are too many risks involved in base jumping. In the future, I predict skydiving will grow in availability because there is an increasing number of people interested in skydiving. As for technology, though there is always room for improvement, I do not think there will be major technological improvements because the current technology is not significantly lacking anything. In short, skydiving is no longer a horrifying thrill because technology and physics has proved it to be a safe sport.
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The Physics of the Electric and Magnetic Phenomena
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Physics 112 – 35884
Dr. Casey Durandet
Abstract: The main purpose of the paper is to have an in depth understanding of electricity and magnetism. First, the paper discusses the history and the facts about electricity. Second, the paper goes into the history and facts about magnetism. Lastly, the paper briefly goes over about the history and realities of electromagnetism.

It is the reason for the great technological movement of the 21st century. Its applications are used on a day-to-day basis. What is this form of energy? It is electricity. Electricity is defined as a form of energy from the existence of charged particles. The charged particles are either positive or negative (protons or electrons). Moreover, the phenomenon of magnetism is also applied on a day-to-day basis. Magnetism is defined as either an attractive or repulsive force between objects due to an electric charge. To thoroughly understand the strength of electricity and magnets, it is vital to first be cognizant of where and when they were discovered.

The history of electricity is quite astonishing. Because electricity is a natural form of energy, the main focus is on who discovered electricity. The furthest evidence of humans observing the effects of electricity is with the ancient Greeks. It was “about 600 BC, the ancient Greeks discovered that rubbing fur on amber caused an attraction between the two” (Atkinson, 2014). Moreover, ancient batteries were discovered in the ground near Baghdad. Figure 1 illustrates the design of the discovered batteries in ancient Persia. Fast forwarding to the 17th century, it was “in the year 1600, English Physician William Gilbert used the Latin word “electricus” to describe the force that certain substances exert when rubbed against each other” (Atkinson, 2014). Though Benjamin Franklin is largely accredited with the discovery of electricity, it was “Italian physicist Alessandro Volta [whom] discovered that particular chemical reactions could produce electricity” (Atkinson, 2014). In “1800, he constructed an early electric battery that produced a steady current” (Atinkson, 2014). To clarify, current is defined as the “rate at which charge flows through its surface” (Serway and Vuille, 2011). Volta ‘s battery was only the beginning. Michael Faraday made electricity more accessible when he “created the electric dynamo (a crude power generator)” (Atinkson, 2014). This invention “solved the problem of generating an electric current in an ongoing and practical way” (Atinkson, 2014). Faraday’s solution helped Thomas Edison invent his famous light bulb because it was the “first practical bulb that would light for hours on end” (Atinkson, 2014). Edison is well known for creating a direct current (DC) system, but it was Nikola Tesla who designed an alternating current (AC) system. Additionally, Tesla was an “important contributor to the birth of commercial electricity” (Atinkson, 2014). To explain what exactly Tesla and Edison did, it is important to know that in a DC system, electrons flow only in one direction. Where as in an AC system, the electric current reverses its direction many times during the time frame. Over time, it was proven that Tesla’s AC system was superior to Edison’s DC system. To conclude, the use of electricity came from no one person, but actually a combination of quite a few intellectuals. Nonetheless, the reality of how electricity behaves is just as important as its history.

Though there is still much to know about the nature of electricity, there is a basic understanding of its essence. As mentioned previously, electricity is a form of energy from the existence of charged particles. To thoroughly understand electricity, it is imperative to first analyze electric force. Coulomb’s law summarizes the electric force between two objects. Coulomb’s law is written as $F = k_e \frac{|q_1||q_2|}{r^2}$, where $k_e$ is $8.9875 \times 10^9$ N x m$^2$/C$^2$. As shown in the equation, the separation distance of the charged particles is on the bottom of the fraction. This means that a greater distance would yield a lower force. Furthermore, Electricity is classified in two different ways: static electricity or current electricity. In static electricity, there
is a transfer of electrons between two objects. Electrons transfer usually “when you rub things
together” (Woodford, 2014). The process of static electricity involves one object gaining
electrons and the other object losing electrons. Since electrons are in the outer part of an atom,
they are easier to transfer. The object that gains electrons becomes negative, and the object that
loses electrons becomes positive. The two objects come together because opposites attract. An
example of static electricity is lightening. Lightening is caused when “rain clouds move through
the sky, they rub against the air around them. This makes them build up a huge electric charge.
When the charge is big enough, it leaps to Earth as a bolt of lightning” (Woodford, 2014).
Furthermore, current electricity is “when electrons move, they carry electrical energy from one
place to another” (Woodford, 2014). An example of current electricity is electric energy flowing
from a battery to turn on a flashlight. It is important to know that “there are two processes of
transferring charges. One is conduction, involves contact between two objects. The other is
induction, it requires no contact.” (Serway and Vuille, 2011). Moreover, there is another facet
with electricity called the electric field. The electric field is defined as “being present in any
region where a charged object experiences an electric force” (Wagon, 1999). In other words, to
see if an electric field exists, a physicist must place a test charge in the area to see if it
experiences any force. For this reason, it is mathematically defined as $E = \frac{F}{q_{\text{test}}}$. Using the
equation, it is known that the units for “field intensity for an electric field is measured in
Newtons per Colomb” (Wagon, 1999). Electric fields are illustrated with a charge and a force
vector. For example, the longer the force vector, the stronger the electric field. Furthermore,
there are many ways electricity is produced in the world. For instance, there are wind turbines
and solar power. With wind turbines, “the wind turns the propeller, which spins the generator
inside, and makes a steady current of electricity” (Woodford, 2014). As for solar cells, “when
light falls on a solar cell, the material it is made from (silicon) captures the light's energy and
turns it directly into electricity” (Woodford, 2014). Moreover, there are numerous practical
applications that come from produced electricity. For example, in a common household, there
would be no power for a television set, no power to turn on light bulbs, no power for a
refrigerator, and no power for air conditioning without electricity. Speaking of power, it is
defined as “the rate at which energy is supplied to the device” (Serway and Vuille, 2011). Not to
mention, the typical voltage of a refrigerator is around 120V. To clarify, voltage is “the kinetic
energy that an electron gains when accelerated through a potential difference of 1V” (Serway
and Vuille, 2011). Mathematically speaking, in a circuit with resistors, voltage is the electric
current multiplied by resistance or $V = IR$. Additionally, a Multimeter can be used to measure
voltage. Furthermore, outside a household, there are other ways to efficiently use electricity. For
instance, there is an electrostatic precipitator and a Xerography. An electrostatic precipitator
“removes air pollution by removing particulate matter from combustion gases” (Serway and
Vuille, 2011). Though an electrostatic precipitator is not widely used for commercial use, the
precipitator is mainly used in industrial situations that give off much pollution. A major
advantage of the electrostatic precipitator is that the filters do not need be changed, only washed.
Additionally, the precipitator manipulates the idea that opposite charges attract. The process of a
precipitator is depicted in Figure 2. Furthermore, a Xerography is “used to make photocopies of
printed materials” (Serway and Vuille, 2011). On the subject of applications using electricity,
what happens when electricity becomes out of hand? Specifically, how does the human body
react to electricity? The reaction of the human body to electric current is called electric shock.
Electric shock takes place when there is “contact of a human body with any source of voltage
high enough to cause sufficient current through the skin, muscles or hair” (Prasad, Sharma,
Furthermore, “the minimum current a human can feel is thought to be about 1(mA). The current may cause tissue damage or fibrillation if it is sufficiently high” (Prasad, Sharma, 2010). The human body responds quite negatively to a small amount of current. Where as when currents approach “100 (ma) [they are] lethal if they pass through sensitive portions of the body” (Prasad, Sharma, 2010). Though the numerical values do sound intimidating, the “effects of an electric current passing through the vital parts of a human body depend on the duration, magnitude and frequency of this current” (Prasad, Sharma, 2010). Not to mention, just like electricity is based upon charged particles, there is a force that also has a dualistic nature.

To fully account for the magnetic force, it is only reasonable to begin with the history of magnetism. Just like electricity, the history of magnetism consists of discoveries because magnetism is a natural phenomenon. As far back as known, the ancient Chinese brought forward the first use of magnets. The Chinese figured out “a steel needle stroked with such a lodestone became magnetic, and around 1000, the Chinese found that such a needle, when freely suspended, pointed north-south” (Stern, 2001). Furthermore, the “magnetic compass soon spread to Europe. Columbus used it when he crossed the Atlantic Ocean” (Stern, 2001). In those days, the compass provided direction, but it left an unanswered question, what causes the needle to move? It was “around 1600, William Gilbert, physician to Queen Elizabeth I of England, proposed an explanation: the Earth itself was a giant magnet, with its magnetic poles some distance away from its geographic ones” (Stern, 2001). Gilbert’s radical proposition gave birth to the modern idea that Earth has a magnetosphere. On ground, the magnetosphere provides an explanation for “magnetic storms and the polar aurora or northern lights” (Stern, 2001). On the other hand, satellites in space are able to detect the “radiation belts, magnetic structures, fast streaming particles and processes which energize them” (Stern, 2001). Undoubtedly, the earth’s magnetic field is crucial for life because it protects living organisms from the sun’s radiation. For that reason, an explanation for the origin of earth’s magnetic field is necessary. The most current theory lies with Earth’s core. Earth’s core is made up of liquids and “it is widely believed this core is made up of molten iron, perhaps mixed with nickel and sulfur” (Stern, 2008). The theory states that because the “molten metal is believed to be circulating, it creates a system of electric currents, spread out through the core” (Stern, 2008). This electric current is thought to “create a magnetic field, a distribution of magnetic forces” (Stern, 2008). Ironically, “the geographic North Pole of Earth corresponds to a magnetic south pole, and the geographic South Pole of Earth corresponds to a magnetic north pole” (Serway and Vuille, 2011). Moreover, when studying earth’s magnetic field, it is necessary to grasp how magnetism is defined. In the past, it was thought iron magnets only produced magnetism. Only until 1821, a “Danish scientist, Hans Christian Oersted, while demonstrating to friends the flow of an electric current in a wire, noticed that the current caused a nearby compass needle to move” (Stern, 2001). Later, Andre Marie Ampere studied this phenomenon and concluded, “the nature of magnetism was quite different from what everyone had believed. It was basically a force between electric currents: two parallel currents in the same direction attract, in opposite directions repel” (Stern, 2001). Though the history of magnetism is very brief, it is only suitable to apprehend magnetism in depth.

As mentioned previously, magnetism is defined as either an attractive or repulsive force between objects due to an electric charge. Moreover, in theory, “the two poles (north and south) cannot exist independently” (Serway and Vuille, 2011). For instance, if a magnetic block was sliced into atomic pieces, there would still be two opposite poles within the material. As stated before, magnetism is a force rather than an energy form. It turns out that “the magnetic force is
always directed toward the center of the circular path” (Serway and Vuille, 2011). The fact that there is a circular motion means a magnetic force causes a centripetal acceleration. In algebraic terms, the magnitude of the magnetic force is \( F = qvB\sin(\theta) \) where \( q \) is the value of the charge, \( v \) is the velocity, \( B \) is the magnetic field and \( \theta \) is the angle between the direction of the velocity and the magnetic field. Additionally, the maximum amount of force happens when the particle’s motion is perpendicular to the field, where as the force is equal to zero when the particle’s motion is parallel to the field. The force equation best illustrates why the maximum force happens at a perpendicular angle and the minimum force happens at a parallel field. Looking at the formula, when the angle is perpendicular (90 degrees), the \( \sin \) of 90 is equal to 1. The \( \sin \) of any other degree would yield a smaller value. Also, when the angle is parallel (180 degrees), the \( \sin \) of 180 degrees is zero; hence the value of the force is zero. Moreover, similar to electricity, there are two kinds of magnetic materials. There are soft magnetic materials and hard magnetic materials. With soft magnetic materials, it is “easily magnetized but tends to lose its magnetization easily” (Serway and Vuille, 2011). In contrast, “hard magnetic materials are used in permanent magnets” (Serway and Vuille, 2011). An example of soft magnetic material would be iron. When it comes to magnetic domains, the soft magnetic materials are said to be ferromagnetic because “the magnetic fields produced by the electron spins don’t cancel completely” (Serway and Vuille, 2011). An example of hard magnetic material would be magnets in computer hard drives or loudspeakers. Moreover, the hard magnetic materials are considered a permanent magnet because the “domains remain aligned even after the external field is removed” (Serway and Vuille, 2011). Certainly, the concept of a magnetic field is just as important as individual magnets. Magnetic fields are simply an area of space that undergoes magnetic influence. To illustrate, magnetic field lines are drawn to explain the interaction. Figure 3 is an example of how magnetic field lines are drawn. The figures show that the field lines connect with each other in a positive/negative situation, but repel each other in a positive/positive situation. Another way to tell the direction of the magnetic field is with x’s and dots. When there are dots, the magnetic field is pointing towards the observer on the paper. On the other hand, when there are x’s, the magnetic field is pointing into the paper. Figure 4 best illustrates how the symbols show the direction of the magnetic field. Furthermore, the concept of magnetic field lines came from an English physicist named Michael Faraday. Though Faraday called it magnetic lines of force, the same notion of using the lines to understand the direction and strength applies. Looking at the line, the strength of the magnetic field depends upon how close the lines are to one another. The closer the lines are together, the greater the strength of the field. Conversely, the farther away the lines are from each other, the weaker the strength of the field is. Speaking of strength, there is a mathematical expression for the strength of the magnetic field in a straight wire. The formula is \( B = \frac{\mu_0 I}{2\pi r} \). This equation reinforces the previous fact that the magnetic force moves in a circular path because the circumference of a circle is in the equation. Granted that magnetic forces have a circular behavior, magnetic materials can also be broken down to how they react. That is to say, there are paramagnetic materials and also diamagnetic materials. Paramagnetic materials have “magnetic moments that tend to align with an externally applied magnetic field” (Serway and Vuille, 2011). This definition is similar to ferromagnetic materials, but the difference lies in the strength of the response. The response of paramagnetic materials “is extremely weak compared with that of ferromagnetic materials” (Serway and Vuille, 2011). Examples of paramagnetic materials are platinum, calcium and aluminum. On the other hand, in diamagnetic materials, “an externally applied magnetic field induces a very weak magnetization that is opposite the applied field” (Serway and Vuille, 2011).
An everyday example of diamagnetic materials is water and air. Before understanding how magnetism can be applied to technology, it is key to first know the source of a magnetic field. The source of a magnetic field is an electric current. The electrons that carry an electric charge produce a magnetic field. When it comes to practical applications, magnets are used “in meters, motors, loudspeakers, and magnetic tapes and disks” (Serway and Vuille, 2011). In terms of medicinal uses, doctors use magnetic resonance imaging or MRI devices to get a better visual interpretation of the human body. Furthermore, doctors can now “stimulate brain tissue using external magnetic fields and injected magnetic nanoparticles” (Chandler, 2015). This type of brain therapy has shown improvement in patients with Alzheimer’s disease, depression and epilepsy. Magnetism and electricity both affect the human body. That is because both concepts are closely related.

Electromagnetism is the interaction between magnetic fields, electric fields or currents. Michael Faraday was the physicist who brought this phenomenon to the scientific world. It turns out that a current is produced when there is a change in a magnetic field. When looking at a circular electric wire connected to an ammeter, a magnet moving towards the wire causes a positive current. Where as when the magnet is not moving, there is no current. Moreover, the ammeter shows a negative current when the magnet moves away from the electric wire. Alternatively, when the wire is moved and the magnet is held stationary, a current is also produced. All of the following scenarios fall under electromagnetic induction. Electromagnetism brings the concept of a magnetic flux. Magnetic flux is the number of magnetic field lines that pass through a surface. The formula for a magnetic flux is \( \Phi_B = BA \). When looking at an illustration of magnetic field lines, the magnetic flux is calculated quite easily because “the value of the magnetic flux is proportional to the total number of lines passing through the loop” (Serway and Vuille, 2011). There would be no induced current if there were no change in magnetic flux. The produced current is called an induced current “because it is produced by an induced emf [or voltage]” (Serway and Vuille, 2011). The concept of a magnetic flux is crucial in calculating the induced voltage. Faraday’s law mathematically defines emf as \( \varepsilon = -N \frac{\Delta \Phi}{\Delta t} \) where \( \Delta t \) is the time interval and \( N \) is the number of the circuit’s loops. Furthermore, there is a negative sign in the formula because it “indicates the polarity of the induced emf” (Serway and Vuille, 2011). Moreover, Lenz’s law explains the direction of the currents flow. Lenz’s law states “the current caused by the induced emf travels in the direction that creates a magnetic field with flux opposing the change in the original flux through the circuit” (Serway and Vuille, 2011). Lastly, electromagnetism is of utmost importance because of its applications. It turns out that “generators and motors are important practical devices that operate on the principle of electromagnetic induction” (Serway and Vuille, 2011). Generators produce electricity, and by inference, electromagnetism is responsible for every major technological system.

In summary, electricity and magnetism are two very closely related topics. In terms of the history of electricity, the observance of the effects of electricity began with the Ancient Greeks. Real progress began with an Italian physicist Alessandro Volta. He discovered that a particular chemical reaction could produce electricity. Moreover, Michael Faraday, Thomas Edison, and Nikola Tesla all contributed to the development of using the phenomenon of electricity. Key concepts of electricity include its definition, classifications, and the electric field. To define, electricity is a form of energy from the existence of charged particles. Moreover, electricity is classified as static and current. Lastly, the electric field is present in any region where a charged object experiences an electric force. Moving on to magnetism, the history of magnetism is just as brief as electricity. The ancient Chinese were the first to pave the way for the use of magnets. To
put it in another way, the Chinese were responsible for launching the use of the compass because they figured out that a needle coated lodestone moved north or south. This discovery led itself into Europe and opened the doors of uncovering the mystery of earth’s magnetic field. When it comes to key points, its important to reconcile the definition of magnetism, its natural qualities, the magnetic field, and magnetic materials. To begin with it’s definition, magnetism is defined as either an attractive or repulsive force between objects due to an electric charge. When it comes to the natural qualities of magnets, the magnetic force always follows a circular path. Moreover, magnetic fields are simply an area of space that undergoes magnetic influence. Finally, magnetic materials are categorized as paramagnetic or diamagnetic materials. As for electromagnetism, its history generally began with Michael Faraday. He studied the phenomenon of an electric current becoming produced by a change in a magnetic field. Key points related to electromagnetism are magnetic flux, induced current, and induction. Magnetic flux is the number of magnetic field lines that pass through a surface. Secondly, an induced current is a current produced by an induced voltage. Lastly, induction, in summary, is the different ways a current can be produced by either changing the magnetic field near an electric wire or a change in current near a magnetic field. When it comes to my personal thoughts regarding the subject of electricity and magnetism, I think the science behind the subjects is revolutionary to human progress. Understanding the physics behind particular objects at a micro-level changes the way humans interpret the world around them. Specifically with electricity and magnetism, the knowledge behind them expanded their applications. In terms of future progress, I believe scientists will one-day use the power of electricity/magnetism to their full potential. With electricity, technology will utilize electricity with minimal or no waste. Where as with magnetism, regardless the weakness of the field strength, I believe scientists will figure out how to efficiently manipulate earth’s magnetic field to generate an abundant amount of power. The use of electricity and magnetism truly changed the course of history. In addition, it is highly unlikely there would even be a Physics’ research paper without electricity.
Figures

Below is a figure illustrating why “ancient Persians may have also used an early form of batteries” (Atkinson, 2014).

Figure 1: A diagram and replica of a battery found by Archeologists near Baghdad (Atkinson, 2014).

Below is a diagram demonstrating the process of how the electrostatic precipitator removes air pollution.

Figure 2: A diagram illustrating to steps to cleaning air polluted material (BBC).

Below is a figure of how field lines are drawn to explain an attractive and repulsive interaction.

Figure 3: An illustration of how field lines are drawn with repulsive and attractive magnets. (Magnet Field Basics)
Below are figures illustrating the symbols used to determine if the magnetic field is pointing in or out.

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Figure 4: The left illustration is the magnetic field pointing towards the paper. The right illustration is the magnetic field pointing towards the observer. (Irregular)
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Electricity and its Physical Interaction with the Human Brain

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Physics 112 – College Physics 2
Class #: 35844
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Abstract: The paper first elaborates the phenomenon of electricity and magnetism by noting its principals. At second, the brain model is detailed. Furthermore, technological advancements that led to brain stimulation are explained. Also, brain stimulation is detailed by reporting its history and principals. Lastly, the implications of quantum and classical physics towards psychology are noted.

The fact that most of present technology is reliant on electricity makes it seem as if humans are dependent on electricity. But, what if human life depends on electricity regardless of technology? In fact, survival would not be possible without electricity. Like a computer, the brain communicates with itself using electrical signals. To complicate the matter further, it is known that the “human brain is composed of about 100 billion nerve cells (neurons) interconnected by trillions of connections, called synapses” (Dougherty, 2011). To comprehend how neurons are interconnected, the fundamental concern is how neurons communicate. Whenever something may trigger the brain to send an electrical signal, whether it be eating, walking, or simply seeing, that same electrical signal “propagates like a wave along the long threads called axons that are part of the connections between neurons” (Dougherty, 2011). Furthermore, “when the signal reaches the end of an axon, it causes a release of chemical neurotransmitters into the synapse, a chemical junction between the axon tip and target neurons” (Dougherty, 2011). The process of brain communication is still not over. That same target neuron ends up responding with “its own electrical signal, which, in turn, spreads to other neurons” (Dougherty, 2011). This entire process happens in less than a second. Having the knowledge of how the brain functions not only favors technological advancements, but also medical progress. Besides the complicity of the biology and chemistry of the brain, the physics of the brain may be a simpler approach in medical applications. Dr. Francisco Ponce, a neurosurgeon and a neuromodulator at the Barrow Neurological Institute, describes how physics contributed to brain stimulation. He said scientists came from an “anatomical approach to a circuit based approach of the brain.” Instead of approaching the brain as a complex interaction between chemicals, the circuit-based approach simplifies the perspective because it applies straightforward laws of nature. To thoroughly elaborate the medical practice of brain stimulation, it is substantial to be informed of what exactly is electricity.
Although the nature of electricity in its entirety is unknown, there is an understanding of its basic principles. Electricity is defined as a form of energy from the existence of charged particles. Beginning with a definition of electric force, Coulomb’s law summarizes the electric force between two objects as \( F = k_e \frac{|q_1| |q_2|}{r^2} \), where \( k_e \) is \( 8.9875 \times 10^9 \text{ N m}^2/\text{C}^2 \). Moreover, electricity is classified in two different ways: static electricity or current electricity. In static electricity, there is a transfer of electrons between two objects. Electrons transfer “when you rub things together” (Woodford, 2014). The process of static electricity involves one object gaining electrons and the other losing electrons. In addition, the object that gains electrons becomes negative, and the object that loses electrons becomes positive. The two objects come together because opposites attract. An example of static electricity is the zap people feel when touching a doorknob. Furthermore, current electricity is “when electrons move, they carry electrical energy from one place to another” (Woodford, 2014). An example of current electricity is the process of turning on a car. It is important to know “there are two processes of transferring charges. One is conduction, involves contact between two objects. The other is induction, it requires no contact.” (Serway and Vuille, 2011). Moreover, another feature with electricity is the electric field. The electric field is defined as “being present in any region where a charged object experiences an electric force” (Wagon, 1999). In other words, to see if an electric field exists, a physicist must place a test charge in the area to see if it experiences any force. It is mathematically defined as \( E = \frac{F}{q_{\text{test}}} \). Moreover, there are numerous practical applications that come from produced electricity. For example, generally a household needs electricity to power a television set, to turn on light bulbs, and for air conditioning. Power is defined as “the rate at which energy is supplied to the device” (Serway and Vuille, 2011). Not to mention, the typical voltage of an air conditioner is around 220V. Voltage is “the kinetic energy that an electron gains when accelerated through a potential difference of 1V” (Serway and Vuille, 2011). Mathematically, in a circuit with resistors, voltage is \( V = IR \).

Referring back to the applications using electricity, what happens when electricity becomes out of hand? Specifically, how does the human body react to electricity? The reaction of the human body to electric current is called electric shock. Electric shock takes place when the “human body comes into contact with any source of voltage high enough to cause sufficient current through the skin, muscles or hair” (Prasad, Sharma, 2010). Furthermore, “the minimum
current a human can feel is thought to be about 1(mA). The current may cause tissue damage or fibrillation if it is sufficiently high” (Prasad, Sharma, 2010). The human body responds quite negatively to a small amount of current. Where as when currents approach “100 (ma) [they are] lethal if they pass through sensitive portions of the body” (Prasad, Sharma, 2010). Though the numerical values do sound intimidating, the “effects of an electric current passing through the vital parts of a human body depend on the duration, magnitude and frequency of this current” (Prasad, Sharma, 2010). Moreover, it is not only electricity that affects the body, but also the force of magnetism.

Magnetism is defined as either an attractive or repulsive force between objects due to an electric charge. To clarify, magnetism is a force rather than energy. It turns out that “the magnetic force is always directed toward the center of the circular path” (Serway and Vuille, 2011). Moreover, similar to electricity, there are two kinds of magnetic materials. There are soft magnetic materials and hard magnetic materials. With soft magnetic materials, it is “easily magnetized but tends to lose its magnetization easily” (Serway and Vuille, 2011). In contrast, “hard magnetic materials are used in permanent magnets” (Serway and Vuille, 2011). An example of soft magnetic material would be iron. When it comes to magnetic domains, the soft magnetic materials are said to be ferromagnetic because “the magnetic fields produced by the electron spins don’t cancel completely” (Serway and Vuille, 2011). Where as the domain of hard magnetic a material is considered a permanent magnet because it “remained aligned even after the external field is removed” (Serway and Vuille, 2011). By all means, the concept of a magnetic field is just as important as individual magnets. Magnetic fields are simply an area of space that undergoes magnetic influence. Furthermore, there is a mathematical expression of the strength of the magnetic field in a straight wire. The formula is \[ B = \frac{\mu_0 I}{2\pi r} \]. Even more, magnetic materials can also be broken down by how they react. That is to say, there are paramagnetic materials and also diamagnetic materials. Paramagnetic materials have “magnetic moments that tend to align with an externally applied magnetic field” (Serway and Vuille, 2011). This definition is similar to ferromagnetic materials, but the difference is the strength of the response. The response of paramagnetic materials “is extremely weak compared with that of ferromagnetic materials” (Serway and Vuille, 2011). Examples of paramagnetic materials are platinum, calcium and aluminum. On the other hand, in diamagnetic materials, “an externally applied magnetic field induces a very weak magnetization that is opposite the applied field” (Serway and Vuille,
An everyday example of diamagnetic materials is water and air. Before understanding how magnetism can be applied to technology, it is vital to first know the source of a magnetic field. A magnetic field’s source is an electric current. The electrons that carry an electric charge produce a magnetic field. When it comes to applications, in terms of the medicinal uses of magnetism, doctors use magnetic resonance imaging or MRI devices to get a better visual interpretation of the human body. In the figure’s page, figure 1 illustrates how MRI’s detail out the brain. Furthermore, doctors can now “stimulate brain tissue using external magnetic fields and injected magnetic nanoparticles” (Chandler, 2015). They do this based off of the concept of electromagnetism. Electromagnetism is briefly a current being produced by a change in a magnetic field. Moreover, this magnetic brain therapy has shown improvement with patients with Alzheimer’s disease, depression and epilepsy. These improvements in brain therapy invoke a question, when and how did doctors start to use the force of magnetism to help patients?

Prior to making sense of the history of electric/magnetic brain stimulation, the brain model must be understood. Generally speaking, the brain is broken up into the brain stem, the midbrain, and the forebrain. In detail, the brain stem “contains the medulla oblongata. This controls breathing, heart rate, and digestion. It also contains the cerebellum, which coordinates sensory input and maintains muscle movement and balance” (Mandal, 2013). Moreover, the midbrain is the “region of the brain responsible for vision, hearing, temperature control, motor control and alertness” (Mandal, 2013). Additionally, the midbrain also includes the limbic system. Lastly, the forebrain is where the cerebral cortex is located. The cerebral cortex has four different lobes: the frontal lobe, parietal lobe, temporal lobe, and the occipital lobe. To clarify, the forebrain is “involved in thought and problem solving” (Mandal, 2013). Though the brain is much more complex, these are the basic functions of the forebrain, midbrain and brain stem. To get a better picture of the brain model, Figure 3 illustrates the different areas of the brain. With a basic understanding of how the brain functions, scientists were soon enough ready to toy around with its physical properties.

The history and evolution of brain stimulation is made up of brilliant minds. To begin with the history of electrical stimulation, it all began in the 18th century with Luigi Galvani. Galvani was the one who discovered “nerves and muscles are electrically excitable” (Sabbatini, 2004). He mainly experimented with frogs. Other scientists confirmed Galvani’s experimental outcome, and “became convinced that nerves and muscles were electrically excitable and they
also generated a kind of electricity themselves” (Sabbatini, 2004). Though, experimenting with electricity was not limited to the body. It was another Italian scientist who first “applied electrical stimulation directly to the exposed brain” (Sabbatini, 2004). Through various experiments with lesions and stimulating the surface of the central nervous system, Luigi Rolando “proved that the central nervous system was indeed electrically excitable” (Sabbatini, 2004). In the 19th century, it was “Gustav Theodor Fritsch and Julius Eduard Hitzig who carried out experiments of localized electrical stimulation of the brain cortex of several animals” (Sabbatini, 2004). In the late 19th century, an American neurologist was the first person to stimulate the human brain electrically. Moreover, Roberts Bartholow ended up experimenting with 22 more patients by 1902. Indeed, the electronic revolution in 1920 “led to the development of much more sophisticated stimulus generators” (Sabbatini, 2004). Notwithstanding, real progress began in the 1940’s. It was the “number of stereotactic devices for neurosurgery that led to a great increase in the number of clinic-experimental studies and surgical treatment methods” (Sabbatini, 2004). Furthermore, the history of brain stimulation did not stop at electricity.

The brief history of magnetic brain stimulation began with Physicists. To be clear, Michael Faraday’s discovery of an electric current created by a change in a magnetic field paved the way for neurologists. Scientists were to utilize this phenomenon to stimulate the brain. It was as early as the 1900’s when “Adrian Pollacsek and Berthold Beer filed a patent to treat depression and neuroses with an electromagnetic device” (George, Mayberg, Schlaepfer, 2010). Only 30 years later, the “development of electroconvulsive therapy ushered the field of brain stimulation into the modern ‘medical device’ age” (Forrow, Fregni, Horvath, Pascual-Leone, Perez, 2011). Furthermore, technological advancements did not stop there. Anthony Barker and his colleagues developed “the first reliable transcranial magnetic brain stimulator in 1984” (Forrow, Fregni, Horvath, Pascual-Leone, Perez, 2011). Transcranial magnetic stimulation or TMS relatively became popular in clinics to treat motor cortex disorders. Undoubtedly, the history of stimulating the brain with either magnetism or electricity largely depended upon development in technology.

As far as the technological advancements took place in the past, brain stimulation technology is still progressing. In terms of directly using electricity to stimulate the brain, Dr. Francisco Ponce explained that scientists “are developing wires that direct a low current in a specific direction.” Additionally, with magnetism, Dr. Ponce stated that researchers are
developing a “focused ultrasound. It works without cutting through the skin.” Lastly, in TMS, researchers are working with “theta-burst stimulation, a high- frequency stimulatory pattern. It can induce rapid long-term potentiation and long-term depression effects via the mimicry of powerful brain oscillations” (Forrow, Fregni, Horvath, Pascual-Leone, Perez, 2011). With the history and technological progress of brain stimulation in place, the underlying facts about each method of stimulation must be covered.

There are many realities behind stimulating the brain through the use of electricity. For instance, there is the process of electrical brain stimulation, how the brain reacts, the clinical applications, and its effectiveness. To examine the process of electrically stimulating the brain, it is important to first note that there are various methods of doing so. For example, to name a few, there is vagus nerve stimulation, deep brain stimulation, trans-cranial direct current stimulation and electroconvulsive therapy. Vagus nerve stimulation or VNS “involves intermittent repeated stimulation of the left vagus nerve with a small electrical pulse from an implanted neurostimulator” (George, Mayberg, Schlaepfer, 2010). Moreover, the neurostimulator is programmed in the doctor’s office and it can also be reprogrammed. Secondly, there is deep brain stimulation or DBS. Deep brain stimulation is a procedure where a “surgical placement of a thin wire into a very specific and carefully selected brain region” (Okun, Zeilman, 2014). In addition, the wire is “connected to a pacemaker-like device that is implanted in the chest region below the collarbone” (Okun, Zeilman, 2014). Figure 3 is an illustration of the pacemaker attached to the wire that connects to the brain. Like VNS, the doctors can control the amount of stimulation in DBS. Thirdly, trans-cranial direct current stimulation defined by Dr. Ponce is “a type of neurostimulation that uses direct electrodes to deliver a constant and low current to the brain.” Lastly, there is electroconvulsive therapy. This is the most popular method of stimulation because as Dr. Ponce remarked, “in the United States, it is the only type of brain stimulation in the market.” Electroconvulsive therapy or ECT is a “procedure in which electric currents are passed through the brain, intentionally triggering a brief seizure” (Mayo Clinic Staff, 2012). Additionally, the patients are usually under anesthetics. This method of stimulation “seems to cause changes in brain chemistry that can quickly reverse symptoms of certain mental illnesses” (Mayo Clinic Staff, 2012). In order to know where to stimulate the brain, doctors use a micro wire. A micro wire helps neurosurgeons know how far they have to go into the brain. Figure 4 is a picture of a micro wire. It is important to note the
wire in the figure is not the thick part; it is the very thin area at the end. With all of the various types of electrically stimulating the human brain, it is key to understand how the brain reacts to such a stimulus. To see how the brain reacts to electricity, viewing the brain as a circuit clarifies the process. For instance, in deep brain stimulation, when the brain is touched by electricity, the brain reacts the same way as if there was a chemical reaction. It is thought that the electric current “inhibits cell firing, which excites the axons, and releases a chemical called calcium from brain cells called astrocytes. Calcium seems to trigger a series of reactions that lead to the release of chemicals called neurotransmitters and the stimulation of blood flow” (Okun, Zeilman, 2014). Though, the reason why this process works out so well is currently unknown.

Although the reason behind the physical process is unknown, electric stimulation still has many clinical applications. For example, deep brain stimulation has treated problems such as “tremor, rigidity, Bradykinesia, Dyskinesia, Dystonia, low or high energy levels, and a general sense of well being” (Okun, Zeilman, 2014). Where as with Electroconvulsive therapy, it treats problems such as “severe depression, severe mania, and catatonia” (Mayo Clinic Staff, 2012). In terms of its benefits, it turns out electrically stimulating the brain increases learning speed. For instance, experiments show “that stimulating the motor cortex increases the speed of learning motor skills” (Battison, 2011). This technique may help athletes conquer a sport faster. Taking all of these treatments and benefits into thought, how effective is this method of healing? With the help of DBS, there was a “35 percent decrease in obsessive-compulsive symptoms and improvements in depression, anxiety and quality-of-life measures” (Patoine, 2010). Additionally, in December of 2008, data indicated, “seizure activity was reduced by a median of 38 percent in patients” (Patoine, 2010). Not to mention, the effectiveness of this type of therapy is further supported by the amount of labor doctors have to do. Firstly, as Dr. Ponce stated, in order to become a neurosurgeon, one must have “a bachelors degree, a medical degree, and six to seven years of residency.” Secondly, when stimulating the human brain, there are three steps doctors and patients must follow. For instance, Dr. Ponce affirmed that at first there must be a “good patient selection.” This is done by neurologists who examine if the patient can have deep brain stimulation. Moreover, psychologists then review the patient to see if they are ready. At the second step, the neurosurgeon must plan a good surgery. As Dr. Ponce said, the doctor prepares by “planning where to plant the electrode, so that there can be a safe trajectory.” At last, the doctor makes sure there is good programming. Good programming, as Dr. Ponce said, is “putting
the patient in the frame and calculating the direction of the drill hole.” Furthermore, doctors keep track of their patients by recording the patient’s brain fluctuations throughout the day. Figure 5 is a patient’s chart of the fluctuations throughout the day. Truly, electricity played a tremendous role in medical advancements, but so did magnetism.

Stimulating the human brain with magnetism is on another level compared to electricity. First of all, with magnetic technology, there is no need to implant anything into the human body. The change in magnetic fields stimulates a current itself. To go into depth about magnetic brain stimulation, one must understand what is the process of transcranial magnetic stimulation, the brain’s reaction to such a method, and the scientific applications. What happens in transcranial magnetic stimulation, there is a “brief intense magnetic field applied to the scalp. This field induces electrical activity in the cortex, effectively disorganizing neural processing in that region of the cortex and thus disrupting normal functioning for a few milliseconds. This effect has been termed a virtual lesion” (Leone, Walsh, 2015). TMS utilizes the phenomenon of an induced current caused by a magnetic field. To specific, it is reported that a “magnetic field (up to about 2 Tesla) is induced with lines of flux running perpendicularly to the plane of the coil” (Cogiamanian, Marceglia, Priori, Rossi, Torre, 2015). That same magnetic field “penetrates the skull, and reaches the brain (maximum range: 2-3 cm), where it induces an electric field able to excite neurons and axons 1/100000 as large as the primary current” (Cogiamanian, Marceglia, Priori, Rossi, Torre, 2015). Additionally, Figure 6 is an illustration of the specs behind a TMS device. Although doctors are able to direct the area of stimulation, its effect on the neural function is unknown. Moreover, there are cases in which TMS showed improvement in patients suffering from depression. Aside from the medical world, TMS helps researchers understand the brain. That is to say, experimenting with TMS “has been used to study perception” (Hallett, 2000). This is done by stimulating the occipital cortex and asking the subjects what they see. Furthermore, TMS can be used to map out the brain. For example, researchers used “focal TMS to perform detailed mapping of the motor cortical areas” (Hallett, 2000). Additionally, scientists have been able to “use magnets to control brain activity in mice” (Underwood, 2015). Indeed, Physicists created TMS, and so Physics played a large role in the new approach of understanding the human brain.

There are two different ways to understand the psychology of the brain: through classical and quantum physics. To begin with classical physics, it is defined as more of a deterministic
approach. This means that classical physicists believe “the interactions in the state of the physical world at any time is completely determined by the state at any earlier time” (Beauregard, Schwartz, Stapp, 2005). In other words, the outside world is not dependent on human observations. Moreover, this approach causes one to think themselves as “a mechanical automaton: every physical action was predetermined by mechanical interactions” (Beauregard, Schwartz, Stapp, 2005). When it comes to applying this reasoning to psychology, scientists would consider thoughts and emotions as simply a byproduct of brain activity. The problem with this method of reasoning is that it “if mental processes and consciousness have no effect upon the physical world, then what keeps a person's mental world aligned with their physical situation?” (Beauregard, Schwartz, Stapp, 2005). However, Physics is not limited to only focusing on the outside world. In terms of psychology, Quantum theory relates the individual to the outside world. To be clear, Quantum theory states, “conscious free choices affect the physically described world” (Beauregard, Schwartz, Stapp, 2005). With this methodology, free choice is emphasized. Moreover, the quantum model helps psychologists because it puts more responsibility to the individual. Granted that in Physics there are different approaches to understanding the psychology of the human being, the classic and the quantum approach are both relative in the world today.

All things considered, when looking at electricity and its physical interaction with the brain, it is necessary to break it into components. That is to say, one must comprehend the nature of electricity and magnetism, the model of the human brain, the history and reality behind brain stimulation, and how Physics can be used to explain Psychology. Beginning with the nature of electricity, it is defined as a form of energy from the existence of charged particles. These charged particles interact with other particles to either make an object positive or negative. As for the nature of magnetism, it is an attractive or repulsive force between objects due to an electric charge. In other words, magnetism is a force and electricity is an energy form. Moreover, the source of a magnetic field is an electric current. When it comes to the model of the human brain, the brain is broken up into the brain stem, the midbrain, and the forebrain. More importantly, each region has different functions that are critical for survival. Furthermore, the history of brain stimulation is comprised of stimulating the brain directly with electricity and indirectly with magnetism. With electrically stimulating the brain, its history began with Luigi Galvani. Galvani discovered the body can be electrically stimulated, and he also indicated the
body itself has electricity. Moreover, Luigi Rolando uncovered the fact that the brain can also be affected by electricity. Lastly, Roberts Bartholow was the first person to help patients through electric stimulation. Where as with the history of magnetically stimulating the brain, it was Adrian Pollacsek and Berthold Beer who used an electromagnetic device to treat depression. Furthermore, the devices upgraded when Anthony Barker and his colleagues developed a transcranial magnetic brain stimulator. Indeed, in order to conclude the realties of brain stimulation, it is first necessary to account for directly stimulating the brain with electricity.

The different methods of electric brain stimulation include vagus nerve stimulation, deep brain stimulation, and electroconvulsive therapy. Furthermore, the most popular method of brain stimulation is ECT. The way the brain reacts to this sort of stimulation is by a change in brain chemistry. As for magnetic brain stimulation, the main method is called transcranial magnetic stimulation. With this method, a device using a strong magnetic field induces a current on the outer region of the brain. This method has causes of which depression was treated. Lastly, Physicists explain the psychology of a person through classical physics and quantum physics. In classical physics, scientists have a brain over mind approach. To be clear, classical physicists believe thoughts and emotions are simply caused by brain activity. Where as with quantum physics, scientists believe thoughts and emotions directly affect brain activity. Speaking of physics and psychology, when it comes to my personal thought on this subject, I believe brain stimulation is truly a revolutionary method of healing. No longer will people be confined to the side effects of prescription pills. Moreover, although brain stimulation therapy comes from an outside force, it is much more natural because it does not involve a variety of different recipes. In terms of the future, I believe one-day transcranial magnetic stimulation will overcome electroconvulsive therapy. That is to say, because TMS is much more simplistic in terms of procedure, we could develop a device that goes deeper into the skull at a specified area and surgery no longer is required to electrically stimulate the brain. By the same token, electricity’s interaction with the human brain may not only benefit humanity in terms of medical progress, but also in knowledge-based progress. Indeed, electricity and brain stimulation is not limited to its benefits for patients, but also to the sustenance of the common man.
Figures

**Figure 1:** MRI pinpointing area of attention (Taken at Barrow Institute)

**Figure 2:** A model of the general areas of the brain (Mandal 2013)
Figure 3: A Pacemaker and the wire that connects to the brain (Taken at Barrow Institute)

Figure 4: Dr. Francisco Ponce holding a micro wire (Taken at Barrow Institute)
**Figure 5**: A patient diary of the brain fluctuations throughout the day.

**Figure 6**: A hypothetical illustration of a TMS device affecting a human brain (Patoine, 2010)
References


X-Ray Diffraction: A Milestone in Scientific Technology

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Abstract:
X-Ray Machines are a huge part of the medical field. Doctors use them very often in order to get a better picture of what's going on inside the body. X-ray diffraction also plays a big role in identifying unknown substances which is very important in the research of medicine, biology, engineering, and environmental sciences. This paper will explore the X-Ray machine in more depth, who discovered the concept behind it, who actually created the first X-Ray machine, the physics behind the making of the X-Ray machine. We will also look at X-ray powder diffraction and its contribute to science.

Introduction:
In 1895 a German physicist named Wilhelm Conrad Roentgen was working in his lab, in Wurzburg Germany trying to test if cathode rays could penetrate glass. As he was conducting his test he realized that there was a glow coming from a chemical coated screen that was near his work station. He named them X-Rays because he didn't know anything about them and it was his first time observing them. He continued to study these rays and develop a better understanding of them. He then was able to conclude that the X-rays were electromagnetic waves that act very similar to light rays; the difference is that the X-Rays had a one thousand time shorter wavelength. Also through his experiments he discovered that these X-Rays can penetrate human flesh. This was a huge advancement in medical technology and it will change the course of medicine forever. (reference 2)

First let's look at what an X-ray is. An X-ray is a machine that uses electromagnetic waves that are able to pass through many materials that are opaque to light. These waves have high energy and a very short wave length. The waves of the X-ray can penetrate through the body and are absorbed in varying amounts depending on the density. The question now is: How are the X-ray beams able to penetrate through our skin and provide a clear image of the inside of our body including our bones? The answer is refraction; Refraction is responsible for the image formation of what we see through the X-ray machine.

Refraction of light:
Refraction is defined by the bending of a wave once it enters a certain medium, and then the speed changes; So the refraction of light is when the light passes from a fast medium to a slow medium which would then cause the light to bend. When it bends it bends towards the normal to the boundary between both of the medias. “As the speed of light is reduced in the slower medium, the wavelength is shortened proportionately. The frequency is unchanged; it is a characteristic of the source of the light and unaffected by medium changes” (reference 4). The amount of bending depends on the index of refraction for that specific medium. The index of refraction is found using the following equation:
\[
\frac{C}{N} = \frac{V}{V}
\]

Where \( N \) represents the index of refraction, \( C \) represents the velocity of light in a vacuum, and \( V \) represents the velocity of light in the medium.

Snell's law is used to relate the indices of refraction of the two media to the directions of the propagation in terms of the angles to the normal. This law was named after the Dutch astronomer Willebrord Snellius, it is said that the law was actually described by the scientist named Ibn Sahl at the Baghdad court in 984. Snell’s law states: “that the ratio of the sines of the angels of incidence and refraction is equivalent to the ratio of phase velocities in the two media, or equivalent to the reciprocal of the ratio of the indices of refraction” (reference 4). The mathematical express is as follows:

\[
\frac{\sin(\text{angle 1})}{\sin(\text{angle 2})} = \frac{V_1}{V_2} = \frac{N_2}{N_1}
\]

Where \( V \) is equal to the velocity of the light for each media, and the \( N \) represents the refractive index.

Snell’s law is derived from Fermat’s principle, that states, that the light travels the path that takes the least time. “By taking the derivative of the optical path length, the stationary point is found by looking at the path taken by the light so that it could be noted that the result does not show light taking the least time path, but rather one that is stationary with respect to small variations as there are cases where light actually takes the greatest time path, as in a spherical mirror” (references 4) Physicist sometimes refer to this concept by comparing it to the beach and the ocean by saying “the area of lower refractive index is replaced by a beach, the area of higher refractive index by the sea, and the fastest way for a rescuer on the beach to get to a drowning person in the sea is to run along a path that follows Snell's law. This relates to the X-ray machine because the rays are refracted when passing areas with different opt densities” (reference 5). So snell’s law help physicist relate the concept mathematically.

**X-ray Powder Diffraction:**

X-ray powder diffraction is a little different from regular X-ray diffraction because it uses Crystalline material to help improve the appearance of the cells. This technique is a more rapid analytical technique. These crystals were discovered in 1912 by Max von Laue, these crystalline particles act as three dimensional diffraction gratings for X-ray wavelengths that are similar to
the spacing of planes in crystal lattice. The X-ray diffraction is based on constructive interference of monochromatic X-rays and crystalline sample. “These X-rays are generated by a cathode ray tube, filtered to produce monochromatic radiation, collimated to concentrate, and directed toward the sample. The interaction of the incident rays with the sample produces constructive interference (and a diffracted ray) when conditions satisfy Bragg's Law $n\lambda=2d \sin \theta$” (reference 1).

This law relates the wavelength of electromagnetic radiation to the diffraction angle and the lattice spacing in a crystalline sample; the diffracted rays then get detected and counted. The sample is then scanned through 2 different angles in order to help find the possible deflection directions that the lattice should deb attained because of the random orientation of the crystal powder material. “The conversion of the diffraction peaks into a d-spacings format allows the identification of the mineral because each mineral has a set of unique d-spacings associated with it. Typically, this is achieved by comparison of d-spacings with standard reference patterns” (reference 1)

**X-ray powder diffraction Instrumentation:**

X-ray diffractometers are made up of three major components; the X-ray tube, a sample holder, and an X-ray detector. The X-rays are generated in a cathode tube by heating the filament to help in producing electrons. Then the electrons are accelerated towards the target by applying voltages. Once the electrons have sufficient energy to dislodge the inner shell electrons of the target material, the characteristic X-ray spectra are produced. The spectra produced can consist of multiple components “The most common being Kα and Kβ. Kα consists, in part, of Kα1 and Kα2. Kα1 has a slightly shorter wavelength and twice the intensity as Kα2. The specific wavelengths are characteristic of the target material (Cu, Fe, Mo, Cr). Filtering, by foils or crystal monochrometers, is required to produce monochromatic X-rays needed for diffraction. Kα1 and Kα2 are sufficiently close in wavelength such that a weighted average of the two is used. Copper is the most common target material for single-crystal diffraction, with CuKα radiation = 1.5418Å” (reference 1)

“These rays are then directed onto the sample, when the sample and detector are rotated, the intensity and strength of the reflected rays are then recorded” (reference 1). Once the shape of the incident rays impinging the sample satisfies the Bragg equation, constructive interference occurs and a large peak appears due to the increase in the intensity. The detector then records and processes the rays signal and converts the signal to a count rate which is then recorded on to a computer screen. The geometry of the X-ray diffractometer needs to be oriented in a way where the sample rotates in the path of the collimated X-ray beam at an angle, so that while the X-ray detector is mounted on an arm to collect the diffracted rays.
**Strength and Limitation of X-ray powder diffraction:**
Now that we have discussed X-ray power diffraction and how it works, let's look at the pros, and cons of this technique. One very important trait associated with X-ray powder diffraction is that it is very powerful and much faster than regular diffraction which helps in the identification of unknown materials. Also for this type of x-ray technique, a large sample is not required to receive good results, and the data is very straightforward and easy to use. As this technique provides very good strengths it also has its limitations. When using the X-ray powder diffraction technique, it can be limited because homogenous material is the best type to use for this technique. Also the material used to be detected must be ground into a fine powder in order for it to be read. This can be hard to accommodate depending on the type of substance you are using so it causes some limitation in scientific research.

**Applications of the X-ray machine.**
The X-ray machine has been around for many years, and has been perfected as well through history. The discovery of this concept was a huge milestone in medical and scientific history. Some scientists were able to take that concept and come up with other uses for it, and some took the old ideas and perfected them more and more. Let's look at the medical field first.

If a patient presented to the Emergency room right after a bad fall they had outside, and you examine their arm and you find that it is swollen and red, you also notice that the patient can hardly move it and it is painful to the touch. What would be the first thing a doctor would order? An X-ray would probably be the first thing ordered, because thanks to the German physicist Wilhelm Conrad Roentgen we are now able to use a machine that will release electromagnetic waves that will penetrate through the skin and provide the doctor with a clear picture of a broken bone. In many cases fracture and infection show very well on an x-ray. The doctor can also see what type of fracture it is, how long it is, and how deep it is. This is all due to the X-ray machine.

The X-ray machine also contributed to the discovery of the MRI, and CT scam technology we have today. The X-ray machine could also show fluid or a mass that a patient might have somewhere in their body. These machines are very common and are usually found in many medical offices, for some x-rays the doctor might have a contrast such as Iodine or Barium introduced into the patients system in order to provide a better quality of image. X-ray machines are important in the diagnosis of Arthritis, the rays can show clear pictures of the joints in the body showing any evidence of arthritis, and the doctor can also follow up with patients by doing frequent x-rays of their joints in order to keep track of their arthritis to make sure it isn't worsening (reference 3)
X-ray Machines are also very important in diagnosing chest issues such as lung infections. An image from an x-ray machine can clearly show any evidence of lung cancer, tuberculosis or pneumonia. They would in most cases show either masses (which so in a darker grey) or fluid which will also do the same. There is also a special type of X-ray test that is used to help examine the breast tissue for breast cancer. Heart disease show up very well also on an X-ray machine.

**Applications of X-ray powder diffraction:**
The powder diffraction technique as we discussed previously is mostly used for determining an unknown substance. This is very important when it comes to research. For example when the come up with new drugs, it is crucial that they review the substances being use and test them to be sure they are safe for a patient to consume. Or if we look in the geography field, this is a very important technique because when new elements or substances that are found, they can use the X-ray powder technique and be able to test it in order to help figure out what the substance is, and the best thing about it is that they would only need a very small amount of the substance ignored to figure out what it is. So it is a very rapid process that gives accurate results that are useful for scientist.

**Conclusion:**
In conclusion I believe that the X-ray diffraction technique is a very useful advancement in the scientific field. The X-ray machine was extremely important in the medical fields because it allowed doctors and medical staff to explore the body even more and to be able to diagnose patient with more accuracy. Doctors now had a way to see inside the body and get a clearer picture of what’s going on in the inside. Doctors were able to use this technique to help diagnose lung cancer, breast cancer, bone tumors, body tumors, teeth infections, bone infection, heart issues such as heart enlargement, or fluid in the heart.

The concept of light diffraction was key in this development because it explains what Wilhelm Conrad Roentgen saw in his laboratory back in 1895. It was the concept of light diffraction that allowed scientist to recreate what had happened in Roentgen’s laboratory and understand it better, and also perfect it so that they can put it into good use. This was also a big advancement in the physics field because it helped create a new phenomena that future physicists were able to build on and then create and discover new medical advancement that came later on in history such as MRI machines, and CT scans, where the basic concept is the same but more advanced and new concepts intertwined as well.

The X-ray powder diffraction was also another huge advancement in the science field because it helped many different fields of science such as geology, environmental sciences, biology,
chemistry, physics, and medicine. This concept was also a build on the general X-ray diffraction technique, only this one was faster and required powdered samples, this type of x-ray was not meant to be used on humans, but instead it was meant to help scientist discover new things. The X-ray powder diffraction was used to identify unknown substances which was useful in the development of new drugs, chemicals, elements, and much more. For this reason i believe that the X-ray diffraction concept is extremely important in science overall because it impacted and influence all the different fields, and helped them all grow and develop.
FIGURES:

This picture shows the angles and the direction of the beams in X-ray diffraction. (reference 5)

This picture shows a X-ray Powder diffraction machine used to help identify unknown material (reference 1)
This picture shows the waves and data collected by the X-ray powder machine showing the simulated waves in comparison to the actual wave they used to see if they are the same material (reference 1).

Picture of the first X-ray image, (reference 2)
References:


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The War of Currents: The History, Benefits, and Future of
Direct and Alternating Currents

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ABSTRACT

The term “War of Currents” is a reference to an infamous feud between the well-known American Thomas Alva Edison and Serbian futurist Nikola Tesla about what the standard type of power (AC or DC) should be in the United States during the late 1880s. Thomas Edison who was supporting DC or “Direct Current” Power systems which at the time were the dominating the American market but had known issues and Nikola Tesla who was lobbying for the implementation of AC or “Alternating Current,” a more dangerous but much more efficient medium of energy. Researching each current type shows that both types have many benefits and a few cons associated with them, but both are needed in the world. Although a “War” broke out to decide which would become the standard, you could very easily argue that both types won. Predicting the future with the two currents is difficult but it’s known that some stuff will change but we won’t know what until technology advances to warrant change.

The term “War of Currents” is a reference to an infamous feud between the well-known American Thomas Alva Edison and Serbian futurist Nikola Tesla about what the standard type of power (AC or DC) should be in the United States during the late 1880s to early 1900s. Thomas Edison, a well-known American inventor credited with creating such machines as the light bulb, phonograph, and moving picture, was supporting DC or “Direct Current” power systems which at the time were the dominating the American market but were unfortunately known as a luxury because of major efficiency problems. Nikola Tesla, a physicist and former electrical engineer for Edison, who today is revered for his absolutely brilliant and radical ideas in the field of High Voltage Electricity and Electromagnetism, lobbied for the implementation of AC or “Alternating Current”, which is known as a more dangerous but much more efficient medium of energy. Both types of Power System have benefits and hindrances that these men debated and fought over and the outcome of this “war” has forever changed the way the world sees electricity and has helped propel us into the future. In this paper, the histories of each inventor will be discussed, the events of the war, the benefits and cons, as well as what the future looks like for each current will also be brought up and discussed.

In order to get a better idea of these inventors’ beliefs and why the feud played out the way it did, it’s important to look at each of their backgrounds up to the point of the “war”. The backgrounds segment in the paper will include details each inventor’s early life, and their career accomplishments.

Thomas Alva Edison was born in Milan, Ohio February 11th, 1847. He began working with machines in the late 1860s as a telegraph operator when he filed for his first patent, an automatic vote recorder for legislatures, at the age of 21. In the following decade, Edison had filed over 120 more patents including such inventions as the quadruplex telegraph, carbon transmitter, and even the phonograph. Not soon after Edison builds his first scientific and industrial research facility in Menlo Park, New Jersey. There he filed patents for such things as the carbon-filament lamp and even a DC generator for incandescent electric lighting which would soon be the biggest invention in America in the 1870s. With his success and notoriety only growing, and his DC generators powering large areas of New York, Edison decided to
Before moving on from Menlo Park and building new facilities in New York and then in West Orange, New Jersey in 1886, Nikola Tesla was able to focus on improving the DC generator and continue to work on inventors.

Nikola Tesla was born in what is now modern-day Croatia on July 10th, 1856. From a young age, he showed signs of immense genius, being able to perform advanced integral calculus in his head. Tesla went to study engineering at Graz’s world-renowned polytechnic institute in Austria. After graduating, he moved to Budapest and worked as an engineer at the Central Telephone Exchange. It was there that the idea for the Induction (Alternating Current) motor was born. Nikola Tesla spoke about the moment he came up with the idea that would later change the world: “One afternoon, which is ever present in my recollection, I was enjoying a walk with my friend in the city park and reciting poetry. At that age I knew entire books by heart, word for word. One of these was Goethe’s Faust. The sun was just setting and reminded me of a glorious passage:

The glow retreats, done is the day of toil;
It yonder hastes, new fields of life exploring;
Ah, that no wing can lift me from the soil;
Upon its track to follow, follow soaring!

As I uttered these inspiring words, the idea came like a flash of lightning and in an instant, the truth was revealed. I drew with a stick on the sand the diagram shown six years later in my address before the American Institute of Electrical Engineers.” After constructing the AC motor, Tesla decided to come to America, where he worked as an associate of Edison in 1884. After two years of working with Edison and failing to gain Edison’s interest in the AC induction motor system, Nikola decided to establish his Tesla Electric Light and Manufacturing Company, where he would be able to focus on his AC Induction machine.

After the fallout between Edison and Tesla regarding what power system, AC or DC would be more functional, Tesla left his job working for Edison and created his own company, the Tesla Electric Light and Manufacturing company, where he began working and focusing on his induction motor. A storm was brewing between these two inventors, and only one man would come out on top. Already controlling the electricity empire in New York city with his high voltage direct current transformers, Thomas Edison was looking to expand. But with the establishment of Tesla’s new company along with George Westinghouse’s patents for induction motors, AC power systems were looking more and more like the right option in the eyes of the public, considering they would provide electricity to many more people, making electricity no longer a luxury for the rich but more of an “amenity” for the middle class. Worried that he might lose his financial backing from government grants and lead investors like John Pierpont (JP) Morgan, Edison launched a public scare campaign that played to the public’s fear of the new and upcoming electricity. Stating that Alternating currents were “as unnecessary as they were dangerous” and “Just as certain as death, Westinghouse will kill a customer within 6 months after he puts in a system of any size.” He began his campaign of using alternating currents to electrocute animals in public demonstrations to show the looming danger of the induction generators. Edison would collect strays and baby animals to appeal to the crowd and execute
them on giant metal stands to show the gruesomeness of the AC if not properly treated. There are two famous incidents caused by these public demonstration. The first being the very first execution by electric chair. Not initially an idea by Edison, but one of his associates, they decided to use the same technology (AC generator) to kill convicted murderer William Kemmler as part of his death penalty. This is famous because it is the very first reported use of an electric chair and it went horribly. They did not give the dynamo enough time to build any current and Kemmler sat there for an extended period of time being electrocuted. Still alive and screaming reporters broke the story and Edison’s reputation was severely damaged. Westinghouse quoted on the event “It has been a brutal affair,” he said. “They could have done better with an ax.”

Still reeling from the nightmare that happened with Kemmler and trying to save his reputation as well as destroy Westinghouse’s and Tesla’s, Edison set the stage for his last and biggest public demonstration. The execution of Topsy the elephant, a former circus elephant that had recently killed three of it’s trainers. The demonstration was performed in front of thousands of people in Coney Island, New York.

Even this demonstration wasn’t enough to push public to DC, at the Chicago World fair in 1893, at the peak of the “war,” Edison lost the bid to be the electricity supplier to Westinghouse and Tesla with their cheaper and more efficient power system. Alternating Current electrifying the fair ending up swaying the public opinions as well as multiple big companies allowing AC to become the US standard. Despite his best efforts, Thomas Edison had lost and Nikola Tesla along with George Westinghouse ended up winning the “War of Currents.”

After seeing what happened during the war, it is important to understand what AC and DC power systems actually are. Both currents have multiple benefits as well as some cons associated with them. Both currents also act in different ways, but it’s important to realize that both are needed in todays age, so there can never just be one.

DC or Direct Current is described as the unidirectional flow (Current that continually flows in one direction) of electric charges from a source outward, it never changes. Direct Current is mostly associated with batteries and Solar Cells. Although mostly associated with these there are two different types of Direct Current, the regular (Low voltage) which include batteries. and Transmission (High Voltage) which is more recent and include long distance power lines. Some of the major components of HVDC systems include: Thyristor valves, VSC valves, Transformers, AC Filters, capacitor banks, and lastly DC filters. Even though its known as a safe and quality means of energy, DC has had a reputation to not be the most efficient choice. Although DC once ruled the New York cityscape with the invention of Thomas Edison’s DC power transformer in the 1870s, it is not as widely used now for long distance power transmission. Today DC power systems are mostly used in commuter type machines (cars, boats, planes, etc) and backup power generators. Direct current circuits differ from Alternating induction current circuits in that that usually require different connectors, converters, and fixtures because of the difference in voltages.

It’s been mentioned before that direct current power systems have a reputation of being safer and more reliable, and although this is true, Alternating current seem have been the top choice for long distance power transmissions. That is because at the time, direct current technology wasn’t as advance as alternating current. Alternating current might have been more dangerous but it was much more efficient and able to reach a much farther distance from the
power stations without needing additional stations along the path. This was very true for a time but due to more recent advancements, long distance direct current power transmissions have become much more efficient and the benefits of it may very well outweigh the costs. The first and most apparent benefit of direct current power transmission systems would be the cost. High Voltage direct current transmission systems have become cheaper over the years and now the total cost of building plus maintenance of the transmission stations actually costs less than the conventional alternating current transmission systems. Another huge benefit to HVDC transmission is because direct current is still a safer mode of transport of energy, that allows for underground transmission in urban areas to become a viable option. Underground energy transmission in urban areas drastically lowers the cost of maintenance. Related to the underground transmission, underwater transmission is also possible with HVDC.

Alternating current is in reference to the flow of electric charge intermittently changing directions unlike the continual path that direct current flows. Alternating current is still more widely used than direct current because of its superiority in ability to lower and raise voltage to what was needed during the infancy of the electrical age. Most sockets and household fixtures are built to deliver alternating current, but depending on the appliance in the house, there are a majority of the time DC converters present. In North America, alternating current is usually around 120 volts with about 60 hertz. A hertz is a cycle, it basically means that the electric charge changes direction about 60 times a second.

Alternating current is still considered the standard in US power transmission. This is mostly because of two or three main reasons. First of which is the history and infrastructure behind it. After becoming the US standard in the late 19th century, the technological era hit full force. Houses everywhere were getting electricity, something not that far in the past was considered a luxury. Because of this AC induction generator stations were built all over and started providing power to the country. The next major reason it is still the standard is the ease of lowering and raising the voltage through the use of transformers. A big problem for direct current generators, the higher the voltage the stronger the transmission, the stronger the transmission, the farther the charge will travel before a transformer has to be installed on the path.

What does the future look like for AC as well as DC power systems? That’s an interesting thought. After writing this paper it seems that my tune has changed a little bit. I was under the assumption that direct current circuits were inferior, lower power output, lower usable distance from a station. I guess I just always thought of direct current systems as backup systems for when the alternating current power lines wouldn’t work. But after researching all of this, it has come to my attention that direct current might actually become superior in the future. With the technological advances to direct current, it is now cheaper to set up and cheaper to maintain, And I don’t have a feeling that that will change. One of the many articles that I used for research mentioned that we might soon be moving on from AC and coming back to the newer high voltage direct current (HVDC) but in areas when AC is already established, it would not be worth it in the long run. That same article mentioned that AC will almost guaranteed always be our primary mechanism for the majority of transmission because of how much easier it is for alternating current to drop voltage over direct current. I personally foresee a future with direct
current becoming a bigger part of society. From what I read, high voltage direct current transmission systems seem like just a better choice. again, they’re cheaper, they’ve always been more safe, and the fact that they are able to run underground is a fantastic benefit that will only become more and more useful as we progress into the future.
Works Cited


Regenerative Brakes and Their Potential

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Abstract:

The purpose of this research paper is to express the potential of a process that is implemented in hybrid and electric vehicles called regenerative braking. Regenerative braking is a process that takes kinetic energy that is formed when a vehicle is braking, by converting the energy so that it can be stored. Then it is converted back to kinetic energy to accelerate the vehicle from rest when needed. In the first portion of the paper it will cover how regenerative braking works in greater detail. How regenerative braking can help save on fuel costs, potential of the reduced wear on your vehicle, as well as reducing emissions in the atmosphere. Finally what different types of road systems have the greatest potential to save/store energy.

Introduction:

Energy can never be destroyed nor created, only converted into many different forms; this is called the conservation of energy law (Dictionary). This statement is very important for this research paper. Since energy can never be destroyed, it has to be manipulated so that it can be stored and then used in the future for the vehicles own purpose. That is the basis behind regenerative braking. In traditional vehicles, when braking the vehicle creates friction between the wheels and the brake pads in order to slow down to a stop. The friction created is a form of heat or kinetic energy that is dispersed into the air since it has nowhere else to go. Unlike conventional vehicles hybrid and electric vehicles use a regenerative braking system to take the energy, when braking and convert it into storable energy. Once it is stored by using an electric motor (Figure 1), it can be re-used by the vehicle for its own gain. The electric motor acts in the opposite direction of the wheels when braking (Simpson and Stuebing 2011). Since the reverse motion is occurring the electric motor turns into a generator by the process of creating an alternating current. The alternating current then sends the energy to the vehicles battery to be stored. Once the vehicle is ready to accelerate, the energy that was stored from braking will then be converted back to kinetic energy to help propel the vehicle forward. Regenerative brakes will benefit its consumer by creating greater fuel economy and a cleaner environment due to less carbon emissions.

How Regenerative Brakes Work

According to a paper by Evans and Hilger there is a basic theory: that the internal combustion engine produces chemical energy from fuel, a generator is then activated to use the chemical energy to produce or convert it into electrical energy that the electric traction motor drive will use to creating torque to the wheels, to cause the vehicle to move forward (Evans, Hilger ). In order for the vehicle to store or give an initial charge to the battery a breaking process occurs. That process is expressed as; when the vehicle is breaking the wheels will create a negative torque that will be sent to the electric traction motor. This will cause the motor to go in reverse causing it to act as a generator and produces energy that will charge the vehicles battery. If there was one key component to the regenerative braking system it would be the
electric traction motor drive. The reason is because the electric traction motor drive determines the efficiency in the regenerative braking system because it is directly involved in converting the energy that is being stored or used to charge the vehicles battery. In Figure 2 expressed the basic model of a regenerative braking system.

Factors that contribute to Regenerative brakes efficiency

Granted regenerative breaks are only in effect if they are being used. For example if one travels to work every day and only uses the freeway with no stopping the regenerative breaks are not being used all at. On the other hand if that same person is driving in the city with frequent stops they will be getting the most use out of their regenerative system and have the highest potential savings (Clegg 1996).

Edinburgh Napier University Research

To better show the potential of regenerative breaks, research that was done at Transport Research Institute at the Edinburg Napier University, shows how effective the regenerative braking system can truly be. What the research team did was take a Škoda Fabia which is a small conventional vehicle that has no regenerative braking system. The team then applied a regenerative braking system and ultracapacitors to the Škoda Fabia and tested for energy efficiency and emissions (Williamson, Emadi, Ragashekara). All of the following research can be expressed with Figure 4. They found that after testing the vehicle in both city and rural environments in which the Škoda Fabia originally consumed 5.24 Liters per kilometer and emitted 140 grams of fuel for every kilometer traveled. That with the new regenerative braking system implemented within the vehicle, it performed with a 54% increase in fuel efficiency in the while in the city, and a 29% increase while in a rural area, as expressed in Figure 3. The research also presented another crucial finding, which was that the regenerative braking system reduced the carbon dioxide emissions. The findings showed that the carbon dioxide emissions had a 52% decrease when adding the regenerative braking system. (Williamson, Emadi, Ragashekara). This research was able to express, that not only can this system save the consumer money, but also the environment.

Calculations of Potential Savings

Also a practical application with all the calculation will be expressed down below.

The amount of recoverable energy is dependent on the amount of energy required to stop the moving vehicle (Jason Fenske 2013). In the following equation below, we are not including air resistance.

\[ \text{Kinetic Energy} = \frac{1}{2} \text{mass} \times \text{velocity}^2 \]

1600kg car - 3530lbs
100km/hr -62 mph
At 70% efficiency (regenerative braking efficiency)
½(1600kg)(100km/hr)^2
= ½(1600kg)(100km/hrX1000m/kmX1hr/3600s)^2
= 22,222.2kgm^2/s^2 = 222.2kJ

Now we take the efficiency losses
222.2kJ x 70% = 155.54kJ Recovered

Now we will use that to find out how much gas is 155.5kJ (how much gas saved)
Energy removed = mass x energy density of gasoline and efficiency of converting energy into usable energy (n)
E = m x (E/m) x n
155.54kJ = m (43,000kJ/kg)(35%)
M = 0.010kg

Then convert the mass from kg into L
1 L gas = .75 kg
.010kg/.75kg = .0133 L saved every time stopped from 100 km/hr
So if you stopped 76 times, the savings would be 1 L of gas.
(Serway, Vuille), (Fenske)

Helping prolong the atmosphere

Over time this not only saves on gas but also on the wear and tear that would normally be applied to the engine and the vehicles brakes. Another important note would be that since there is less fuel being consumed it is reducing the amount of carbon dioxide emissions that is entering the atmosphere. Another contributing factor to carbon emissions is the initial startup of a vehicle. As expressed by Brandenburg and King “It also becomes preferred to preheat the engine before start up, since it has been determined that approximately 80% of the emissions from a typical vehicle are created during the ignition and initial driving of a cold engine” (Brandenburg, King). One of regenerative brakes goal is to help solve this problem by using the energy stored to warm up the engine before ignition, which would reduce the amount of carbon emissions (Evans, Hilger).

Conclusion:

Throughout this research paper I expressed how regenerative brakes work, the factors that play a role on the efficiency of the system, how the system has great potential to save, and what type of transit route could have the greatest potential to save/store energy. I believe that in the future, regenerative braking will become so popular that it will be hard to find a vehicle that does not have the system implemented in it. Due to the advancement in technology, the continuing needs to save money and the environment. As stated above in the Edinburgh Napier University research, that there was at 54% increase in fuel efficiency in city which will save the consumer money. Also the research expressed that the Škoda Fabia had a decrease in carbon dioxide emissions of 52%, and that is just for one vehicle. Imagine if majority of the world started to use regenerative brakes, how clean our atmosphere could potentially be.
Figures:

Figure 1

http://www2.hesston.edu/physics/201112/regenerativeenergy_cw/paper.html#Flywheel

Figure 2

<http%3A%2F%2Fdocs.lib.purdue.edu%2Fcgi%2Fviewcontent.cgi%3Farticle%3D1005%26context%3Dtechmasters>.

Figure 3

Figure 4

References


https://www.youtube.com/watch?v=kNDWzFaiDZQ


http://www2.hesston.edu/physics/201112/regenerativeenergy_cw/paper.html#Flywheel
Physics of Horseback Riding

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PHY 112
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Research Paper
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Abstract

Horseback riding may not come to mind when talking about physics. When riding a horse and even just doing simple groundwork with the horse, the rider and the horse are experiencing forces on them. By studying the amount of work, power and energy that is involved with horseback riding, it can show just how physics is incorporated.

Horsepower

Horseback riding to some is a way of life, for others, it is a hobby or a passion that allows a person to escape from the everyday reality. Whether out riding in a competition, leisurely riding or doing groundwork, forces are acting upon the horse and rider, so physics directly relates with horseback riding. When the horse is just running wild in the arena, there are forces, power and energy being converted by the horse, which is something that many people do not realize. Horsepower, which is a unit of power is directly related to horses. Engineer James Watt, who was conducting experiments in the eighteenth century, created the term horsepower. He found that one horsepower is equal to 746 watts; the equivalent in heat is 2,545 British Thermal Units per hour. Another way to measure power is metric horsepower, this equals to 4,500 kilogram-metres per minute, which is equivalent to 0.9863 horsepower. The experiment Watt conducted was observing the amount of work a large Draft horse could sustain in just one day. He found that these draft horses were able to conduct 50% more power than an average horse can do per day. This was pretty interesting because it shows just how powerful an animal can be. Horsepower is a remarkable discovery because it has helped build better and stronger engines. There are different horsepower in each part of an engine, for example, the electric motor horsepower can be found from the electrical input watts. This allows for heat and friction losses in the motor. Another type of horsepower is thrust in jet engines and rockets. The thrust is found by multiplying thrust in pounds by the speed in miles per hour divided by 375 (one horsepower). Horsepower is used in many different ways in everyday life, from cars to jet engines; James Watt really made a very important impact in the world of science and physics.

Understanding the gaits of a horse

To understand just how riding is affected by physics, one must understand the basics of horseback riding. First, the horse has four different gaits at which there are different speeds at each gait. These gaits are, walk, trot, canter/lope, and gallop. Walking is the slowest of the gaits and a gallop is the fastest. At the trot, the horse is at a light jog, and the diagonal pair of legs is lifted. At the lope, the horse is running faster now and there instead of a two beat gait, it is now a three beat gait. In the canter, the horse has a leading side, where the two left or right legs are extending further than on the opposite side. At the gallop, the horse expands its gait from the canter, allowing the three-beat gait to become a three-beat gait. Usually, to get the horse to go from one gait to the next, the rider will apply a force to the side of the horse with their heal, or kicking the horse gently. While trotting, the rider may feel more comfortable by rising to the trot, or posting. This helps the rider get into more of a rhythm of the horse, so both the rider and the horse are comfortable. When doing the posting trot, the rider must rise or stand when the horses
outside leg (assuming the rider is in a circular arena) is extended. This can get quite tiring on the riders legs due to the constant up and down motion. The physics that is applied to the posting trot can be explained by the weight of the rider multiplied by the displacement or the measurement of the length the rider traveled upward from the saddle. The amount of work that is done is usually in the units of joules. This shows how much work the rider at the posting trot does.

Velocity at different gaits

Velocity is considered a vector, which means that it has magnitude and direction. The units that represent velocity are meters/a time unit. Average velocity is also a vector quantity, but it also has direction involved with it and can be either a positive or negative value. Horses can have many different velocities because they have more than just one gait. While walking, which is the four beat gait, the horse’s velocity is about 6.4 kilometer/hour. When watching a horse walk, it is clear to see that the horse’s head moves up and down. The horse’s head moves up and down because it helps balance the horse since their balance is not the best at this gait. This can be seen shown by looking at the horses vector components; the horse’s x-component is smaller than its y component due to the movement of their head going up and down to help keep their balance. When the horse moves faster, the movement of the horse’s head is smaller because the horse is more balanced at the faster gaits. At the trot, the horse moves at about 13 kilometers/hour, so almost double the velocity of the walk. The horse is more balanced at the trot because two of the horse’s legs are always on the ground at the same time. The next gait, which is the lope, is a three beat gait and the horse moves at a velocity of about 16-27 kilometers/hour. At the gallop, which is the fastest gait, the horse’s velocity is 40-48 kilometers/hour. The horse’s velocity is always changing and the riders must know how to be balanced with their horse at every gait. If the rider is not balanced, then the whole ride can be compromised. It is important that the riders understand the different velocities of the horse, because different types of competition deal with the horse’s gait, and some riders can be disqualified for going too fast. In barrel races, many of the young age groups, such as the riders that are in the walk, trot, lope group cannot go faster than the lope. If they are moving faster than the lope, they will be disqualified. Some competitions go as far as having a break out time for their specific group because it makes the competition fairer.

Acceleration while the horse is running

As the horse is running in a circular ring, the horse is experiencing different changes in acceleration, because the direction is constantly changing. This type of acceleration that the horse is experiencing is called centripetal acceleration. Acceleration is defined as the change in velocity divided by the change in time. So, when the horse is running around the arena, the x and y velocities constantly change, which means the acceleration, is changing. Imagine the horse in a ring running in the counterclockwise direction. When the horse is running in the counterclockwise direction and hits the eastward direction, the horse’s eastward velocity has decreased, but the northward velocity increased. There is also centripetal acceleration the horse is experiencing...
because at any one point in the ring, one of the horse’s velocities is decreasing or moving towards zero. So, this means that the acceleration is pointed inward, to the center of the circle. To find the magnitude of the horse’s acceleration, the velocity of the horse and the radius of the circle must be known. For example, if a horse is trotting at five miles per hour and the radius of the arena is ten feet, then the magnitude of acceleration can be found by squaring the velocity and dividing it by the radius. So, the magnitude of acceleration for this scenario would be, $2.5 \text{ m/h}^2$.

Newton’s Laws of Motion

Newton’s first law of motion states that, an object will stay in motion unless an outside force acts upon the object. When talking about horses, Newton’s first law of motion comes into play when the horse’s velocity changes. This means that when the horse is trotting at a constant speed, the forces in the vertical direction is constant due to gravity, but in the horizontal direction, there are not any forces. This can be seen when the horse is jumping. As the horse is airborne, the only force acting on the horse is gravity. The rider must learn balance when riding a horse, balance is very important because if a rider’s balance is off, then both the horse and rider will be uncomfortable. While trotting in a straight line, the rider is balanced, but if the horse suddenly makes a sharp left turn, the rider will lose their balance, and want to keep going in the direction of the before the horse made a sharp turn. Newton’s first law affects barrel racers probably the most because the horse is traveling at such a high speed and must make sharp turns around three barrels. The rider then must brace themselves for the turn and either hang onto the horn, or sit as deep as they can in the saddle. Practicing these turns for barrel racers is very important because it helps them learn balance and once the rider knows how to brace themselves for the turn, their run will be much smoother and faster. Riders have to constantly change their balance because they want to remain at equilibrium with their horse. Whenever there is a change in the horse’s gait or direction, the rider will have to adjust himself or herself to make sure they are balanced with the horse.

If a horse’s acceleration is changing, this means that there has to be a net force to change acceleration. Net force, mass and acceleration are all needed to understand Newton’s second law. Newton’s second law of motion is defined as the acceleration of an object produced by a net force which is directly proportional to the magnitude of the net force. This law can be shown in the equation: force equals mass multiplied by acceleration. So, the net force that is acting upon the horse to cause the change in acceleration is due to its hooves when they land and push against the ground. Due to the force of the horse’s hooves against the ground, and the direction the horse is moving, the horse’s acceleration is changing.

Newton’s third law of motion states: for every action, there is an equal and opposite reaction. The third law can be put into play and shown when the horse is jumping over an obstacle. As the horse jumps, it pushes itself off the ground and into the air, over the obstacle. The force the ground applies to the horse to help propel the horse over the obstacle is enough for the horse to clear the obstacle. Another example is when a horse kicks a fence over with its hind legs. Since the fence’s mass is much smaller than the horse, the fence will experience a larger acceleration as compared to the horse. The horse will have a very small or zero acceleration, but the fence will be moved much further due to the extreme differences in the mass of the fence and horse.
Another example of Newton’s third law has to deal with every rider’s worst nightmare—falling off of his or her horse. Newton’s third law entails momentum as well, and can be found by multiplying the mass of the rider by the velocity of the horse. When the horse bucks or makes a sudden stop where the rider is not prepared for it, then the energy from the horse will be transferred to the rider. The motion of the rider (if moving in the forward direction) will continue to move in the forward direction until the rider lands on the ground. Some people say that their momentum made them do something, and it is true, and in this case, it is true. Momentum is related to the velocity the object is traveling at. So, the greater the velocity, the greater the momentum the object will have. Serious injuries can occur due to falling off a horse, so it is important that riders wear helmets to protect themselves from any life-threatening injury.

Energy

Energy is all around, and is defined as the capacity for doing work. Energy cannot be created nor destroyed; it can only be transferred from one object to another. When dealing with motion, anything that is motion can say there it contains kinetic energy. This means that it is energy in motion, and that an object in motion can indeed do work on anything that it comes in contact with. Potential energy, which is stored energy, is a scalar, and not a vector. Kinetic energy is only a scalar quantity, which means that it does not contain direction as well. This can be seen when a car is moving at a high velocity, and it hits into a wall, or another object. After the car has hit the wall, it can be seen that there is a great deal of damage to the car and the wall has been damaged as well. Kinetic energy can be quantified through the equation: \( \frac{1}{2} \text{mass} \times \text{velocity}^2 \). Energy is all around us, and is a very interesting component to everyday life.

Potential Energy

Kinetic energy deals with horses just like how kinetic energy pertains to moving cars. When a horse is galloping at 40 kilometers/hour, and if there is a barrel the horse is going to come in contact with, it can be said that the horse is going to do work on the barrel. By using the example of the horse kicking the fence that was mentioned above, it can be said that the energy the horse possesses before kicking the fence is potential energy and is changed to kinetic energy when the horse performs the kick. That energy is then transferred to the fence, causing it to fall down, or bend. Potential energy is defined as the energy that is stored in an object because of its position. The potential energy can be quantified by using the equation: mass times the gravity times the height of the object. So if the horse is about 4.5 feet tall, weighs about 1000 pounds, and gravity is 9.8 m/s\(^2\), the only piece left to find is the horse’s mass. Different breeds of horses will have varying masses, but usually a horse weighs about 1000 pounds; this converts to about 453,592 grams. So by multiplying the mass, gravity and height of the horse, the potential energy of the horse is 20,003,407.2 joules. To find the kinetic energy of a horse moving at 40 kilometers/hour, the equation \( \frac{1}{2} \text{mass} \times \text{velocity}^2 \) will be used. The horse’s mass is going to be needed as well, just like in potential energy; so the kinetic energy of a horse that weighs 1000 pounds and is running at 40 kilometers/hour will have about 362,873,600 joules of energy. 300,000,000 and 20,000,000 joules of energy are both quite large amounts of energy for one object to have, but it shows just how much
energy that can be transferred from one object to another. It is interesting to see how strong a horse is, it really puts in to perspective just how strong this animal is, and people tend to underestimate them.

Inertia

Inertia goes hand in hand with Newton’s first law of motion, which states: “an object at rest stays at rest and an object in motion stays in motion with the same speed and in the same direction unless acted upon by an unbalanced force.” When the object that is in motion, and resists any change in its motion has inertia. Inertia is based off of the mass of the object, so if an object has a large mass, then it its inertia will be greater than an object with a significantly smaller mass. Inertia is related to Newton’s second law of motion, incorporating the equation, force equals mass times acceleration. So, for an object to change its position, the force has to be greater than the object’s mass. So, inertia and mass are related because inertia is directly related to mass, the greater the mass, the greater the inertia an object possesses. Objects do not just move in straight lines, so what happens to the inertia when an object is moving in a circular path? When an object is moving in a circular path, the object is experiencing rotational inertia. Rotational inertia can be quantified by using the equation mass times the radius squared. So, if a horse is moving in a straight line and in a circular path, their inertia may change slightly, but not a huge amount because the horse’s mass is the same.

Barrel racing

Barrel racing can be a very fun competition to watch, and it seems very simple to run the cloverleaf pattern, but it takes skill and lots of practice. The pattern consists of three barrels set up in a cloverleaf pattern. The rider can either go to the left or right barrel first, and make one left hand turn, and two right hand turns, or do one right hand turn and two left hand turns. Usually, the decision to go to the left or right barrel fist is determined during the practice, and most of the time, the horse favors making left or right hand turns. Since this is a timed event, and the best time wins, it can be stressful on the horse’s legs and joints due to the sliding stops and the three quick, tight turns around the barrels. Depending on how big the pattern is set up in; the average time to run the pattern would be eighteen seconds. There are many forces that are acted upon the rider and the horse, like stated above, but one part of physics that is forgotten while riding is determining the torque that is applied to the horse’s joints while turning.

Torque

Torque is the measure of the force applied to an object that causes that object to rotate, and is a vector. Since a pivot point in need for torque, the pivot point will be the horse's hock in its back leg (the hock is equivalent to the human knee). When barrel racing and turning around the barrels, the horse pivots around its hind legs, and then pushes itself around the barrel with its hind legs. By doing type of activity, the horse has a good amount of torque being applied to its hocks. The amount of torque a horse’s hock is experiencing is found by multiplying the force and the distance from the force to the axis of rotation. The unit for torque is Newton-meter; even though it is equivalent to a joule, it is important that torque is not a form of energy. By applying torque to the horse’s hocks, it could result in the horse developing arthritis. This is why it is important
for the rider to make sure they know exactly what they are doing because they could potentially injure their horse, and themselves if they are not riding correctly.

Conclusion

Horseback riding which has been a part of life for many years, and involves a great deal of physics that many people do not realize. By researching the necessary material, I have learned a great deal about horseback riding. The research that I have conducted has opened my eyes as to why certain things happen when I am riding. For example, I now understand why I have to hang onto the horn when turning around a barrel during a barrel race. The equations and research could be directly applied to horseback riding, and it made sense when I was applying the equations to riding. Now, when I ride I find myself contemplating what velocity, and how much force, momentum, and acceleration the horse has. I believe that with increasing knowledge in the physics and science world, many more injuries will be able to be prevented and more protective gear will probably released for both the horse and rider. Doing this research, I have greater understanding of just how much work, and energy a horse can have, and I am amazed that an animal can be that powerful.
This circled part in the horse’s legs is the hock, which is the area where the horse experiences torque when turning.8

This is the standard cloverleaf pattern that the barrel racers run when competing.2
References


Retiring the Drill: The New Age of Dentistry

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Abstract

Approximately 5.0% of Americans abstain from seeking dental treatment based solely on perceived feelings of apprehension that is attributed to the dental drill. Although fear and anxiety is manageable, the current practice of using dental drills is not only invasive but can be destructive to the anatomy of the tooth. New insight, however, in the advancement on modern dental practices suggests that the implementation of non-thermal atmospheric gas plasmas could provide patients with a more comfortable environment while eliminating the need for a dental drill. With sufficient evidence supporting the efficacy of non-thermal atmospheric plasma application in bacterial eradication as well as the treatment of tooth decay, intraoral diseases, root canal therapy and composite restoration, the need for the dental drill may become obsolete.

Introduction

It is not a misbelief that fear, anxiety, and dental visitation coincide with one another. According to the American Dental Association, evidence suggests that nearly 60 million Americans avoid visiting a dental professional because treatment is perceived as painful. As a result, dental practitioners are disliked and uphold a notorious reputation. This is mainly attributed to the dental methods that are currently in practice, specifically, the use of the dental drill. With respect to all facets of the medical industry there is an aspiration for producing innovative technology that could further enhance the quality of modern medicine. In the case of dentistry, this revolves around the implementation and utilization of non-thermal atmospheric plasma (N-tap) technology. Over the last decade, biomedical engineers and researchers have developed and tested several devices that are able to convert specific gases at atmospheric pressure into cold plasma. Proven to be effective in the eradication of specific bacterial species as well as the treatment of tooth decay, intraoral disease, root canal therapy and composite restoration, N-tap technology may become the limelight of modern dental practices.

The Dawn of a New Age

Plasma: An Overview

Most people who’ve acquired a basic understanding of science are familiar with the three states of matter: liquid, solid, and gas. However, a fourth state of matter exists and is referred to as plasma. Even though plasma’s name is not as renowned as its counterparts, it contains highly unique properties that are unlike those of the other states previously mentioned. In the most simplistic definition, plasma can be defined as an electrically neutral medium of unbound electrons and neutrons orbiting its nucleus. In other words, the overall charge of plasma is equivalent to zero. And although it was described that plasma contains unbound electrons and neutrons, one must not interpret “unbound” as being “free”. From a physical standpoint, these subatomic particles are not considered free due to the influence that external forces can have on them. One such example can be shown through the production of electrical current with magnetic fields where electrons and neutrons attraction for one another can be altered. As a result, external forces have the ability to govern the collective behavior of these subatomic particles. Due to the distinct nature of these behaviors, more than one class of plasma exists.

Thermal vs. Non-Thermal Plasma

Classification of plasma can be divided into the two following categories: thermal or non-thermal. The basis for classification is dependent on the two variables: (a.) the relative
temperature of the subatomic particles within the plasma and; (b.) the level of ionization those subatomic particles hold. When accounting for these two variables, one can gain a better understanding of the differentiation between thermal and non-thermal plasma. Thermal plasma, commonly referred to as hot plasma, exists when both neutrons and electrons are equivalent in temperature ($T_n = T_e$) and are therefore known to be in thermal equilibrium. Alternatively, non-thermal plasma, or cold plasma, exists when neutrons and electrons vary in temperature. In the case of non-thermal plasma, neutrons are known to be at a lower temperature than that of electrons ($T_n < T_e$).

Coinciding with relative temperature, subatomic particles must attain a certain degree of ionization in order to generate plasma. It is here that the term “plasma density”, more commonly known as “electron density”, is used to describe the number of free electrons per unit volume. In the case of plasma, the degree of ionization is proportional to the number of atoms that have gained or lost one or more electrons. This variable is ultimately controlled by the relative temperature. The degree of ionization, $\alpha$, can be formulized mathematically as $\alpha = \frac{n_i}{n_i + n_n}$, where $n_i$ represents the number density of ions and $n_n$ represents the number density of neutral atoms. Furthermore, plasma or electron density, which relates to the degree of ionization by the average charge state, $Z$, can be formulized mathematically as $n_e = (Z)n_i$, where $n_e$ represents the number density of electrons. Comparatively, thermal plasmas contain a much higher degree of ionization than of non-thermal plasmas. Due to their lower level of ionization, non-thermal plasmas are able to be used in plasma technology, earning them the reputation as “technological plasmas”.

Non-Thermal Atmospheric Plasma

As it was previously discussed, N-tap exists at lower temperatures. In addition to this characteristic, N-tap also contains a lower degree of ionization, which can be found at or below atmospheric pressure. In order to create N-tap, a compound must be converted into gas following ionization through the application of one of five different forms of energy. Forms of energy include direct or alternating current, heat, radiation or utilization of a laser light. Moreover, researchers have discovered that certain elements of the periodic table, when heated, can be utilized as a gas source. The more common elements currently in use include oxygen (O), nitrogen (N), hydrogen (H) and argon (Ar).

Since N-tap is produced and performs optimally at lower temperatures, its use in material science is worthy of recognition. One of the more notable applications of N-tap is shown through its ability to modify properties of biomaterial surfaces. Some of the more common examples of this unique capability are shown through alteration of electrochemical charge and degree of oxidation as well as bondage modification of surface-induce chemical groups. Stemming from that, application of N-tap has the ability to improve the durability of biological structures, resistance to both chemical corrosion and physical abrasion, enhance water absorption potential, as well as extend affinity towards specific molecules. Most importantly, because low temperature plasma is easily confined to small spaces it can be modulated and performed with precision and efficiency when utilized for the improvement of biomaterial surface properties.

The capabilities of N-tap do not only extend to the improvement of biomaterial surfaces, it is also found to be an effective bactericide. Similar to that of disinfectants, antiseptics, and antibiotics, low temperature plasma is efficient in killing harmful microbes. As a result of this unique capability, non-thermal plasma has extended its use in a new era of dental applications.
Regarded as a novel tissue-safe technique, plasma treatment will allow for irregular structures and narrow channels, normally found within a diseased tooth, to be completely eradicated of harmful bacteria. With time, application of N-tap technology in modern dentistry will prove to be more effective and efficient than the conventional methods currently in practice.

**Mechanism of Non-thermal Atmospheric Plasma**

As it has already been discussed, N-tap is said to be “non-thermal” because its electrons are at a higher temperature than the more heavy particles, which are found at room temperature (73°F)\(^4\). When generated, the overall temperature of N-Tap is less than 104°F at the point of contact\(^4\). There are several different methods by which N-Tap can be produced. These methods include but are not limited to radio frequency, microwave frequency as well as high voltage alternating-current (ac) and direct-current (dc) circuits\(^2\). Utilizing these methods, researchers have engineered a variety of different devices, specifically for the production of N-tap and include a Dielectric Barrier Discharge (DBD), an Atmospheric Pressure Plasma Jet (APPJ), a plasma needle and a plasma pencil\(^5\). In addition, both a gas and energy source is required for maintaining production of plasma within these devices. The more commonly used gas sources are Helium (He), Argon (Ar), Nitrogen (N), heliox (HeO\(_2\)) and air (O\(_2\)). Moreover, energy sources commonly used are in the form of thermal, electrical or light. To better understand how these devices perform as a means of production of N-tap, a basic overview for each device is provided below.

**Dielectric Barrier Discharge (DBD)**

DBD’s have many applications in the biomedical industry and is mainly used for the sterilization of living tissue, microbial inactivation, formation of new blood vessels, surface preparation, and dimer formation\(^4\). This particular device consists of two electrodes, wherein one electrode is covered with a dielectric material, and both are separated by an insulating dielectric barrier. The two electrodes are connected to high voltage AC generator, which acts as the electrical source. A gas source of interest (e.g. He, Ar, N, HeO\(_2\) or O\(_2\)) then passes through the two electrodes and is converted into plasma. This process only occurs when one electrode is at a higher voltage than the other, which is grounded. The need for a higher voltage is required to produce the discharge necessary to create plasma. AC high voltages power DBD’s in the kHz range\(^4\). The amount of power consumed when utilizing a DBD falls within the range of 10-100W\(^4\). A schematic of this device is shown in Fig. 1.

**Atmospheric Pressure Plasma Jet (APPJ)**

There are two different categories of APPJ’s: Radio frequency plasma jets and Pulse direct-current plasma jets. Radio frequency plasma jets, also known as plasma needles, consist of two coaxial electrodes. The outer electrode is grounded while the central electrode produces a discharge driven by a radio frequency of 13.5MHz and a power of 50W\(^4\). Located between the two electrodes is a small linear chamber where a gas source of interest (e.g. He, Ar, N, HeO\(_2\) or O\(_2\)) is fed through at a high rate. The reactive species produced within the chamber is then directed toward the nozzle of the device where it takes the form of plasma. This particular class of APPJ has been proven to be effective in eliminating *Escherichia coli* of several *Staphylococcal* species\(^4\). A schematic of this device is shown in Fig. 2.

Pulse directed-current plasma jets, also known as plasma pencils, consist of two disk electrodes that are inserted into a dielectric cylindrical tube and are separated by a small gap. The two disk electrodes are connected to a high voltage pulse generator by a DC voltage supply and are
attached to a dielectric disk by a thin copper ring. Using sub-microsecond high voltage pulses, the two disk electrodes produce a discharge while a gas source of interest (e.g. He, Ar, N, HeO₂ or O₂) is injected through the two holes of the electrodes. As the discharge and gas mixture react with one another, a plume of plasma is formed and exits the nozzle of the device at a high rate of velocity. Similar to that of the plasma needle, the plasma pencil has been used in the treatment of E. coli infection as well as the inactivation of leukemia cells and P. Gingivalis. A schematic of this device is shown in Fig. 3.

The Silver Lining: N-tap Technology in Dental Medicine

The study, diagnosis, prevention and treatment of both oral related diseases and disorders are the focal points of dental medicine. And while dentistry is misrepresented as a medical profession, it plays a pivotal role in disease prevention, which contributes to the overall health and well-being of people. According to one source, each year $60 billion is spent in the United States to treat oral-related disease. Some of the more common diseases include tooth decay, periodontitis and oral candidiasis, and root canal failure, all of which are attributed to bacterial infection. Even though bacterial infection can be prevented and treated, utilization of the dental drill has caused habitual fear among so many prospective patients. There may, however, be a silver lining to this dilemma. Renowned for its abilities to act as a bactericide, initiate cellular detachment, and cause programmed cellular death, N-Tap has taken the interest of many researchers and professionals in dental medicine. Specifically, in the application of bacterial inactivation, tooth decay, oral disease treatment, root canal therapy and composite restoration.

Application in Eradication of Bacteria

The human mouth harbors a vast array of both harmful and beneficial microbiota. According to one source, it is estimated that the average mouth of an adult may contain anywhere from 500-1,000 different species of bacteria as part of their microflora. However, only 100-200 may be alive and thriving at a single given time. This highly diverse microflora inhabits the oral mucosa, or mucous membrane, where stratified squamous epithelium tissue lines the inside of the mouth and throat.

N-TAP efficacy in the eradication of oral bacteria is influenced by three components: (1.) gas composition, (2.) driving frequency, and (3.) bacterial strain. Utilization of N-tap devices is shown to be more a more efficient bactericide than conventional non-thermal methods like that of UV radiation, which on its own is only partially effective. The mechanism by which the N-tap devices act as a bactericide relates to the combination of reactive oxygen species, ions and electrons, as well as UV and electromagnetic fields as a whole.

To begin, the effect that reactive oxygen species (ROS) have on the eradication of microbes plays a pivotal role in the use of plasma. To support this claim, several studies were performed to show the positive effects that ROS can have in bacterial eradication. In one study, APPJ was used with and without oxygen. What researchers were able to conclude was that the time needed to inactivate 90% of the bacterial species was much higher when oxygen was not present. In a second study, researchers used a three different mixture of gases to determine the effect they would have on Bacillus spores. Mixtures included pure helium (He), a mixture of 97% helium and 3% oxygen (HeO₂), and pure air (O₂). In this study, researchers concluded that pure He took 20 minutes to deactivate the bacteria, the mixture of HeO₂ took 10 minutes, and pure O₂ only took 20 seconds.
Charged particles like that of ions and electrons are also crucial to bacterial eradication. Laroussi et al. claimed that charged particles inactivate microbes by rupturing the outer membrane of cellular walls. The study showed that charge accumulation on the surface of a cell's membrane will cause an electrostatic force to overcome the tensile strength of the membrane, therefore causing it to rupture. While this method of bacterial inactivation is effective in eliminating gram-negative specimens due to their irregular cellular membrane, it is not as efficient in gram-positive species, a minor drawback.

The role of UV radiation in combination with plasma is still unclear. What is known is that in the presence of plasma, UV radiation depends heavily on the operating pressure. According to one source, plasmas at very low pressures can produce wavelengths of UV radiation involved in bacterial inactivation. While it would seem as though this would pose implications for N-tap technology efficacy, considering devices operate at atmospheric pressure, utilization of vacuum plasma overcomes that barrier.

**Application in Tooth Decay**

Tooth decay, more commonly referred to as dental “caries” or “cavities”, occurs when the structure of one or more teeth becomes compromised due to the activities of bacteria. On the surface of each tooth is a thin layer of microbial organisms embedded within an extracellular polymer substance, also referred to as a biofilm. This biofilm covers the entire surface of the each tooth where the enamel is present. The nutritional source for these microbes comes from the byproducts (e.g. sugar and starch) of solid and liquidity substances that have been broken down by enzymes. When bacteria use these byproducts as a food source, acid is produced resulting in mineral loss of the enamel. The location of dentin is shown in figure 4. Due to over extensive mineral loss, small holes develop and allow for bacteria to harbor and repopulate. At this point, damage is permanent and tooth structure cannot be restored without dental intervention.

While brushing and flossing are the current means by which tooth decay can be prevented, N-tap technology may be able to replace these conventional methods with the plasma torch toothbrush. Similar to that of the plasma pencil, the plasma torch toothbrush uses sub-microsecond high voltage pulses to create plasma. The plasma produced is precise and confined to the area of contact. This allows for treatment and sterilization of irregular surfaces and makes it a suitable alternative for the decontamination of cavities without the need for a drill. According to one source, substantial evidence shows that N-tap technology like that of the plasma needle is efficient in eliminating *E. coli* and the gram-positive cariogenic bacterium *S. mutans* in tight and narrow spaces. Moreover, since the plasma torch toothbrush is capable of operating at room temperature it will not result in loss of surrounding tissue. While it may be years before the plasma torch toothbrush will be marketed for personal use, similar technology is already available for purchase by dental professionals.

**Application in Intraoral Disease Treatment**

One of the more commonly occurring intraoral diseases is oral candidiasis. Oral candidiasis is a condition where the fungal class *Candida albicans* accumulates on the stratified squamous epithelium tissue which lines the surface inside the mouth. Those who are prone to developing this disease are children, elderly, and immunocompromised individuals. N-tap devices, like that of DBD and APPJ, have been implemented and studied in treatment of this disease. Koban et al., reported that the efficiency of cold plasma could extend to the treatment of fungal infections and
inactivate fungal species like that of *C. albicans*. In this study, Koban et al. points out the cold plasma was able to inactivate the fungal species in less than 30 seconds.

**Application in Root Canal Therapy**
Root canals are similar to cavities in that they are caused by bacterial infection. The main difference, however, is that the location of infection is centralized in the pulp or pulp chamber rather than the dentin. The location of the pulp and pulp chamber is shown in figure 4. In a series of studies, N-tap technology has been proven to be valuable in the therapy of root canals. One study of particular interest showed that APPJ could not only generate plasma in a small confined space but that plasma could come into contact with the nerves of the canal and not cause any pain to the patient. Utilizing a heliox (HeO₂) mixture of 20%, the plasma generated was efficient in eliminating the bacterial species *Enterococcus faecalis*, a main contributor to root canal failure. Evidence, such as this, proves that N-tap technology is valuable to root canal therapy.

**Application of Plasma in Composite Restoration**
Composite restorations, more commonly referred to as dental fillings, are a type restorative material and adhesive that is comprised of synthetic resins. The role of composite restoration in dental medicine is to fill cavities within the dentin surface that are caused by bacterial infection. One of the main issues this restorative material presents in dental medicine is their short-lived lifespan, which is known to only last 4-7 years. Implementation of N-tap technology in dentin surface preparation has proven to extend the lifespan of these synthetic resins through interfacial bonding improvements. According to one study, plasma used in dentin surface preparation increases the interfacial bonding strength between dentin and synthetic resins by up to 60%. This was shown through plasmas ability to improve synthetic resins performance in durability and longevity.

**Conclusion**
In effort to repute the general public’s opinion of dental professionals, utilization of N-tap technology in dental treatment could be the silver lining to a serious dilemma. Renowned for its ability to eradicate bacterial infection, N-tap is proving to be effective in the treatment of tooth decay and oral related diseases as well as improving root canal therapy and extending the lifespan of dental composites. Regarded by its developers as noiseless and painless, N-tap devices like that of DBD’s and APPJ’s will not only provide patients with a more comfortable environment but will ease feelings of apprehension. It may be years before N-tap technology reaches the offices of all dental professionals but it is clear that the day will come when the dental drill can be retired.

Although N-tap technology is proving to be a valuable asset to future dental practices, it does draw up some concern from a public health perspective. The most notable concern is whether or not such technology will have adverse health effects due to prolong occupational exposure of radical oxygen species and charged particles. Known to have carcinogenic effects, radical oxygen species and charged particles could result in a higher number of cancer cases among the dental profession. While there is no indication that utilization of N-tap technology among dental professionals has resulted in a higher number of cancer cases, it is still too early to tell. Nonetheless, it is vitally important for biomedical engineers and researchers to take into account
the possibility that utilization of such technology could result in harm to those who use it for extended periods of time.

In addition to the possibility of adverse health effects, N-tap technology has some limitations. For example, the technique is highly sensitive and would require dental professionals to obtain additional training. This could be costly and time-constraining. Another limitation that has presented itself is the cost of purchasing and maintaining the equipment. Since N-tap technology is rather new, there are a limited number of resources available and familiar with its engineering. As with any new technology, however, limitations such as these can be resolved given time.
Figures

Figure 1. (A.) Schematic of Dielectric Barrier Discharge and (B.) floating Electrode Dielectric Barrier Discharge (Hoffman C, Berganza C, Zhang J.)

Figure 2. (A.) Schematic of Atmospheric Plasma Jet and (B.) Plasma Needle (Hoffman C, Berganza C, Zhang J.)

Figure 3: Schematic of Plasma Pencil created by Laroussi et al. (Hoffman C, Berganza C, Zhang J.)

Figure 4: Diagram of the different components of a tooth (Hoffman C, Berganza C, Zhang J.)
References


AGE AND DISTRIBUTION OF STARS WITHIN THE MILKY WAY

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AGE AND DISTRIBUTION OF STARS WITHIN THE MILKY WAY

The Milky Way galaxy is filled with hundreds of billions of stars. From these stars, every single type of spectral class is able to be found. Using the H-R Diagram as well as taking a sample of the stars around us, the distributions of the types of stars in the Milky Way can be extrapolated. Making observations of the types of stars found in the Galactic Zones of our own galaxy as well as others, it can be determined that the Galactic Zones are populated by different types of stars.

The galaxy is filled with stars. About three percent of the Milky Way’s total mass is made up of stars, which then adds up to over one hundred billion stars in total. Not all of these stars are the same, however. When studying the Milky Way, it is clear to see that there are stars of every size, temperature, and age present. What is interesting, though, is the patterns that can be found in relation to age and distribution throughout the galaxy. It is crucial to understand that the location of a star within a galaxy relates to the age and composition of that star.

Stars are categorized into groups based on their temperature and color. There are seven spectral types of stars found in the universe: O, B, A, F, G, K, and M. O type stars burn the hottest; burning upwards of 30,000 K (Evans). These stars burn so hot that they are a dark blue in color, as they mainly emit blue-violet light. M type stars are the coolest types of stars. These stars are the coolest; having a maximum temperature of 3,500 K (Evans). These stars are a dark red in color and emit light into the infrared spectrum. Although all of the stars within a spectral type have similar temperatures and emit within the same light range, these stars may have extremely different sizes. Red dwarfs that are a thousand times smaller than our sun can be in the same spectral type as Red Giants, which may be hundreds of thousands times larger than the sun. All of these spectral types are found on the Hertzsprung-Russell (H-R) Diagram. This diagram shows the relationship between luminosity, temperature, absolute magnitude, and spectral class. How does this relate to the ages of stars, though?

On the H-R diagram there are several groups of stars: Main Sequence, Giants, White Dwarfs, and Supergiants. Throughout the lifetime of a star, that star will change so that is falls within almost all of those groups. When a star is formed, it starts as a protostar until it begins nuclear fusion. Once fusion has been engaged, that star is categorized as a Main Sequence star. It will spend a majority of its life in this stage. When that star runs out of hydrogen, it collapses and then expands into a Giant or Supergiant, depending on the original mass of the star, and starts fusion of Helium, Carbon, Oxygen, etc. (depending, again, on the size of the star). Once there isn’t enough fuel to maintain hydrostatic equilibrium, the star collapses and then explodes into either a planetary nebula (low-mass stars) or a Type II Supernova (large-mass stars). At this point, there is only the super-dense core left, which will become one of several things, depending on the mass. For low mass stars (Mass < 8solar masses), the
core will shrink into a White Dwarf. If the star is a high mass star (Mass $> 8$molar masses), it will either compress to become a Neutron star or a Black Hole (NASA, 2015).

Understanding the lifecycles of stars, it is then simple to recognize how old certain stars are by comparing what sequence they are in with what their mass is. The more massive a main sequence star, the shorter its lifespan will be. This is because smaller stars generate fusion at a slower rate, allowing them to burn for a much longer period of time. Therefore, a K-type main sequence star will have a lifespan of several billion years, while a B-type main sequence will have a lifespan of just a few million years. However, when looking at the ages of main sequence stars, there currently no way to determine their exact age. All stars have an estimate range of what their age could potentially be, but the exact age of stars can only be determined within a range of one billion years.

Although there is no exact way to pinpoint exactly how old a star is, there is a method that can be used to determine a relatively accurate age. All stars are made of hydrogen, a small amount of helium, and minute traces of the heavier elements which are called metals. In the early universe, none of the heavier elements that are around today were present as they had yet to be created by stars. As time went on, stars created these metals through atomic fusion. When those stars died, they turned into novae, infusing the early universe with metals (Mateer, 2015). This process means that the younger stars contain a higher percentage of metals than older stars. This then allows scientists to put main sequence stars into age groups. When classifying the stars by age, there are four categories that stars fall into: Population I stars, Intermediate stars, and Population II stars, and Population III stars.

Population I stars are the youngest stars in the universe, having been formed less than one billion years ago. These are very metal rich stars, with metallic elements being about four percent of their total mass. Intermediate stars formed between two and eight billion years ago. These stars’ mass range from four tenths to two percent metals (Reike, 2014). Sol is an Intermediate star, having formed roughly four and a half billion years ago and having a mass that is roughly two percent (1.78 to be exact) metal (Strobel, 2011). Population II stars are the oldest stars in the current universe. These are stars that formed over eight billion years ago. Because of the scarcity of metals during that period of the universe, these stars have less than four tenths of their total masses made up of metals. The last classification of stars is the Population III. These stars were created within ten million years after the Big Bang. There are currently none of these stars in existence anymore, as they would’ve gone through their life cycles and perished before the creation of the Milky Way. Being so ancient, these stars had absolutely no metals found within their masses. Population III stars were the first to infuse the universe with metals (Redman, 2007).

To be able to relate the age of stars to their distribution, a step needs to be taken back to look at the galaxy as a whole. The Milky Way galaxy is a large spiral galaxy about 100,000 light years in
diameter. Spiral galaxies are unique because of their complex structuring, and can be broken up into several different galactic zones. When looking at the galaxy from a side-on view, it is easy to see these zones: The Galactic Disk, the Bulge, and the Halo. These three zones, though all part of the same galaxy, have entirely unique properties and compositions. The same can also be said about the stars found within them.

The most obvious zone of the galaxy is the Galactic Disk; the enormous ring surrounding the center of the galaxy that lays along the galactic plane. Stars orbiting within the disk do follow the galactic plane, but exhibit eccentric orbits. Instead of being eccentric along the plane (like planets around their stars), their orbital eccentricities are perpendicular to the plane. The Galactic Disk is the relatively younger section of the galaxy. It can be divided up into several different sections: The arms, the thin disk, and the thick disk.

The arms are the youngest section of the galaxy overall. Composed mostly of gas and dust, the arms are the star factories of the galaxy. The stars found in this section are young Population I stars, a majority of them being under one hundred million years old. These stars all vary in type, but it is here that short-lived O and B supergiants can be found. A large amount of pre-main sequence stars can be found here as well, including T-Tauri stars and protostars (Grant & Lin, 2000). These stars are either found solo, or can be found in loose open clusters. These stars have orbital eccentricities that fall within four hundred light years of the galactic plane. Stars found within the arms account for less than one percent of the stars within the Milky Way. The next section is the thin disk. The thin disk is all of those stars that have orbital eccentricities within one thousand five hundred light years of the galactic plane. These are the older of the Population I stars, being as old as a billion years old. It is within the thin disk that primarily A and F type stars are present, as well as a few giants and White Dwarfs. The stars found in the thin disk account for about nine percent of the total stars in the galaxy. The oldest section of the Galactic Disk is the thick disk, which is primarily populated by Intermediate stars. These stars all orbit within three thousand light years of the galactic plane. This section is primarily populated by G, K, and M type dwarfs, but also contains giants, White Dwarfs, and planetary nebulae. This is the most populous section of the entire galaxy, contributing to about eighty-eight percent of the Milky Way’s stars.

Although not as aesthetically pleasing, the Galactic Bulge, or Bulge for short, is still an essential piece of the galaxy. The Bulge is the spherical center of the galaxy, surrounding what is thought to be a supermassive black hole. The Bulge is the brightest section of the galaxy because of the high density of stars located there. These stars are very interesting, as they have spherical orbits around the Galactic Center. While the stars located in the Galactic Disk follow a circular orbit following the galactic plane, the Bulge stars orbit along random planes in extremely elliptical orbits, creating a spherical effect. The stars located in the Bulge are primarily young
Population II stars, being older than eight billion years old, but younger than ten billion years. These are primarily cool, red dwarfs that are K, M, and occasionally G type stars, with a few giants found interspersed throughout the Bulge. This gives the Bulge a reddish tint when viewed from afar. The Bulge stars are low in metals due to their formation occurring before many metals were present in the universe.

Although the Bulge stars are old, the oldest stars in the universe can be found in the Halo. The halo is the spherical region that surrounds the galaxy, having the same diameter as the Galactic Disk. The Halo is primarily empty space and dark matter with a very small amount of stars found interspersed throughout it. The stars found in the Halo are ancient and are believed to have formed just shortly after the formation of the Milky Way. These stars can be found in either globular clusters throughout the Halo, or as rogue stars independently drifting. The orbits of the stars and clusters in the Halo are elliptical and random in the same fashion as the stars found in the Bulge. The stars in the clusters are all M dwarf stars, with a few K dwarfs mixed in. There are also White Dwarfs that are found within the globular clusters. All of these stars contain extremely low amounts of metals because of how old they are.

Looking at the zones of the Milky Way, it is apparent that every spectral type of stars can be found throughout the galaxy. When understanding where the types of stars are located throughout the galaxy, one may then ask what the ratios of those stars are. When observing the galaxy from earth, it is almost impossible to look at the entirety of the Galaxy due to distance and objects blocking our view. This makes it difficult to determine what the ratios of the stars are. To overcome this issue, astronomers simply take a sample of the stars surrounding our sun. When taking a sample of the nearby stars, astronomers then extrapolate their findings to assume that the stars represent that galaxy as a whole.

When a sample is taken, the ratios of the stars can then be extrapolated. The easiest stars to form are the smallest stars, because less material is used up in their consumption. This is reflected in the fact that M type stars are the most abundant, making up eighty percent of the stars present in the Milky Way\(^5\) (Powell, 2006). These stars are also the longest lasting, so there are red dwarfs that are anywhere from newly created to over billion years old. Over time, these stars have added up to the abundance that is present today. The next most abundant type of star is the K type, but these stars make up only eight percent of the total stars in the galaxy. The further left on the H-R Diagram, the less abundant the stars are (in terms of main sequence stars). This comes down to O type stars having an abundance of 0.00001% of the stars in the Milky Way. White Dwarfs are very abundant, making up five percent of the Milky Way’s stars. That shows about how many low mass stars have died in the Milky Ways’ history.

The Milky Way galaxy is filled with billions of different types of stars. These stars come in all all age groups, from newly formed to the dying remnants of ancient
stars. Understanding the ages of the different types of the stars, as well as their
distribution throughout the galaxy, is important for knowing more about the universe in
general. Knowing these things helps determine how old the Milky Way is, and even aided
in the determination of where the Sol system is located in the galaxy.
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The formulas with which to measure the mass and radius of stars

Measuring the mass of stars is difficult, but possible and since binary star systems are the most common, I will begin with methods for those systems. One method for a binary star system, is to use Kepler’s third law of motion 

\[ (m_1 + m_2) P^2 = (d_1 + d_2)^3 = R^3 \]

where \( P \) is the orbital period, \( m_1 \) and \( m_2 \) are the respective masses, and \( R = r_1 + r_2 \), and the "seesaw equation" for the center of mass:

\[ d_1 + d_2 = R \]

where \( R \) is the total separation between the centers of the two objects” (“Measuring the Mass of Stars,” n.d.). The individual distances of each star from the common point they orbit is used to determine the stars’ individual masses using this formula.

Typically, this method is difficult because we may not be able to map the orbits of the stars, and in these we can only find the sum of their masses and not the individual masses of each star, or just place limits on the masses instead of actually determining them.

For individual stars, their masses can sometimes be determined from the orbits of the bodies about them using

\[ M + m = \frac{4\pi^2}{G} \frac{a^3}{p^2} \]

where \( M \) and \( m \) are the masses of the two bodies, \( a \) is the semi-major axis of the orbital ellipse, and \( p \) is the orbital period in “solar units,” which is \( M + m = \frac{3}{2} \).

“The Hertzsprung–Russell diagram is the key to determining masses of individual stars. For stars on the main sequence, their properties are essentially determined by their mass” (“How do we know the masses of single stars?,” March 17, 2012).
Calculating radius is a lengthier process, as radius cannot be measured directly. “Even the largest star is so far away that it appears as a single point from the surface of the Earth - its radius cannot be measured directly. Fortunately, understanding a star's luminosity provides you with the tools necessary to calculate its radius from easily measured quantities” (“Calculating the Radius of a Star,” n.d.). The luminosity of a star is “given by the equation $L = 4\pi R^2 s T^4$, where $L$ is the luminosity in Watts, $R$ is the radius in meters, $s$ is the Stefan-Boltzmann constant ($5.67 \times 10^{-8}$ Wm$^{-2}$K$^{-4}$), and $T$ is the star's surface temperature in Kelvin” (“Calculating the Radius of a Star,” n.d.).
The calculation is a bit easier if we try to find the ratio of a star’s radius to that of our Sun, for which the formula is \( \frac{L}{L_s} = \frac{(4\pi R^2 s T^4)}{(4\pi R_s^2 s T_s^4)} = \left( \frac{R}{R_s} \right)^2 \left( \frac{T}{T_s} \right)^4 \), where \( L_s \) is the luminosity of our Sun, \( L \) is the luminosity of another star, \( T_s \) is the temperature of our Sun, \( T \) is the temperature of the other star, \( R_s \) is the radius of our Sun, and \( R \) is the radius of another star. “Solving for the ratio \( \frac{R}{R_s} \) yields \( \frac{R}{R_s} = \left( \frac{T_s}{T} \right)^2 \left( \frac{L}{L_s} \right)^{1/2} \)” (“Calculating the Radius of a Star,” n.d.).

The whole process is time consuming, and more than a bit complicated, but it helps when you have charts that can give you approximate information, like temperatures, just by looking at them. This table, for example, gives the approximate temperatures by looking at the b-v values.

<table>
<thead>
<tr>
<th>b-v</th>
<th>Surface Temperature (Kelvin)</th>
</tr>
</thead>
<tbody>
<tr>
<td>-0.31</td>
<td>34,000</td>
</tr>
<tr>
<td>-0.24</td>
<td>23,000</td>
</tr>
<tr>
<td>-0.20</td>
<td>18,500</td>
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<tr>
<td>-0.12</td>
<td>13,000</td>
</tr>
<tr>
<td>0.0</td>
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<td>0.15</td>
<td>8500</td>
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<tr>
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<td>3700</td>
</tr>
<tr>
<td>1.61</td>
<td>3000</td>
</tr>
</tbody>
</table>

“To find the ratio \( \frac{L}{L_s} \), we can use the absolute magnitudes of the stars. The magnitude scale is a logarithmic scale. For every decrease in brightness of 1 magnitude, the star is 2.51 times as bright. Therefore, \( \frac{L}{L_s} \) can be found from the equation \( \frac{L}{L_s} = 2.51^{Dm} \), where \( Dm = m_s - m \)” (“Calculating the Radius of a Star,” n.d.)
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Your Pharmacy is Radioactive

Written by: Alanda A Barash
April 16, 2015
Physics 112
Section 35884/35886
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Abstract: This research paper discusses the use of radioactive pharmacology to promote health with the distribution of radioactive pharmaceuticals. This paper discusses the need for radioactive drugs and the benefit of the radioactive medications that are being distributed for different types of therapy. The history of radioactive technology in the medical field through pharmaceuticals is highlighted throughout this research with some brief explanations about why radioactivity medication is important and its future in the advances of medicine.

If a stranger is asked, “What do you think of radiation?” More then likely, there will be a change in their facial expression. There may be a look of sadness or a frown and, typically it will be followed by a response about how dangerous and bad radiation is for mankind. Most people do not know that radiation is a part of daily life and comes in many forms. Some forms of radiation include the electromagnetic radiation that is emitted from radios, televisions, and microwaves. The common forms of radiation are usually ignored due to their every day use in life and the only radiation people tend to think or worry about is the radiation associated with what is heard about from cancer.

Radiation is used in the field of pharmacy. More specifically, radiation is used in nuclear pharmacy. Nuclear pharmacy is a specialty area of pharmacy practice dedicated to the compounding and dispensing of radioactive materials for use in nuclear medicine. In the early 1970’s nuclear pharmacy as a specialty area was developed. After nuclear pharmacy was developed, nuclear medicine was recognized. As specified previously, there are many types of radiation, but the radiation that concerns nuclear pharmacy and its research is the radiation that comes in the form of radionuclide.

Radionuclides are the excess energy that atoms “give off” when an atom has too many protons and neutrons in its nucleus and wants to return to its stable state. Radionuclide can come about in two different forms. The first form that these radiation particles exist in is the radionuclides that occur naturally. The second source of radionuclide is man-made radionuclide, which are created in labs. The naturally found radionuclide are the elements with an atomic number that is greater than 83. The man-made radionuclides are made in a particle accelerator. When a particle is in the accelerator, it is given extra protons and neutrons to cause a lack of stabilization of the nucleus. The particle wants a stable nucleus, so the extra protons or neutrons manipulate the particle into emitting radiation to allow it to stabilize itself. When the radiation is emitted, that is how the radioactivity can be captured.

The emission that is given off by the radioactive particle will determine whether or not the radionuclide will be useful for the treatment of a patient. Nuclear medicine uses small quantities of radioactive materials with a known type of emission to determine its specialty. The radioactive source gets “tagged” to some compound that is known to localize in a specific area of the body. When the “tagged” compound carries the radioactive material to the desired site a gamma camera detects the emissions given off by the radioactive material and creates images of the relative distribution of the radioactive source in the body.
In a retail setting, when a patient needs a prescription to be filled, they go to the desired pharmacy of choice and drop off a script written from the doctor. The pharmacist then, types, verifies, and fills the prescription based on what it reads. At nuclear pharmacies, it is almost very similar to the procedures of a retail chain. The pharmacist is responsible for obtaining the desired radioactive material, either from a manufacturer, or from an in house generator system. A common isotope that is used in nuclear medicine is Technetium-99m. Technetium-99m is usually available from a generator system. According to Purdue University, the way the isotopes are created is by the generator forming the radionuclides that are retained on an internal column until the generator is "milked". When "milking" the generator, sodium chloride is passed over the column, which removes the radioactive material. The eluate gets collected in a shielded evacuated vial. To make sure a certain “batch” of radioactive material is safe for patient use, it is tested and undergoes an accuracy exam and once that is completed, the eluate can then be used in the preparation of the final radiopharmaceutical products. A majority of the compounding is done behind leaded glass shielding and by using leaded glass syringe shields and lead containers to hold radioactive material. This ensures that the pharmacist that is making the nuclear compound is safe. After the compound is made and checked for its accuracy and effectiveness, it is labeled and then it is ready to be dispensed to the patient for use.

Another question a person could be asked is, “Who created such a field of pharmacy work? How did these people master a science using radiation to not kill a human, but help in the therapy of HEALING a person?” The answer to such a question is not as simple as picking a favorite president or deciding which candy bar can be eaten for dessert. The answer lies with the fine and humble beginnings with a French man who was a physicist and went by the name of Henri Becquerel. This man chose to experiment with uranium salts and start the research that lead up to the use of radioactive materials in drugs. The research of radioactivity continued with another scientist that went by the name of Marie Sklodwska Curie. She was able to discover different techniques about isolating radioactive elements. In 1972, University of New Mexico was the first to be recognized by the State Board of Pharmacy for its practice in radio nuclear pharmacology.

Biology, anatomy, and general chemistry are often courses that are associated with pharmacy degrees. It is often overlooked to have physics when obtaining a degree in the studies of pharmacy. It is believed for students to have basic and proper knowledge in all fields of science to determine a positive therapeutic plan for patients of all kind. A physicist will often tell a student that physics is the foundation of all science. Radioactivity has growth and decay. Understanding the growth and decay of radioactivity is crucial to understanding the formulation of the medications being created in radioactive therapy.

Nuclear decay is described with an equation that is similar to the equation of chemical reactions. Decay activity is the measurement of decay per time. A half-life of an isotope is when a population of radioactive material is decreased by exactly one half.
The equation that is often used for the description of exponential half-lives is indicated as: \( R(t) = R_0 e^{-\lambda t} \).

Three types of radiation are often used to compare all the radioactive particles that are used in the field of physics. Alpha particles are the most massive of the radiation particles. The absorption of these particles is so easy because of their larger particle size. Beta particles are less massive and move faster than alpha particles, which make them a lot harder to absorb.

The Pharmaceutical Research and Manufacturers of America reported that 646 medications were being developed for cancer in 2006. The cost to research and the cost to bring a new drug to the drug market have been estimated to be between $0.8 billion and $1.7 billion according to DiMasi and to Mullin in 2003. A lot of time is spent on research and trials to make a new drug to safely enter the market. For a drug to be entered to the market, it can take up to 12 years. Nuclear technology allowed it to be easier to develop new drugs since it aids in the advancement of trials moving along to pass along in the approval process of dispensing. The reaction of receptors in the body with radioactive medical imaging allows for scientist to gather accurate data for distributions of medications that are in the process of being created. The types of studies being conducted right now to develop new medications are related to labeling drugs of interest and molecular imaging for key processes. Medical imaging instrument companies have been aiding with one another in radiopharmaceutical development to create an advancement in radiopharmaceutical development. The companies that develop narcoleptic drugs, have had a heavy emphasis on studies directed at evaluating saturation of key receptors in vivo. These studies are being done because there has been a realization of doses being greater than those required to saturate the target receptor that needs to be attacked which results in more neurotoxicity. All these advancements in radioactive medicine becomes more beneficial for the public since there can be aid for numerous health conditions with these medicinal research drugs.

Radiation therapy uses high energy-radiation to shrink cancer cells. Not only can radioactive medicine be delivered to a human through a machine but also it can be delivered by having radioactive material placed nearby cancer cells. Radioactive iodine is an example of a radioactive way to kill cancer without the use of a machine. The toxicity of radioactive iodine is not harmful to the body but it is very harmful to cancer cells. Radiation therapy kills cancer by damaging DNA inside cells that carry genetic info that is usually passed onto offspring. When the cancer has been destroyed the body takes care of the broken down cells by its own natural processes. Not all patients that receive radiation therapy will cure their cancer condition.

Radiopharmaceuticals can be defined by the cancer society as “drugs that contain radioactive materials called Radioisotopes.” These types of medication can either be given by oral route or venous injected route. The way of absorption of each medication is different depending on the medication being used and the type of cancer being treated. Strontium 89 which is the chemical used in the drug called Metastron®, samarium 153 which is the chemical in Quadramet®, and Radium- 223 which is known as Xofigo® are
all radiopharmaceuticals that can be used for tumors that have spread to the bones \(^6\). These types of medications are usually given to patients though an IV so they can enter the bloodstream as fast as possible \(^6\). The radiation of these medications kills the cancer cells which provides relief of pain for the patients suffering with the disease.

Not only do these drugs have an affect on bone cancer but it affects many other cancers as well. Some examples of other cancers these drug can treat includes thyroid cancer as mentioned previously \(^6,5\). Iodine 131 is the chemical name of radioiodine that is used to treat thyroid cancer. Thyroids cancer absorbs all the iodine in the blood and that why the radioiodine is used to destroy the thyroid gland. It is used to treat any thyroid cells left behind after surgery or any cancer that may have spread to the lymph nodes and other various parts of the body \(^6\).

Phosphorus 32 is used to damage hollow brain tumors without hurting the “good” and healthy parts of the brain \(^6\). (hollow) to kill the tumor without hurting the healthy parts of the brain. This element is also used to treat two other types of cancer; ovarian cancer and polycythemia vera which is classified as a blood disease \(^6\).

For some cancer conditions, it could be very useful to destroy malfunctioning cells with radiation. The radioisotope that generates radiation could be localized in the designated organ through a radioactive element following its usual biological path \(^7\). Beta radiation is usually the cause of any destruction to damaged cells in therapy \(^7\). The ideal therapeutic radioisotope would be a beta emitter with a significant enough amount of gamma rays to enable medical imaging \(^7\). Some examples of these radioisotopes would include: lutetium-177. This isotope is prepared from ytterbium-176. After the ytterbium-176 become Yb-177, it then decays very quickly to Lu-177. Yttrium-90 is used to treat lymphoma and liver cancer \(^7\).

An experimental procedure is taking place in this day and time that uses boron-10. This isotope concentrates INSIDE the tumor. The patient is irradiated with a neutron that gets absorbed by the boron in order to produce alpha particles, which kill the cancer. According to research done by world-nuclear.org. Another radionuclide used in cancer therapy is lead-212, which has a half-life of 10.6 hours, could be attached to monoclonal antibodies for a therapeutic treatment \(^7\). The alpha decays of Bi-212 and Po-212 are active in destroying cancer cells over the course of a couple of hours \(^7\).

Another interesting point that can be made about radioisotopes is that they do not only help cure problems that arise from cancer. The radioisotope Erbium-169 has a half-life of 9.4 days and that isotope is used primarily for relieving arthritis pain in synovial joints \(^7\). In a study that was conducted to treat rheumatoid arthritis, Erbium—169 was compared with triamcinolone hexacetonide for topical treatment of 32 patients. Erbium—169 was injected into 83 joints and triamcinolone hexacetonide into 54 joints \(^8\). Both treatments provided better grip strength for the patients and alleviation of some pain and swelling. At every check-up percentage of remissions was higher after triamcinolone hexacetonide injection than after erbium—169 \(^8\). Therefore, this study concluded yet another good use for a radioisotope for the health and healing of patients globally.
Xenon-133 is used for pulmonary ventilation studies. This isotope has a half-life of roughly 5 days\(^9\). This particular isotope is used for VQ imaging and helps get better visuals of the human chest cavity. If this gas is inhaled it can be used to access the cerebral spinal fluid\(^9\).

As it has been seen, physics has played a very important role in the field of medicine and pharmacy. The advancing of radioisotopes has created another branch to pharmacology that is advancing daily. New drugs are always being created and new methods are being implemented to find better and more efficient ways to care for patients. Each disease now is being able to receive treatment with some sort of radioisotope.

In conclusion, with all this research and all the facts that are presented I think there can be a cure for cancer sometime into the near future. Radiation can cause cancer but it can also be manipulated enough to cure the disease. By creating a drug that can manipulate a cancer cell and allow the cell to stop replicating and progressing its damaging ways in the human body shows the great advancements that have just been made in the last decade. Different isotopes of elements are being used to be injected into the human body to target cancerous cells can cause patients to eventually be terminally ill. The research of the radioactive drugs is still being done and will not take a break any time soon.
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The Development of Wafers from a Silicon Ingot

Tyler Becker

4/23/15

PHY111 – College Physics I

Dr. Casey Durandet
Abstract

How was silicon discovered? How is impure silicon turned into wafers for electronic use? Why does it matter? Silicon is the main reason most electronics on Earth can function. Many have wondered what makes a computer function or makes a video game console more than just a plastic box. The answer is silicon-based microelectronics. This paper will discuss how silicon can transform from simple sand on a beach to being the cornerstone of a multi-billion dollar industry.

History of Silicon

The discovery of silicon was a puzzling situation for scientists. Even though silicon is the second most abundant element in the Earth’s crust, it was not thoroughly studied until the late 1700’s. Silicon is mainly found in the compound silica that was originally thought to be its own element. In reality silica is the compound Silicon Dioxide (SiO2). SiO2 was considered to be a unique element because the bonds were extremely difficult to break, even with the best methods of the time. Antoine Lavoisier was the first scientist to proclaim that silicon was its own element. His proclamation was found to be correct when a scientist by the name Jöns Jacob Berzelius finally was able to separate silicon and oxygen in 1824. He did this by reacting silicon tetrafluoride with potassium to produce potassium fluoride and silicon. The formula for this reaction is SiF₄ + 4K → 4KF + Si. The potassium fluoride and silicon then needed to be chemically separated and this was done by adding water at room temperature. The water reacts with the potassium fluoride to make a liquid byproduct of hydrofluoric acid and potassium hydroxide. The formula for this reaction is KF + H₂O → HF + KOH. This reaction results in behind pieces of pure silicon. This process is now outdated with the technological advances of today, but this was still an important part of electronic history.

The Making of Semiconductor Grade Silicon

The modern day process of getting semiconductor grade silicon is a lengthy but necessary part of wafer manufacturing.

Metallurgical Grade Silicon

The manufacturing of silicon utilizes silica (SiO₂) as the raw material. Silica is most commonly found in sand and quartzite. Between the two options, quartzite is purer than sand. The first step to getting to pure silicon is converting the silica to metallurgical grade silicon. This silicon is a crude silicon that is about 98% pure. While an average person may think that is extremely pure, it has too many impurities for semiconductor processes. Impurities include metallic and organic contaminants. This silicon, however, would be sufficient for industrial use. The raw materials for this step are silica and charcoal (C). The two raw materials are placed in an electric arc furnace to be melted down. An electric arc furnace is a machine that creates an arc by
sending an electrical current between two electrodes that are placed in the raw materials (see figure #1). This process occurs at about 1800ºC. The chemical formula for this process is, \( \text{SiO}_2 + 2\text{C} \rightarrow \text{Si} + 2\text{CO}; \Delta H_{2100} = +695 \text{ kJ} \). Since this silicon is not pure enough for electronics, it needs to be processed further.

**Making Trichlorosilane**

The chunks of metallurgical grade silicon are first ground to small pieces. The small pieces of silicon are shot into a fluidized bed reactor along with hydrochloric acid. The silicon and hydrochloric acid react chemically to create trichlorosilane and hydrogen gas. The formula is, \( \text{Si} + 3\text{HCl} \rightarrow \text{SiHCl}_3 + \text{H}_2; \Delta H_{298} = -218 \text{ kJ} \). The hydrogen gas is a byproduct that is released into the air while the trichlorosilane is kept as a liquid at room temperature. The trichlorosilane is then distilled to clean it. Distilling is the process of boiling the liquid to a gas and then cooling it back down to a liquid in a different container.

**Trichlorosilane to Polycrystalline Silicon**

The next process takes place in a device called a quartz bell jar. The structure is an upside down jar with an exhaust vent and a u-shaped pure polycrystalline silicon seed inside it. The pure silicon seed is heated to 1150ºC using an electric current. The jar is placed on top of a reservoir of liquid trichlorosilane that is connected to a supply of hydrogen gas. The hydrogen gas is pumped into the reservoir of trichlorosilane, and the mixture is heated to a gaseous state. This gaseous solution rises into the bell jar with the pure silicon seed, and the reaction begins. At this high heat the trichlorosilane and the hydrogen will react which gives the chemical formula, \( 4\text{SiHCl}_3 + 2\text{H}_2 \rightarrow 3\text{Si} + \text{SiCl}_4 + 8\text{HCl}; \Delta H_{1400} = +964\text{kJ} \). When this reaction occurs, the pure polycrystalline silicon undergoes deposition and grow on the poly-Si seed. A deposition is the chemical process where a gas transforms into a solid. This process will continue until the ingots have grown to the specified diameter for the customer, this normally takes about a week to grow fully. The poly-Si is now 99.99999999% pure and this means that there is only one foreign atom for every 10 billion silicon atoms. It is now semiconductor-grade silicon. Every year, there is approximately 20,000 tons of pure poly-Si produced around the world (refer to figure #2). The poly-Si ingots are now crushed into multiple sized pieces and shipped to a variety of industries ranging from microelectronics manufacturers to solar panel manufacturers.

**Ingot Growth – Czochralski**

The primary growth process in silicon ingot manufacturing is the Czochralski method. This method was developed by Jan Czochralski in the 1917. He discovered this phenomenon by accident while he was trying to dip his pen into his inkwell and accidentally put it into a crucible of molten tin. When he pulled the pen out the tin stayed on the pen like a thread and after further testing he realized it was a single crystal. This discovery led to the experimentation in the field of semiconductor technology, and when scientists eventually that silicon acted in a similar manner,
they knew they had discovered the way of the future. There are many steps to this method that need to be precisely done (refer to figure #3).

**Preparation and Melting**

The process starts with a quartz crucible. The quartz crucible holds the polysilicon discussed earlier and a substance called “dopant.” Dopant is an element that is added to modify the electrical characteristics of the silicon, such as resistivity and the type of charge that can go through the wafer. If more dopant is added, then the resistivity goes down because it creates a better conductor of electricity. There are two different types of dopant: p and n dopant. The p dopant is made up of the elements in Group III but primarily boron, aluminum, and occasionally gallium. These elements are lacking electrons in their valence shell which, when added to the silicon, will help electricity move smoothly through the material by lowering the resistivity. The n dopant is made up of the elements is Group V but primarily phosphorous, arsenic, and antimony. These elements have extra electrons in their valence shell which also lowers resistivity. Every ingredient that is placed in the crucible is positioned precisely where it needs to be, meaning that operators need to reduce the amount of empty space in the crucible as much as possible to allow for more product. While it is true that grinding up all of the poly-Si would produce the densest load in the crucible, there has to be a mix of small pieces and big pieces to control how fast everything will melt. During a site tour of SUMCO Phoenix, Dr. David O’Farrell strongly emphasized the importance of this key point. After the pieces have been carefully placed in the crucible, it is then heated to 1420 ºC which is the melting point of silicon. When the silicon is melting the crucible walls will melt a certain amount, adding oxygen as an impurity. Though the amount of oxygen is small, it has some positive benefits such as: making the wafers stronger for future processes. The true pulling process starts after the silicon is completely melted.

**Pulling**

The ingot shape is determined by the seed. The seed is a square piece of monocrystalline silicon that is hanging on the puller rod. The seed is the only thing that holds the ingot throughout the entire pulling process. Although the seed can be reused up to five times, it costs about $550. The machine lowers the seed crystal into the molten silicon with the dopant and begins to spin. While the seed is rotating and pulling, the crucible also rotates in the opposite direction. This spinning creates the cylindrical ingot that is seen at the end of the process. During the pulling of the crystal, the rate at which it is pulled is regularly monitored to keep a consistent diameter. The average pulling speed is about .5-1mm/minute. Note that the dopant is not distributed equally throughout an ingot; the bottom is always more concentrated which means that it will have a lower resistivity than the top of the ingot. The finished ingot displays four distinct lines around the ingot called node lines. These lines are based on the orientation of the seed since the seed is square the ingot needs to represent that shape in its final form. Because a
wafer needs to be a circle, these node lines are ground off in a later step. At the end of the pulling process, the ingot is about 2 meters in length and is about 160 kilograms in weight. The entire weight of the ingot is held by the small seed, the weight is found by multiplying the mass times the acceleration of gravity \((9.8 m/s^2)\). This means that a seed weighing about 1.67N can hold up an ingot with a weight of 1,568N. There are many diameters and doping levels that the ingots can be grown to depending on customer specifications. The typical sizes are 150mm (6 inch), 200mm (8 inch), and 300mm (12 inch). Companies are also working on the development of a 450mm (18 inch) ingot. The variation in sizes determines the amount of chips that can be produced by a single wafer. The entire pulling process takes about three days. After the pulling is complete, the ingot will sit in the furnace and cool for about three hours. The furnace is then opened and the ingot is cut from the seed and cooled to room temperature for another three hours. This marks the end of the Czochralski method and, after the ingot is completely cooled, the ingot goes on to the next step.

**Ingot Growth – Float Zone**

Float Zone is the second most common method of crystal pulling. Float Zone was a method developed by Siemens in 1953. They gave the license for the process to Wacker Burghausen in the 1960’s and Wacker mass produced ultra-pure silicon from their factory. This method is in some ways similar to CZ. FZ methods are typically used to grow ingots with high resistivity. This process starts with a polycrystalline silicon ingot that was made in the quartz bell jar discussed earlier. This ingot is placed vertically in the float zone machine. Under the ingot, there is a monocrystalline seed that is pointed up towards the bottom of the ingot. The ingot is heated by a high-frequency coil and brought into contact with the seed crystal. By slowly lowering the ingot, the coil will melt it from bottom to top. Just like with CZ the seed determines the shape of the new monocrystalline ingot. This process takes between 10-15 hours to complete, not including cool down time. When the ingot is finished, it will go through the same manufacturing steps as the ingots made from CZ.

**Ingot Shaping**

The next step in the wafer manufacturing process is to shape the ingot and prepare it for slicing. The first step of ingot shaping it to cut off the seed and tail cone. These pieces are cut up and used in the CZ method again for another ingot, this is known as remelt. When remelting ingot an issue arises, there is a certain amount of dopant in the pieces from the cones that have to be taken into consideration. When the operators are making up the batch for the new ingot, they need to make sure that they only add a certain amount of extra dopant to compensate for the remelt pieces. After the cones are cut off, the ingot is cut into about five segments called “blocks.” From these blocks, slices that are 1.2mm in thickness called “slugs” are cut off and sent to the quality control department to be tested for quality inspection and metrology. Each block will produce about 300 wafers. The individual blocks now need to be ground to the
particular diameter that was ordered (refer to figure #4). Grinding also removes any imperfections on the outside of the block, including the node lines. Most orders come with a request for a notch or flat to be put on the side of the block, preferably on one of the node lines. These marks are used for setting and alignment purposes. Flats are used on the wafers that are below 200mm (8 inches) in diameter. Notches are used on wafers that are 200mm (8 inches) and above. The next step is where the wafers start to take their shape.

**Ingot Metrology**

The slugs discussed in the previous segment now need to be tested for their quality. The two primary measurements that are conducted on the ingot slug are infrared absorption and four point probe analysis. The IR absorption method measures interstitial oxygen and carbon contamination. The absorption of infrared light, and the resulting peaks can be correlated to oxygen and carbon levels within the silicon. The four point probe analysis is a standard metrology tool to measure sheet resistance. The sheet resistance directly correlates to the dopant concentration of the ingot. The four probes are placed side by side on the slug with the outer probes passing a current through the charged silicon creating a voltage in the inner probes. Using Ohm’s law the resistivity can be calculated with the formula: $\rho = \frac{R_{ohm} \cdot (s_1 \cdot s_2)}{t \cdot \ln(2)}$. The resulting sheet resistance, oxygen, and carbon data can be entered into software to ensure that the process is stable, consistent, and meets customer specifications.

**Wafer Slicing**

The next step is sawing/slicing. There are two ways companies do this step: ID-sawing and Multi-Wire Slicing. ID sawing, also known as inner diameter sawing, is the process of sawing wafers one by one with a saw. The saw blade itself is a 300µm thick chrome nickel steel sheet coated with small diamonds. Diamonds are the hardest naturally-occurring substance on the Earth and help the blade cut through the silicon with ease. This early process was the most common way to cut blocks to wafers but, due to issues with warping of the wafers and time inefficiency, a new sawing technique was invented. MWS, also known as Multi-Wire Slicing, is the process of slicing a block into all of the wafers in one run (refer to figure #5). Multi-Wire Slicing is done by running a single wire that is about 180µm in diameter over four metal rollers at about 10m/s. The block is put in between the top two rollers and is put down through the wire. The wire itself does not do any of the cutting; this is done by a semi liquid mixture called “slurry.” Slurry is a semi liquid that contains silicon carbide in an oil-based mixture. This process has proven itself to be a more efficient mechanism for cutting the wafers. While the process does take some time, about 8 hours for a 200mm (8 inch) block, it produces better wafers with a much lower percentage of irreversible warping. Once the blocks are cut into wafers, they are sent to the next processes to remove damage and polish.
Wafer Grinding

After all of that rough slicing, the wafers must now get ground to remove sawing damage. Grinding can be done one of two ways, SSG (Single-Side Grinding) or DDG (Double-Disc Grinding). SSG is the process in which the wafer is clamped by vacuum on a ceramic chuck that rotates the wafer during the grinding operation (refer to figure #6). This process has to be done two times because it only does one side of the wafer at a time. The alternative option is DDG which is where the wafer is placed in between two grinding pads, and both sides are simultaneously ground. This process is faster than SSG because it does twice the work in the same amount of time. The pad of the rotating grinding wheel can be customized for abrasiveness by the customer for a coarse or fine grind. The wafers are now somewhat smooth and are ready for the next step.

Wafer Edge Treatment

Before the wafers can endure any other treatment, they must get beveled. This step is essential to the protection of the wafers. The wafers are brought to a machine that has a grinding wheel with an edge that has inverted half-moon shape. The wafer’s edge is pushed up against the half-moon shape to shave off the edges to create a rounded edge. Beveling helps the wafer by reducing the chance of chipping, which later results in wafer breakage. It also has uses for a different process that will not be discussed in this paper called epitaxy. Now each wafer needs to be able to be identified individually; this is addressed in the next step.

Wafer Scribing

Wafer scribing is the part where each wafer gets its own identity inscribed onto the wafer. This process is called laser marking. A high energy pulsed laser engrains a code into each wafer using dots. The code that is placed on the wafer may contain an identification number, vendor, resistivity, dopant concentration, crystal orientation, and check characters depending on customer requirements. The codes can either be a deep mark (20 - 40 µm) or a soft mark (1 - 4 µm). Each code is checked by a technician using a microscope to ensure that it was put clearly on the wafer. The next step is one of the most important processes to ensure product quality, known as wafer lapping.

Wafer Lapping

The next step in the wafer manufacturing process is called lapping. Lapping is a process that improves the wafer’s flatness and TTV (Total Thickness Variation). Multiple wafers are moved into lapping carriers between upper and lower steel lapping plates. The plates rotate in opposite directions to create a fine grinding action. Excess silicon on the surface of the wafers is
taken off by the combination of the lapping plates and the semi liquid slurry. This is the same slurry as described previously, but the slurry used here is considered medium grit. This process will produce what is considered a rough polish. Levelness is critical when it comes to wafer processing. For scale, if the wafer were the size of the state Oregon (about 300 miles across), the maximum deviation from a flat would only be 5 feet \(^4\). The now super flat wafers are taken to the next step.

**Wafer Etching**

The wafers are now brought to etching. Wafer etching is a chemical process that reduces more of the damage done by the previous mechanical steps and removes particles lodged in any cracks of the wafer (refer to figure #7). There are two different methods to this process: Caustic Etching (Anisotropic etch) and Acid Etching (Isotropic etch). Caustic etching, using KOH, means there will be different etch rates in different directions in the material. The chemical formula for this reaction is: \(\text{Si} + 2\text{KOH} + \text{H}_2\text{O} \rightarrow \text{K}_2\text{SiO}_3 + 2\text{H}_2\). Acid etching, HF/HNO\(_3\), results in the same etch removal in all directions. The chemical formulas for this process are: \(\text{Si} + 2\text{HNO}_3 \rightarrow \text{SiO}_2 + 2\text{HNO}_2\), \(2\text{HNO}_2 \rightarrow \text{NO} + \text{NO}_2 + \text{H}_2\text{O}\), and \(\text{SiO}_2 + 6\text{HF} \rightarrow \text{H}_2\text{SiF}_6 + 2\text{H}_2\text{O}\). During this process, typically 15-20\(\mu\)m are removed from each side of the wafer, and the geometry of the wafer is still kept constant.

**Wafer Polishing**

**Edge Polishing**

Edge polishing is applied to smoothen the rim of the wafer to avoid particle generation. While the edges may have been ground previously, the wafer has gone through many processes and needs to have a final polish to make sure it stays rounded \(^10\). Edge polishing is done by placing the wafer towards the pad of a rotating polishing drum.

**Surface Polishing**

The first step in polishing the surface of the wafers is sorting them by their thickness. Sorting is important because when the wafers are placed in the polishing machine the polishing pads need to be level on all of the wafers. The wafers are wax mounted onto a polishing carrier plate. This process is nearly identical to lapping except that the polishing typically only happens on one side, and the medium grit slurry is replaced by a fine grit slurry (refer to figure #8). This fine grit slurry will work to give the wafers a mirror finish. There is a process where both sides of the wafer get a mirror polish, but this is for devices that need the utmost levelness and smoothness \(^9\). Single-sided polish is the most common finish because most devices only utilize one side of the wafer. Surface polishing is the final mechanical process performed on the wafer. The wafers need to be inspected thoroughly in order to be shipped.
**Final Wafer Inspection**

There is an entire science and specialty dedicated to the measurement of critical semiconductor properties, commonly called metrology. For this discussion, simply 100% of wafers need to be measured for their resistivity, flatness, and particles. Resistivity is measured with a contactless resistivity gauge using a method called eddy current evaluation. Flatness of the wafer is measured using a capacitive measurement, which analyzes multiple points on a wafer and determines an overall flatness profile (refer to figure #9)\(^3\). Finally, particles are measured using a laser diffraction tool, which can bin particles into different sizes and also count the particles.

**Wafer Cleaning**

Wafers are cleaned in two different solutions known as SC1 and SC2. SC1 is composed of \(\text{H}_2\text{O}_2\), \(\text{H}_2\text{O}\), and TMAH (tetramethyl ammonium hydroxide). SC2 is composed of HCl, HF, and \(\text{H}_2\text{O}\)\(^3\). These baths are to ensure that there are no remaining foreign chemicals or particles on the wafers (refer to figure #10). The wafers are then dried and checked by hand in a dark room before shipping.

**Wafer Packaging**

Finally, wafers are now ready for shipping. The wafers are placed in holders called cassettes. Cassettes are high-quality plastic containers with multiple slots to hold each wafer. After the wafers are carefully placed in the cassette, they are wrapped in a transparent bag followed by an aluminum moisture proof bag. The bags are then vacuum sealed, which draws out the air to reduce contamination exposure\(^3\). These wafers are then sent to the customer for further processing.

**What are the wafers used for?**

Semiconductor companies such as Motorola, Infineon, Intel, IBM, Texas Instruments and many others use the silicon substrates described above and continue processing them to make very specific devices. These devices are then packaged and installed into many different applications. These applications include automobiles, household appliances, air handling equipment, airplanes, weaponry, satellites and the list continues for many other products. In addition to electronic devices such as diodes, transistors, integrated circuits, and memory these wafers can also be used in solar panel manufacturing. Solar panel manufacturing has to do with making photovoltaic cells from the wafers. There are many different industries and manufacturers that rely on silicon-based products that it would be impossible to list them all.
Interview with Mr. Becker, Infineon Director of Engineering/Manufacturing

An opportunity arose to interview Chandler Becker, the current Director of Engineering and Manufacturing at Infineon Technologies in Mesa, Arizona. The interview was conducted April 1, 2015. As a child, Mr. Becker was always interested in science and medicine. Even in pop culture, his favorite TV show through high school was Star Trek the Next Generation. His skill set in high school revolved around math and science and his father encouraged those interests significantly. Finally, his father was a Mechanical Engineer from the University of Michigan and was very strict towards grades, particularly in math and science.

Mr. Becker was a pre-med major for his first two years of college. However, he changed his major during his junior year because of family commitments and decided to earn a Bachelor’s of Science in Chemistry from Arizona State University. During his sophomore year, he interviewed for and was awarded a job at Motorola as an entry level operator in their SMO epi facility. That fab eventually closed due to business conditions, but he transferred to Motorola’s Gallium Arsenide (GaAs) full fabrication facility. “This facility is where I truly started to develop into a skilled technical worker” stated Mr. Becker.

He became a process technician in the wet etch/cleans and gold plating department, which fit his major perfectly. He then continued to work as a technician until graduation and promptly received a job as a Process Engineer responsible for Gold Plating and back-end wet process. According to Mr. Becker, his first major project was to develop an analytical laboratory to measure key parameters of the sodium gold sulfate plating solution. This involved wet chemical titration, AA/MS spec and UV/VIS techniques. “The lab provided me an opportunity for my first published paper and presentation at an international conference” said Mr. Becker.

Unfortunately, the GaAs fab was to be closed and Mr. Becker was forced to make a life decision on what he wanted to do with his career. The two options were to either go back to school to become a pharmacist or return to school to earn Masters of Business Administration to strengthen his position in the current field. The costs and time associated with pharmacy school were very high relative to the expected payback of an MBA education. Therefore, together with a new job at International Rectifier, Mr. Becker decided to return to school to become a technical manager. Initially at IR, Mr. Becker was a senior process engineer but soon began to take on additional responsibilities. First, he was assigned the Metrology group, then added the Research and Development team and eventually took over the entire Engineering group. Then, two years later, Mr. Becker was promoted to Director of the entire Manufacturing operation. He has held that position for just over two years.

Mr. Becker was then questioned what types of credentials are typically required for his current job and also for any position that may report into his department. The positions that currently report to Mr. Becker are: Cleanroom Operator, Process Engineering Technician, Metrology Technician, Process Engineer, Metrology Engineer, Development Engineer, and
Process Engineering Manager. For the operator, a high school diploma is all that is needed. A Technician typically requires an Associate’s Degree or equivalent college credits. Any of the Engineering positions will require a minimum of a Bachelor’s degree in engineering or a physical science, such as chemistry or physics. In order to become an Engineering Manager, you typically require approximately ten years’ experience as a Process Engineer and good skills as working with people as a team lead. Finally, in order to be a Director, a Process Engineering Manager must have shown good results as a manager and also have earned a Graduate Level Degree. An additional recommendation that Mr. Becker made was that a very useful certification is Six Sigma Greenbelt or Blackbelt, which is a specialization in Lean Manufacturing, Six Sigma troubleshooting and project leading, including Statistical Process Control.

In his current role, Mr. Becker is responsible for a wide range of responsibilities throughout the operation. These include monitoring process stability and performance using Statistical Process Control (SPC). Meeting manufacturing output requirements and developing Key Performance Indicators (KPI). Mr. Becker also sponsors and leads research and development projects (R&D) and manages approximately a $6 Million annual budget for Repair and Maintenance (R&M), headcount, facility, raw material and other costs. He also is responsible for hiring, training, rating, and mentoring the more than 60 employees within his departments. Another critical responsibility of the Engineering and Manufacturing team is to ensure product quality by maintaining fab certifications, supporting customer quality concerns, work with supplier to ensure their product meets corporate standards, and maintain internal fab yields. Finally, Mr. Becker also works with corporate managers to develop long-term strategies and integration a new corporate structure, merging International Rectifier with Infineon.

A final question for Mr. Becker was how the semiconductor industry changed since he started in 1997. Mr. Becker’s initial response was “In the 18 years that I have been involved in the semiconductor industry, there have been many highs and lows.” In regards to technology, the industry has been improving year on year since its inception. For example, in the power industry, which is what he is most familiar with, the defect control levels have been tightened significantly. “When I was processing epi at Motorola as a student, we wore smocks, hairnets, and foot booties. The defect measurement solely revolved around bright light inspection and we counted by sight how many particles were on the wafer. If there was an epi spike, which is formed by epi growth on top of a large particle, we would simply use a modified Dremel type drill to polish off the spikes and then perform a wafer scrub.” This was performed in a Class 1000 or worse cleanroom. Today, a similar technology requires a Class 10 or better cleanroom where epi particles are measured with automated metrology and epi spikes do not even occur. Also, silicon technology in power electronics will eventually reach a physical limit. Therefore, new technologies such and GaN, GaAs, and SiC are being developed. On the bad side, the silicon industry can be quite unforgiving to the employees within it. The industry is prone to major peaks and valleys that result in massive hiring and equal rates of headcount reductions.
He stated “I have been involved in three major layoffs that have been extremely difficult on many levels. I have seen my good friends lose their jobs and even had to lay some of them off myself.”

In conclusion, Mr. Becker has worked very hard to develop a strong career in the semiconductor industry, and aside from some difficult downtimes, the industry has provided excellent advancement and earning opportunities, while being continuously challenged.

**Conclusion**

The beginnings of the semiconductor industry were initially very slow to develop but have made great strides over the past several decades. Today, semiconductor devices can be found in almost every aspect of a modern home. Silicon is used in lightbulbs, satellites, everything in between. In order to start the silicon manufacturing process, it is critical to begin with ultra-pure, ultra-flat, and ultra-clean silicon substrates. The many processes required to make those wafers is extremely complex and expensive. Many careers have been dedicated to only one individual step of the process and this paper only discussed the basics of those processes. Throughout this paper, multiple Professional technical experts were interviewed and provided great insight to this exciting industry. Thank you to both Dr. O’Farrell of SUMCO and Mr. Becker on Infineon for their input to this paper.
Figures

All pictures are from reference #4

Electric Arc Furnace

Silicon Manufacturing Process

Figure #1

Figure #2

Czochralski Process

Figure #3
Grinding
Figure #4

MWS
Figure #5

Surface Grinding
Figure #6

Etching
Figure #7
Surface Polishing
Figure #8

Flatness Gauge
Figure #9

Cleaning Automat
Cleaning
Figure #10

Me at SUMCO with an ingot
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Hubble’s Law and the Hubble Constant

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2015
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Hubble’s Law/Constant is arguably one of the most important cosmological discoveries ever made. A constant and theory developed in 1929 by Edwin Hubble, it states that the universe is expanding constantly. Hubble’s law describes the expansion of the universe outward as evidenced by their red shifts. It also describes the velocity and distance of the galaxies moving away from each other and required astronomers to create new and more dynamic models of the universe.

Made possible in part by Vesto Slipher’s measurements of the apparent radial velocities of nebulae, Hubble’s law was primarily developed by Edwin Hubble and his estimates of distances to nearby galaxies (Huchra 2008). Even though there were papers published by Georges Lemaitre and H. P. Robertson using Hubble’s data before his 1929 discovery, Hubble deserves the credit for the discovery because it was his systematic program of measuring galaxy distances that finally convinced the community at large. Hubble’s law forced cosmologists to create dynamic models of the universe and also implies the existence of a timescale or age for the universe.

Almost as soon as Hubble announced his discovery, he was met with many different types of low level controversy. At the time, Hubble’s initial value for the expansion rate was about 500 km/sec/Mpc, with the expansion age of the universe determined from this being 2 Gyr. However, by the 1930s radioactive dating of rocks had already shown geologists that the age of Earth was 3 Gyr. In 1932 as well, the astronomer Jan Oort contradicted Hubble and his scales in a paper that he wrote. He did this because at the time the scale of the Milky Way was fairly well established, yet Hubble’s calibration implied that the Milky Way was far larger than any other nearby galaxy, but the astronomical community continued to support and use Hubble’s value. A solution came in the form of a combination of effects in the 1950s with Walter Baade’s discovery of Population II stars and his recalibration of the period-luminosity relation for population I Cepheid variables (The Hubble Constant). This also gave rise to a perplexing problem, where what Hubble had thought were individual stars in distant galaxies turned out to be star clusters instead. This meant that Hubble had not been observing standard candles, objects whose absolute luminosity does not vary with distance.

In the 1920s, Edwin Hubble used the new telescope at Mount Wilson Observatory to observe and detect variable stars in several nebulae. At the time, nebulae were a heated discussion topic, posing the question of whether they are interstellar clouds in our own Milky Way galaxy, or whole galaxies outside of our own. This question was difficult for scientists to answer because at the time it was extremely difficult to measure the distance to other astrological bodies for lack of a reference point. The variable stars that Hubble had observed had a characteristic pattern resembling that of a class of stars known as Cepheid variables. In earlier years, a female astronomer had proved that there was a correlation between the period of a Cepheid variable and its luminosity. By measuring the period and luminosity of these stars, Hubble was able to show that the nebulae were not clouds within our own galaxy, but external galaxies beyond the edge of our own (NASA 2014).

Hubble’s law states that the distant galaxies we see are constantly moving away from the Earth. However, this fact does not imply that we are the center of the universe, as all planets and galaxies would see all other planets and galaxies moving away from one another. Hubble’s law is a direct correlation between the distance to a galaxy and its velocity as determined by a red shift. Hubble discovered by comparing measurements of the Cepheid-based galaxy distance determinations with measurements of the relative velocities of galaxies that galaxies that are
farther away from us are also moving away more rapidly. Hubble’s discovery marks the beginning of the modern age of cosmology (Hubble’s Law).

The velocity of a galaxy can be expressed mathematically by the equation \( v = H \cdot d \), where \( v \) is the galaxy’s outward velocity, \( d \) is the galaxy’s distance from Earth, and \( H \) is the constant of proportionality known as Hubble’s Constant. The common unit of measurement to measure the velocity of a galaxy is km/sec, while the most common unit of measurement to measure the distance to a galaxy is called a Megaparsec (Mpc), which is equal to 3.26 million light years or 30,800,000,000,000,000,000 km. Therefore, the units of measurement for Hubble’s Constant are \((\text{km/sec})/(\text{Mpc})\). The exact value for Hubble’s constant is still up for debate, but the value is generally believed to be around 65 km/sec/Mpc. This means that a galaxy that is one Megaparsec away from the Earth will be moving at a velocity of about 71 km/sec. This means that to determine an object’s distance from the Earth, only the velocity need be known, which can be determined thanks to the Doppler shift (Redshift and Hubble’s Law). The velocity yielded from observing Doppler shifts can then be input into the Hubble equation and used to find the distance of the object from the Earth.

The velocities of galaxies is determined largely in part thanks to the use of red shifts and the Doppler effect. The light coming from distant stars and even more distant galaxies is not featureless, but in fact has distinct spectral features characteristic of the atoms in the gases around the stars. When the spectra of these stars are examined, they are found to be shifted towards the red end of the color spectrum, indicating that almost all of them are moving away from the Earth (Nave). The wave nature of light means that there will be a shift in the spectral lines of an object if it is moving. Thus, by using the Doppler shift of objects, their velocity can be calculated and then used to find the distance of the star or galaxy from Earth. An object’s motion causes a shift in the wavelengths of the object. This shift depends on the speed and direction the object is moving, so this shift can be used to estimate the velocity of the object. The “radar guns” used by police officers operate under this same principle as well. The gun sends out a radio wave at a specific wavelength/frequency that reflects off of the passing car back to the radar gun. The device is able to determine the car’s speed from the difference in the wavelength/frequency of the transmitted beam and the reflected beam.

Hubble was able to determine that the universe is expanding through much time and an inconceivable amount of pictures. In his day, the most pressing question of the time concerned the nature of cloudy patches called nebulae. Most of Hubble’s colleagues believed that these nebulae were contained within our own Milky Way, but Hubble succeeded in disproving this by taking many and more pictures of the nebulae, showing that at least some were outside the Milky Way. After Hubble arrived at Mt Wilson observatory, he began taking photographs of the same group of nebulae, now known as galaxies. He needed to take many pictures of the same galaxies over time so that any visible change could be observed. On October 4, 1923, while comparing a photo of the Andromeda galaxy to one he had taken the night before, Hubble identified a Cepheid variable star, a type of star that could provide the means for determining the distance of a galaxy. By comparing its apparent brightness with its actual brightness, Hubble was able to determine that the star was 900,000 light years away, later proved to actually be closer to 2 million light years (1929: Edwin Hubble).
Annotated Bibliography

This site gave a great amount of information on Hubble and his history of studies. It gave a specific galaxy which he photographed and explained his process of how he was able to reach his conclusions. This website gave solid information in easy to understand terms and went more in depth into the history of Hubble, rather than the science of what he did.

This website is obviously more directed at children, which turned out to be a good thing actually. This site was very easy to understand and aided in furthering my understanding of Hubble’s Law and how it came to be.

This website is a Harvard sight that is strongly worded and full of a lot of information. It describes the history of Hubble’s Constant and how it came to be. It describes in detail the timeline of the Hubble Constant and its contributions toward astronomy.

This website was a little shorter than most of the others that I visited. However, it contained useful information on the actual value of Hubble's Constant that I used in my essay.

It describes in fairly simple terms what Hubble’s Law is. It gives good examples that are easy to understand. This is a .edu website that is focused on teaching and making the Hubble Constant easy to understand by putting it in layman's terms.
Ptolemaic Astronomy

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Introduction to Solar System Astronomy

Spring 2015

Professor J. Weitz
**Claudius Ptolemy created a model of the solar system was an alteration of previous models that were previously developed by Greek astronomers. Ptolemy’s major contribution was that of the “Heavenly Sphere”. He concluded that planets orbited the Earth, which was not only the center of the solar system but the center of the entire Universe. The idea itself is not new to astronomers of this time period, however the patterns of orbit were revolutionary and accepted for centuries.**

In today’s society, we have all been told that the Earth revolves around the Sun. One may also probably know that planets other than our own have their own moons, and the way to test to see whether or not something is true is by experimenting. Today we are fortunate enough to have the resources to research any questions we may have about the great unknown. Thousands of years ago, these facts were not widely known. Astronomy was anyone's guess, and the way things were was just the way the gods had made them. It was felt there was no need to truly understand them or put them in any kind of order. Early scholars changed much of that. More specifically, a man by the name of Claudius Ptolemaeus changed the way many believed the heavens to be designed. (Johnson, V)

Ptolemy lived in Alexandria, Egypt around 100 AD. He worked as a Mathematician, Astronomer, and Geographer. Although not much can be said about his personal life, his hard work and legacy speaks for itself. For years, Ptolemy worked to order the Greek geocentric view of the universe. The Greek word for Earth is “geo”, and centric means centered, hence the term geocentric model. Not only was the Earth believed to be the center of our own solar system, but of the entire universe as well. The predictions made by early astronomers such as Ptolemy, seemed to be the first steps to actually understand many of the mysteries of nature. He was also able to rationalize the apparent motions of the planets as they were known in his time. (Ptolemy and Astronomy)

Unfortunately, the Ptolemaic model was extremely complicated to explain. Therefore, not many people actually took the time to learn and practice it themselves. Retrograde motion uses a system of circles on circles to explain the orbits of the planets called epicycles and deferents. The main orbit of the planets is the deferent, the smaller orbit is the epicycle. In the figure below, only one epicycle is actually drawn out; however, over 28 were required to explain the actual orbits of the planets. (The Ptolemaic Model, 2001)
His influence is strictly due to the accident that his predecessors’ works were lost while his survived. Many of their legacies and achievements are known only through him, and when he disagrees with them it is usually Ptolemy who is wrong. One astronomer whom he has a lot of corresponding work with is Hipparchus.

Hipparchus was a Greek astronomer whom Ptolemy had been inspired by. In fact, he spent most of his time elaborating on Hipparchus’ research. To this day much of what we know about Hipparchus, we know because of evidence of Ptolemy’s research. Ptolemy’s model of the solar system and heavenly sphere was a refinement of previous models developed by Greek astronomers (much like Hipparchus). Ptolemy’s major contribution, however, was that his model could so accurately explain the motions of heavenly bodies (planets), it became the model for understanding the structure of the solar system. (HISTORY OF ASTRONOMY)

Mathematically, everything checked out, due to the fact that Ptolemy was able to make it do so. Because of this, no one thought to question the integrity of the Ptolemaic system. While studying the orbits of the planets of our solar system, Ptolemy was able to expand on Hipparchus’ catalog of constellations. By doing so, he found an extra 270 stars. It is said that he studied in an observatory north of Alexandria, for over 40 years working on his observations and discoveries. Experts debate as to whether or not he truly viewed and recorded each star, or if he used various methods, including extrapolation in order to complete his catalog. (HISTORY OF ASTRONOMY)

One of Ptolemy’s greatest legacies is that his astronomical work was divided into thirteen books. His book was known as Ho megiste astronomas (Greek for 'the greatest astronomer’), or Megiste for short. Once it reached northern Europe through the Arab civilization in Spain, it
acquires its eventual title - as “Ptolemy's Almagest”. He decided in order to preserve his life’s work, a series of books that would tell of all he accomplished. Each book was a new chapter and practically an encyclopedia of mathematics to explain the phenomena of astronomy. It contains numerous explanations ranging from earth conceptions to sun, moon, and star movement. It also covered predictions for when eclipses would occur and a breakdown on the length of months. Astronomers across the globe in the middle ages found these to be extremely helpful in developing new hypotheses. Below is a list of the chapters and their contents. (Ptolemy and Astronomy)

“The 13 chapters of the Almagest according to a lecture:

1-2: Motion of the celestial sphere, table of chords
3-4: Length of the year and month, solar theory, theory of the motion of the Moon
5: Construction of the astro-lab
6: Theory of eclipses
7-8: Precession, Catalog of 1022 stars and nebular, Theory of Planetary motion”

(Ptolemy and Astronomy)

In the late 820s Ptolemy's Almagest was translated into Arabic. The Ptolemaic explanation of the motions of the planets remained the accepted wisdom until the Polish scholar Copernicus proposed a heliocentric view in 1543. It should be pointed out as well that Ptolemy's system is actually more accurate than Copernicus's. The heliocentric formulation does not improve on Ptolemy's until Kepler's Laws are also added. It took 1200 years for astronomers to come up with a new system. (HISTORY OF ASTRONOMY)

In practical terms the Ptolemaic system proves adequate for everyday purposes. Although, its complex nature makes it attractive to the exclusive minority of learned men. The details may be hard to master, but once understood they will reveal future positions of the planets. Ptolemy himself preparing charts of the moon's behavior, more accurate than any previously available, which remain in everyday use until the Renaissance. But in the long run the complexity is unconvincing (the alternative proposed by Copernicus is simpler); and the orbiting planets of Jupiter, revealed by Galileo's telescope, inconsiderately smash through one of Ptolemy's crystal spheres. (PTOLEMAIC ASTRONOMY)

For any human being, a 1200 year legacy is unbelievable. It is said that a person dies twice, once when they are buried, and again when their name is mentioned for the last time.
Claudius Ptolemaeus will not be forgotten any time soon. Although today we know that the earth orbits the sun and Ptolemy’s model could not at all be true, his influence in human history will live on for centuries to come. That in of itself is incredible.
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Analysis of Carbon Nanotubes, Polymers, and Applications

Cody Boeckholt
Abstract

The paper will be discussing the strengths, applications, construction methods of carbon nanotubes and their composite forms. Construction methods will be fully analyzed from a theoretical standpoint as not every construction method will exist. Applications will vary from physical to electrical highlighting the potential uses of carbon nanotubes in both categories.

Introduction

A carbon nanotube is a honeycomb lattice cylinder with a diameter size that is relative to a nanometer and a length that is more than a micrometer.\(^1\) Carbon nanotubes have a unique electronic property due to their free moving electrons and its ability to be a semiconductor is based on the way a carbon nanotube is made.\(^2\) Polymers of carbon nanotubes discovered by Alan Heeger, Alan G. MacDiarmid and Hideki Shirakawa to be conductive of electricity in the 1970s started the trend of some of the first conductive non-metallic materials.\(^2\) Conductive carbon polymers have single and double bonds alternating backbone called “conjugated double bonds” that allow it to conduct electricity because the electrons are free to move around.\(^2\) Carbon polymer-based composites however in their conductive polymer forms still require doping.\(^2\)

Carbon nanotubes come in two main types a single-walled(SWCNT) and multi-walled(MWCNT) with diameters that range.\(^1\)

Preparation/Manufacturing of Carbon Nanotubes for composites:

Techniques used for the manufacturing of carbon nanotubes are varied: arc discharge, laser ablation and chemical vapour depositions.\(^3\) Which all yield a mixed result of both single and multi-walled carbon nanotubes along with carbon nanoparticles.\(^3\) The disadvantages of using the three methods above is that they affect several of the properties of carbon nanotubes.\(^3,2\) For all three methods above, a purification method must be present in order to remove impurities and prevent the carbon nanotubes from falling apart.\(^3\)

Preprocessing:

Each one of the steps below discusses the in-depth methods used to prepare carbon nanotubes for insertion or mixing with other materials in order to create a polymer.

Purification:

The purification technique typically relies heavily on one of the properties of carbon nanotubes which is its’ strong resistance to oxidation thus several of the methods involve heavy oxidations and catalyst. The most common purification methods are gas-phase oxidation, intercalation methods, liquid-phase, oxidation and physical separation.\(^5\) Intercalation is the reversible inclusion or insertion of a molecule. Typical intercalation methods involve a metallic ion and an atom of an element that can easily oxidize away leaving just the metallic ion behind.\(^3\) For example, copper chloride is typically used in intercalation because it not only adds a metallic ion, but the chloride oxidizes away taking the nanoparticles away with it.\(^3\) Disadvantage to using intercalation and oxidation is that the final material can possibility be contaminated with residues from the intercalates the material inserted into the carbon nanotube.\(^3\) Physical separation methods of a carbon nanotube first involve a surfactant, after the solution is treated by sonication, filtration, centrifugation or chromatographic methods.\(^3\) Long periods of sonication can lead to damage of the carbon nanotubes splitting them into smaller tubes and some forms of oxidation through acids can damage carbon nanotubes as well.\(^3\) A recent method developed by
Julie L. Zimmerman and co-workers at Yale’s School of engineering takes raw carbon nanotube felts and first prepares them through a two-step technique involving oxidizing-acid reflux. Compared to the laser vaporization process which removes impurities through chemical dissolution, it leaves the carbon nanotubes with less impurities. Disadvantages to using physical or chemical purification techniques is that it involves steps that alter the properties of the carbon nanotubes which can weaken the tubes. In some cases physical separation techniques that involve chlorine gas can leave a hydroxyl group on the cap or top of the carbon nanotubes. The hydroxyl group can then deprotonate causing the carbon nanotubes to fall apart. A fix for the physical separation technique is to add an extra step involving hydrochloric acid and other strong acids which drive any reaction that has the potential to deprotonate in the reverse.

### Graphitization:
Graphitization is a method which treats some of these defects throughout the carbon nanotube through super heating of them at temperatures ranging from a 1000°C (Celsius) to 3000°C. Studies and experiments in multi-walled carbon nanotubes generated through chemical vapour deposition which is the process that involves disposing a thin solid layer of a substance on a surface as a result of vapor-phase chemical reactions in high temperature gas that are close to the surface. Proved to be an efficient method in fixing microstructural defects at 1800°C and even removed residual metals from the tubes. The disadvantage of graphitization is that it does not fix kinked or side grafts in the tubes.

### Covalent Functionalization:
Covalent functionalization is the bond between carbon nanotubes and is a process with which carbon nanotubes solubility is increased consequently it also increases its’ resistance to the van der waals forces. It also increases the bonding between carbon nanotubes and polymers through covalent bonds. It does this through covalent bonds which typically debundles the carbon nanotubes and allows the chemical species to intercalate between the bundles increasing the solubility. Two categories of covalent treatments have been identified “grafting to” and “grafting from”. “Grafting to” involves a reactive groups or radical precursors that have a molecular weight that terminates with the polymer’s. The idea of polymerization is that we are bonding another atom of a another material to the carbon nanotubes with a reaction between the two that will generate a chain of monomers. The most common process starts with a free radical which is a molecule that contains a free electron that is inserted near a monomer (a molecule of lower molecular weight) in this case the carbon in the carbon nanotube which reacts by stealing an electron from the monomer creating another free radical. The process continues until a chain of it is joined together or a free radical steals an electron from a monomer attached to another carbon molecule. One of the advantages to the “grafting to” method is the amount of polymer being bonded to carbon nanotubes can be limited. The “Grafting from” method involves in situ polymerization of the monomers in the carbon nanotube in order to grow polymers from the walls of the carbon nanotubes. One advantage is that the polymers are denser than the “grafting to” method, but at the cost of uncontrolled growth of the polymers.

### Non-covalent Functionalization:
One problem commonly associated with carbon nanotubes is the fact that they are formed in clumps of tubes that need to be dispersed and spread out in order to use them. The non-covalent bonding helps in the dispersion of these tubes because it increases the solubility of them through the prevention of allowing the carbon nanotubes to share electrons. The method involves warping the carbon nanotube in a polymer and surfactant that decreases the surface tension in the carbon nanotube. The main purpose of the surfactant is to prevent the formation of aggregates (masses of carbon) and help the tubes overcome repulsion from the van der Waals forces.

**Solution Processing:**
Solution processing is the most common method for fabrication of carbon nanotube polymers composites. Common process methods involve a carbon nanotube powder being dispersed in a liquid that is vigorously stirred or put under sonication. From this solution a carbon nanotube film is developed from this mixture or mixtures similar to it in nature. Other methods involve the mixture of single walled carbon nanotubes that has been purified and then mixed with metallic ions making it semi conducting. Advantages of using solution processing is that it can produce a variety of carbon nanotube based films that can be laid on top of other materials for reinforcement.

**Synthesis of Conducting Polymers:**
A method in which carbon nanotubes are developed into an electrically conducting polymer is through synthesis. Where carbon nanotubes are put into a strong acid that is consistently stirred throughout the reaction as the carbon nanotubes react with the excessive amount of oxidant in the solvent to develop a chain of monomers. The other method of synthesizing a conducting polymer is through electro-chemistry where carbon nanotubes are placed in a pool of counter and reference electrodes. The diluted monomers and electrolyte which is the dopant then has a voltage run through it that develops a film of polymer over the top of the liquid. This film is electrically conductive and has the potential to be used in small circuit board applications.

**Melt-mixing:**
Melt mixing is a process in which carbon nanotubes are combined with polymers typically found in plastic. Polymers such as polypropylene (PP) are heated past their melting point and combined with carbon nanotubes through typical manufacturing methods used in the process of manufacturing plastics. The process involves extrusion, internal mixing, injection molding, compression injection molding and blow molding. The advantage to using melt-mixing is that no solvents are need and the carbon nanotubes do not have to be soluble in order to be used. The disadvantage to using melt-mixing is that the dispersion of carbon nanotubes throughout the composite is less than methods involving polymerization.

**Applications and purpose of carbon nanotubes:**
Applications of carbon nanotubes vary as they can be used as capacitors, sensors, high tension wire, reinforcement of other materials, connections for other electrical components and ultra chemical sensors. Each category of application for carbon nanotubes will be analyzed along with its theoretical strength and manufacturing process if possible.

**Semiconductors/Electrical devices:**
Carbon nanotubes possess several key factors that make them ideal for electrical devices if the hurdle of doped carbon nanotube polymers being too sensitivity to ambient conditions can be overcome. As computers advance the difficulty of developing smaller transistors, capacitors and connections pathways becomes evident. Carbon nanotube polymers show potential to be part of every single component on a motherboard and possess $\text{mho(Siemens(S))}$ that can be as large $80,000 \text{ S cm}^{-1}$ or as small as $10^{10} \text{ S cm}^{-1}$. Siemens is the inverse or reciprocal of resistance in the case of direct current which most electrical devices are powered under. Referring to the chart below: Conducting polymers expand across all three categories commonly associated with electrical devices such as computers, cellphones, and even TVs. Several of these ideal properties of carbon nanotubes and their composites are decided by several factors. First is how the carbon nanotubes are formed or rolled. Second, is the amount of doping a carbon nanotube and its composite forms receives in order to increases it conductivity. Doping which adds impurities to the material either adds (reduction) or removes (oxidation) electrons from the carbon polymer-based composites this method of doping is specific to polymers themselves. Doping unfortunately makes the conductive polymer carbon composites highly unstable, and highly susceptible to ambient temperatures. Carbon nanotube polymer composites have been observed even at low levels of doping to have a significant increase in the level of conductivity. The four main polymer families at the moment for conductive carbon polymer families are: polyaniline (PANI), polypyrrole (PPy), polythiophene (PT) and poly (phenylene vinylene) (PPV).

One potential application for carbon nanotubes is in super capacitors because of their high mechanical strength, good electrical conductivity and relatively low cost. The two main types of super capacitors that can be developed from carbon nanotubes are double layered, pseudocapacitor, and a combination of double layered pseudocapacitor with regular capacitors. Carbon nanotube composites combined with metal oxides such as RuO2,54 NiO55 and MnO56 have been observed showing promising results of increased charge efficiency. Carbon nanotubes and conductive polymers show more potential than combining carbon nanotube composites with metal oxides. A conductive polymer is an organic polymer that approaches a conductivity closer to metal. Combining carbon nanotube composites and conductive polymers have been observed to increase pseudocapacitance, the amount of electrical energy stored by reversible faradiac redox reaction. Creating one of these capacitors involves taking a solution of monomers of carbon nanotube and conductive polymers nanocomposites depositing them onto a carbon nanotube. In 1999 a carbon nanotube electrodes were deposited onto a film of polyaniline...
Deposition is the direct solidification of a material onto the surface. This resulted in increased current charge and better polymerization. The best result however came from depositing polypyrrole (PPy) on to aligned carbon nanotube arrays and then treating it with sulfuric acid. This resulted in an increased electrode of 180 F(farad) and a 20% loss of charge after 2000 galvanostatic charging-discharging cycles.

Another application carbon nanotubes are being looked at for is rechargeable batteries. Currently mobile technology is continuing forward at pace so fast that current batteries like the lithium ion/polymer based ones cannot keep up. Not only does it take a significant sized battery to power most smart phones today, the consumer market is consistently complaining for a thinner phone leaving manufactures like Motorola looking for a viable option like carbon nanotubes to improve lithium based batteries. At the moment carbon nanotube composites are being looked at for replacing the traditional electrodes in lithium-ion/polymer batteries in order to increase cycle life of the batteries. Research into multi-walled carbon nanotubes polymerized with poly(Nvinylcarbazole) used as positive electrodes with lithium produced a 45 and 115 mA h\(^{-1}\) g\(^{-1}\) discharge capacities and a 1.5 to 2.5 Volts during 20 discharge cycles. Another viable nanotube composite that is ideal for an electrode with lithium is aligned carbon nanotubes polymerized with Poly(3,4-ethylenedioxythiophene)(PEDOT) and Polyvinylidene fluoride(PVDF) which yields a 265 mA h\(^{-1}\) g\(^{-1}\) discharge after 50 cycles. The benefit of using a polymer like this is its’ flexibility, high conductivity and easy integrates into lithium batteries because it does not require any metal substrate. A metal substrate is what the polymer composite is typically coated on in a lithium battery.

**Physical applications of carbon nanotubes composites:**

Physical application for carbon nanotubes vary from reinforcing current polymers to cables that could potential hold up a future space elevator. Currently carbon nanotubes are not used in many physical applications due to its high cost. It is typically found in military jets and expensive aerospace satellites at the moment in variety of different materials. The main materials are typically a form of carbon nanotube yarn, tape, braid conductor, conductor braiding and twisted cables. The primary reason why carbon nanotube composites are found in space applications is because of its lightweight and high thermal threshold. Typically satellites use carbon nanotube materials as the high cost is cheaper compared to the amount of fuel burned to launch the satellite into space at a cost of 20,000 dollars per 10 kg of mass. Carbon nanotubes are also prized for their conductivity used as a tape, braid and cover for other capacitors as it guides charges away from sensitive electronics while also protecting them from extreme temperatures. Just recently has carbon nanotubes been developed into a cable because getting all the tubes to align in the same direction, disperse in a solvent, and be soluble is very difficult as discussed in
previously in the paper. Rice University has just developed a method that produces a cable of carbon nanotubes that is flexible and is four times more conductive than copper. It is apparent very resistive to sheering forces and is as thin as a human hair. Carbon nanotubes are not just limited to cable application and temperature ones, but also strength application. As an unmodified nanotube of carbon is very ridged due to its’ double conjugated bonds along the backside of the structure. This allows carbon nanotube composites to have an endless list of physical applications as long as a better method of dispersing, aligning and purifying the tubes can be found until then the melt-mixing process is the only form of polymerization that allows for physical applications for carbon nanotubes without severe modification of their ridge structure.

Conclusion

While carbon nanotubes possess the ability to be used in nearly every type of polymer in both structural and conductive materials. It severely lacks cost effective larger-scale production methods associated with current conductive polymers and composites. As it is very hard to disperse in solvent through electrochemistry and is damaged in its doped form by ambient temperatures which cause the tubes to fall part. It is hard to purify of defects, align, grow carbon nanotubes with the same structural orientation, straighten and untangle clusters of them. If new and more efficient methods can be developed that not only take advantage of carbon nanotubes wide variety of applications, but efficiently address these sever problems associated with carbon nanotubes and their composites then carbon nanotubes have a chance to become an essential part of our everyday life.
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Major Planets vs. Dwarf Planets

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Introduction to Solar System

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The International Astronomical Union (IAU) demoted the ninth planet Pluto to a dwarf planet on September 13, 2006. The IAU did this when at an annual meeting a majority voted for the definition of a planet to be changed. Pluto did not meet all the requirements for the new definition and was then classified as a dwarf planet. As a result from the demotion of Pluto some of the world showed great emotion, mostly elementary school children. When Pluto is compared to dwarf planets and planets, it is clear that Pluto should have not been a planet and is now labeled correctly as a dwarf planet.

The International Astronomical Union (IAU) has been in charge of the naming and nomenclature of planetary bodies and their satellites since 1919 when the organization was founded. This means the IAU determines what the planets are of our solar system and the definition of a planet. The first draft proposal of the definition of a planet was first debated at the 2006 general assembly. After the assembly the definition of a planet was determined. The definition of a planet consists of three rules. One, it must orbit a star. Two, it must have enough mass to assume a nearly round shape, and three it must clear the neighborhood around its orbit (International Astronomical Union). After this definition was derived the IAU had to reevaluate the solar system to see if any object disobeyed these rules. After looking all planets except Pluto followed these rules. Pluto disobeys a vital rule necessary to be considered a planet, which is why it was demoted to a dwarf planet. If Pluto's size, composition and orbital path are compared to the other planets of our solar system then it is clear to see that made the right decision.

The IAU has an infamous reputation ever since they demoted Pluto however; organizations like the IAU serve a major purpose in astronomy. In 1781 Sir William Hershel was observing what he thought was a comet through his telescope the only reason he thought this because the object he was observing moved very slowly across the sky. But it after more observations he came to the conclusion that it could possibly be a new planet. After other astronomers around the world confirmed that the object was a planet they placed the responsibility of naming the newly discovered planet on Sir Hershel. Sir Hershel, like any good research scientist, decided to acknowledge his funding source, which was King George III at the time. So in honor of King George Sir William Hershel named the planet Georgium Sidus, which translates to George’s Planet (Uranus). A rather comical thought is that in 1781 the planets of the solar system were Mercury, Venus, Earth, Mars, Jupiter, Saturn and George. The name George did not last long and the planet was ultimately named Uranus, which is the Greek god of the sky. So organizations like the IAU serve a very responsible purpose in the science of astronomy.

Before one can compare planets one must learn how planets form around their host star. This is a hot topic for debate in astronomy because it is still not exactly clear on how planets form, but the most widely accepted theory is the Solar Nebula Theory. The Solar Nebula Theory states that stars form in massive clouds of hydrogen also called giant molecular clouds (GMC) they do so by molecules of hydrogen clumping and becoming denser. As the cloud condenses more and more the rotation or angular velocity continues to increase as the clump or ball of hydrogen condenses into what is called a protostar this is true because of the law of conservation of angular momentum. The protostar is the star in its infancy stage and over the course of about 100 million years planets will have the opportunity to form. If a person were to look over our solar system they would notice that there are eight planets in all (yes only eight) and that the inner most four are terrestrial planets and the outer most four are gas planets or Jovian planets.
The word jovian means Jupiter like). Is this a coincidence? Not according so the Solar Nebula Theory. Terrestrial planets are closer to the sun because at the temperature zone in which they formed any gas in at that temperature would have had too much kinetic energy to condense, leaving only rock and metal to condense. And the moment that the hydrogen gas can condense, it does. This is why planets are ordered from terrestrial to jovian and not jovian to terrestrial or even random.

Now looking over the planets of our solar system the first stop is a planet that is only a little larger than our moon Mercury. Mercury experiences extreme temperature fluctuation because it has a very thin atmosphere so thin it is almost undetectable. Temperatures can reach 430 degrees Celsius and drop to -180 degrees Celsius. Mercury also rotates very slowly in comparison to its orbit period. Mercury rotates once every fifty-nine Earth days and orbits the Sun once every 88 Earth days. By the time Mercury has completed one rotation it will already have completed 67 percent of its orbit around the Sun. As of today there have only been two spacecraft that have been sent to Mercury. Mariner 10 in 1974-75 and MESSENGER in 2011, which is still in orbit around Mercury (Mercury: Overview).

58 million kilometers away from Mercury we run into Venus, the second planet. Venus has an equatorial radius of 6,051.8 kilometers, which is ninety-five percent of Earths radius showing that the two planets are nearly identical in size, however, are completely different in weather. Venus’s average temperature is 480 degrees Celsius or 900 degrees Fahrenheit; this is because of the abundance of greenhouse gases on Venus such as carbon dioxide. Venus’s atmospheric pressure is ninety times greater than that on Earth, which makes astrobiologists skeptical that life could survive in these extreme conditions. Another extreme notation of Venus is its rotation motion, rotation time and orbit time. If a person were living on Venus and decided to watch a sunrise and occur they would have to look to the west instead of the east. Why is this? This occurs because Venus rotates clockwise instead of counterclockwise like the majority of the planets in the solar system. Thus causing the Sun to rise in the west and set in the east on Venus. And if this person wanted to watch the sunset on Venus they would have to wait 243 Earth days but also know that Venus’s orbital period around the Sun is 225 Earth days. This means that a day on Venus is longer than a year (Venus: Overview).

After Venus comes Earth, our home and since no one will be reading this that isn’t from Earth it seems like a waste of energy to inform the reader of the obvious. But a fact that is not widely known is that Earth days are getting longer in time and the Moon is getting farther and farther away each year. The Moon is stealing angular momentum from the rotation of Earth causing Earth to spin slower and ultimately making the days longer but this stolen energy from the Moon is also what is causing the Moon to drift away.

The next stop is the “red planet” or better known as Mars. Mars surface is covered with iron rich minerals that have oxidized over time giving Mars a reddish color. Mars is named after the Roman god of war and Mars’s moons Phobos and Demios are from Greek mythology meaning panic and fear. This is because it is said that panic and fear follow war. Over 40 spacecraft have been sent to Mars from orbiters to flybys this is because Mars shows evidence that it once was a place where life could thrive. There are dried up riverbeds and deltas on the surface of Mars. There is also evidence that there was once an ocean due to the erosion of rocks on the surface of Mars. Since the only life that humans know of exists on Earth and the life that we know of liquid water is an
essential ingredient in the mixture of the recipe of life. If there ever was life in the oceans that flowed millions of years ago there’s the chance that there could be a fossil of a Martian organism. Thus this is the reason why a mission to Mars could be beneficial (Mars: Overview).

In 2026 a privately funded mission called Mars One will launch. The mission’s goal is to send four people with a one-way ticket to Mars and send four more people every two years until a small village is living on Mars (Mission - Mars One). People today have such deep aspirations to explore Mars that they are willing to conduct a life threatening experiment to explore Mars. One would think that there would be a hard time finding volunteers for this mission however though, there are numerous people that have volunteered and are hoping to be shipped off to most likely their death. The reason for this is because resource will eventually run out and just because there is plan to send more doesn’t it will fall through.

After one were to journey past the four terrestrial planets they would pass through the asteroid belt that lies between the last terrestrial planet and the first jovian planet or more specifically between Mars and Jupiter. A common misconception is that the asteroid belt is a dense, rock area teeming with activity, but in reality the asteroid belt does not have that much material. According to one source “If you could stand on an asteroid and look around, the next one would be too far away to see very well” (Asteroid Belt). If every asteroid in the asteroid belt were accounted for and all their masses were added up; the total mass of the asteroid belt would only be four percent of our Moon’s mass. There is one body in the asteroid belt that accounts for nearly a third of the total mass of the rocky belt and that is the dwarf planet Ceres. On March 6, 2015 NASA’s spacecraft DAWN entered orbit around Ceres and will study the dwarf planet for fifteen months. Researchers hope to gain more insight on how complex planets like our own form over time and the formation of star systems.

The first of the jovian planets is Jupiter the most massive planet in our solar system. Jupiter is three hundred seventeen times as massive as the Earth and has a circumference eleven times larger than Earth’s. Even with all this mass Jupiter completes one revolution every 9.92496 hours compared to Earth’s twenty-four. This fast rotation gives Jupiter extreme weather; an example is The Great Red Spot on Jupiter, which is a storm that has been observed ever since the 1600’s when Galileo first turned his telescope to the sky. Jupiter’s massive body had more influence on objects in the early solar system allowing it to collect more moons with its gravity. So far Jupiter has sixty-three confirmed moons and others that are up for debate. Jupiter’s four largest moons are some of the most interesting bodies in the solar system. Europa, one of the four, has a sub surface ocean that is quiet larger the total amount of water on Earth. This has scientist very interested in visiting Europa for the potential of finding life (Jupiter).

After Jupiter comes Saturn and its glorious rings. Saturn is the easiest planet to recognize because it has a ring system like no other planet in our solar system. All the jovian planets have rings but Saturn’s are the largest and most noticeable because they reflect the most light from the Sun back. Saturn’s rings branch out 282,000 kilometers from the planet (about ¾ the distance between Earth and the Moon) while the rings are extremely long they are only a kilometer thick. Saturn’s density is around .97 g/cm^3 and the density of water is 1 g/cm^3 this means if Saturn was put into an ocean large enough it would float since it is less dense than water (Saturn).
Up next is the odd ball of the planets Uranus. Uranus is an oddball because it is tilted on its axis at almost a ninety-degree angle. This tilt causes violent weather on Uranus since one of the poles of Uranus constantly faces the Sun. Another odd fact about Uranus is that all of its moons are named after characters in Shakespeare’s plays. And Triton, one of Uranus’s moons, orbits Uranus the opposite direction from which Uranus rotates and the laws of physics tell us that eventually that Triton and Uranus will collide and it should display a spectacular show in the night sky (Uranus).

The last planet of our solar system is Neptune. Neptune is also a place with extreme weather with windstorms reaching speeds of 600 miles/hour. Neptune is offend referred to as Uranus sister planet because they share a very similar composition. Since Neptune’s discovery it has only completed one orbit around the Sun the long journey of 165 Earth years (Neptune).

Finally we have reached Pluto, the famous dwarf planet. In 2006 Pluto was demoted to a dwarf planet and this shocked the world. For some reason Pluto had a special place in peoples hearts and felt that it was unfair that it was demoted but if one is presented with the facts it is clear that Pluto is much happier with its family, the dwarf planets. There are three rules that a body in space must obey to be classified as a planet and they are: 1. Must orbit the Sun or star
2. The body must have enough mass to assume a nearly round shape
3. Must clear the neighborhood around its orbit.

Now if we look at Pluto it has a round shape, it orbits the Sun but it does not clear its neighborhood around its orbit of the Sun. Pluto actually crosses the path of Neptune so at sometimes in Pluto’s orbit around the Sun it is not the farthest body in the solar system. No other planet in the solar system behaves this way.

If you look at the figure then you can see how Pluto not only crosses the path of Neptune but also Pluto does not orbit in the same plane as the rest of the Planets. Going
back to the Solar Nebula Theory it states that everything formed in a rotating disk that means all the planets formed in the same plane and since Pluto is not in the same plane, we can deduce that it is not a planet but a dwarf planet.
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The Physics of Artificial Joints for Orthopedic Applications

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Abstract:
Total knee arthroplasty is currently the most common orthopaedic surgical procedure in the United States. Due to the complexity of the anatomy and physiology of the knee joint, it has taken over fifty years for physicists, chemists, engineers, biomedical and clinical scientists to perfect the design and implementation of this prosthetic implant, its surgical procedure, and its longevity as a weight bearing joint replacement. The purpose of this paper is to take a deeper look at the anatomy and physics of the knee, and the revolutionary scientific advancements of the total knee prosthesis and its future.

The knee is the largest and most complicated joint in the human body. The asymmetrical joint is made up of four bones, four ligaments, and two tendons that articulate with eighteen different skeletal muscles. As a result of such complicated anatomy, the knee is a compilation of three separate joints combined into a single joint cavity: the femoropatellar joint and the medial and lateral tibiofemoral joints. The femur (thighbone) connects with the patella (kneecap), then articulates with the tibia and fibula (lower leg and shin bone, respectively). These bones are secured by a series of assisting structures that hold the joint in place and facilitate movement.

While the knee joint articulates with eighteen different muscles in the lower extremity, the primary muscles involved for controlling the knee are the quadriceps (rectus femoris, vastus intermedius, lateralis, and medialis) and the hamstrings (biceps femoris, semitendinosus, and semimembranosus). All muscles acting around the knee joint bridge two different articular joints: hip to knee, and knee to ankle (Scuderi, 2010). These muscles are attached to the skeletal bone via the tendons, then stabilized and strengthened by four main ligaments (Merieb, 2013). The anterior cruciate ligament (ACL) is critically important as it prevents the tibia from being hyperextended relative to the position of the femur. This can easily be torn during the twisting or bending of the knee. The posterior cruciate ligament (PCL) prevents backward displacement. The medial collateral ligament (MCL) and lateral collateral ligament (LCL) prevent each side of the knee from being bent open due to an applied stress on the opposing side of the knee. The role of the ligaments is act as proprioceptive stress transducers, both allowing and limiting motion (Scuderi). The knee joint itself is only partially enclosed by a capsule, and has alignment in three planes: frontal, sagittal, and horizontal (Merieb, 2013).

The knee is an example of a synovial joint. [See Figure 1.] Synovial (“joint eggs”) joints are comprised of articulating bones that are separated by a fluid filled cavity. The ends of each bone in the joint are surrounded by articular cartilage, separated by an articular cavity. This cavity is enclosed by an articular joint capsule that is composed of two layers: an external fibrous layer, and a synovial membrane composed of loose connective tissue. This capsule is also filled with a slippery, synovial fluid with a viscous consistency. The sole purpose of the synovial fluid is to reduce friction between the cartilages; without it, the joint surfaces would wear away quickly. The knee also has a built-in protective mechanism called the meniscus. The meniscus is a crescent shaped, cartilaginous structure that separates the two articular surfaces. The meniscus serves many functions, including improving joint stability, absorption of shock to the knee joint with movement, protection against soft tissue damage, and distribution of synovial fluid. As the meniscus is attached only at the outer edges, it can easily be torn free. Synovial joints are ideal features among the skeletal system, as they allow for substantial freedom of movement (Merieb, 2014).

Movement of the knee is facilitated mostly by the impulses sent to the adjacent skeletal muscle to move the joint. The femoropatellar joint is a plane-joint, as the “floating” patella
glides across the femur during flexion. The medial and lateral tibiofemoral joints are hinge joints, providing both leg flexion and extension. The quadriceps provide for extension at the knee, while the hamstrings and sartorius provide flexion. These work together with the gastrocnemius, plantaris, popliteus, and gracilis to achieve this effect. The three adductor muscles work with the gracilis for leg adduction, and the tensor fasciae latae (connecting to the iliotibial band) handles abduction. (Merieb, 2013). While these muscles all work cohesively together to facilitate movement and support the knee joint, the quadriceps and hamstring muscles are the two primary muscles responsible for knee movement and locomotion, working antagonistically with each other. Contraction of the quadriceps muscles result in a straight leg and knee; contractions of the hamstrings result in a knee bend (Byrant et al. 2011).

Physics is an underlying theme in the design and function of the knee-joint and its ability to provide structural support and locomotion for the human body, as well as absorb the impact of the forces involved in daily activities. The knee is a prime example of conservation of energy, or the concept that energy is not something that can be created nor destroyed, but rather transferred from one form to another (Serway, 2012). Conservation of energy in the body converts chemical energy into mechanical energy. It begins when skeletal muscles are fueled with energy provided from nutrients in the human diet, creating a form of chemical energy. This chemical energy is then converted into both elastic and gravitational potential energy.

Elastic, or spring potential energy, can also be considered stored energy arising from the amount of work needed to stretch or compress a spring (Serway, 2012). As spring forces are conservative, the work done can be recovered by removing the applied force. In the human body, muscles can be considered the spring, as the tightening of the quadriceps to lift the leg is comparable to the compression of a spring. The knee, and therefore the leg, has gravitational potential energy, or the potential to do work and overcome the gravitational force. Therefore, by displacing the leg vertically, the knee is overcoming the gravitational force. Gravitational energy is a conservative force, and it transfers then into kinetic, or moving energy by placing the foot down on the ground to propel the body forward. As the foot hits the ground, the shock is absorbed by the leg, and friction occurs at the knee to maintain posture, resulting in a loss of energy to its surroundings. This process repeats itself as the adjacent leg contracts the quadriceps muscle to act upon the knee.

Newton’s Third Law states that “for every action there is an opposite and equal reaction” (Serway, 2012), and that forces always exist in pairs. When walking, the leg contacts the ground to propel the body forward. When the foot drives against the ground it exerts a frictional force, and the ground pushes back with the same amount of force generated by the step. The driving force behind this action is the knee; the knee is responsible for creating and transferring the force to the ground, as well as absorbing the impact of the force being pushed back upwards. As Newton proposed, single isolated forces in nature cannot exist (Serway, 2012). The normal elastic force (from the foot) balances the gravitational force (from the ground). Without the third law, locomotion would not exist.

As aforementioned, the knee’s anatomy includes built-in “shock absorbers” (i.e. the meniscus). Moving the whole body requires momentum, which relies on the knee to do most of the work. The meniscus helps the knee hold the force of the body’s weight. When walking, running, jumping, or even falling, the body’s change in momentum will equal the impulse that the floor exerts upon it. The knee has no control over the amount of impulse that will be exerted upon it, and will be forced to compensate in order to stabilize the rest of the body. It is however,
possible to decrease the amount of force to the body by bending the knees while landing (or
landing on a soft surface).

The biomechanics of the knee is the role of kinetics (forces) and kinematics (relative
motion of rigid bodies). Along the patellofemoral joint, the forces acting upon it are the highest
due to the short “arm” of the extensor mechanism (Scuderi, 2010). The required force for the
quadriceps muscle to extend the knee is ten times the ankle load when at a 120° flexion angle.
The forces on the joint range from 385N while walking, to 2500N while climbing stairs, and
6000N when landing from a jump (Scuderi, 2010)! [See Figure 2 and 3.] While walking, the
pressure (dividing the force by the contact area along the femur and patella) range between 1.5-
2.5 MPa. Under a load, however, this pressure can be as high as 55 MPa when landing from a
jump. Tightness in the muscle also has a significant impact on joint reaction forces (specifically
during resting and sitting). The stronger the muscles, the greater support they can provide to the
knee joint, thus limiting the risk for injury. Given the magnitude of the forces and pressures
being applied to the area, the patellofemoral joint is strongly influenced by muscle tightness,
body weight, and impact sports (Scuderi). Any abnormalities, such as the posterior sagging of
the tibia, or the anterior sliding of the femur, can lead to higher patellofemoral reaction forces.

Torque, or the tendency of a force to rotate an object around a fulcrum (Serway, 2012), is
the knee’s Achilles heel. While forces push and pull an object, torque twists and pulls it along a
fulcrum, or axis. In the human body, the fulcrum is the knee. Torque is what can easily lead to
injuries, the injury being proportional to the amount of torque being applied. If the knee is not
stabilized, the femur will move at the same time as the tibia and fibula do, resulting in twisting or
damage at its connection (i.e. the knee joint). The resulting injuries can be as simple as a loss of
balance, to a torn ligament (ACL the most common), or the dislocation of the “floating” patella.
Moreover, chronic “use and abuse” can also lead to a degenerative breakdown of the knee joint.

Another danger to the knee joint, and perhaps the most common condition leading to
knee arthroplasty, is arthritis. Osteoarthritis, a chronic degenerative condition that leads to
compression and abrasion at the joint, resulting in softened and eroded articular cartilage, is the
most common (Merieb, 2013). Rheumatoid arthritis is also a debilitating issue. This
inflammatory disorder affects the knee’s synovial membrane, leading up to a build-up of fluid
that thickens the membrane and forms scar tissue along the bone. This scar tissue then ossifies
and fuses the bones together, completely immobilizing the joint (Merieb, 2013).

The solution to these irreversible injuries is the partial or total knee arthroplasty, or the
surgical replacement of some or all of the weight bearing surfaces of the knee joint. In the
United States alone, over 600,000 total knee replacements are performed every year (Cram et al.,
2010), a statistic that is rapidly growing. In fact, total knee arthroplasty is now the most common
major surgical procedure each year. In the total knee arthroplasty, surgeons enter the knee joint
anteriorly to insert a prosthesis, a metal implant cemented into place, to replace the articulation
between the femur and the tibia, and/or the femur and patella (Kurtz, 2004). During the process,
a sufficient amount of bone is removed to level the joint line, tightened ligaments are relaxed,
and mechanical alignment is carefully checked to allow for optimum load sharing between the
prosthesis and the bone (Palmer et al., 2014).

Although the design, fixation, and materials used in each implant varies upon the need, the
typical design for a total knee implant is made up of three different sections: the femoral, tibial,
and patellar components. The femoral component is typically made of metal and made to fit
around the end of the femur. It contains a groove meant to mimic the intercondylar fossa on the
femur, allowing the patella to glide up and down as the knee bends and straightens. The tibial
component is generally a flat, metal platform that includes a polyethylene (plastic) spacer. The patellar implant is a dome-shaped piece of polyethylene, meant to mimic the kneecap. If the patella needs replaced, this component is included in the arthroplasty (BoneSmart). [See Figure 4.]

There are currently many different types of total knee prosthesis available: [See figure 5.]
- Fixed bearing
- Medial pivot
- Rotating platform and mobile bearing
- Posterior cruciate ligament (PCL)–retaining
- PCL-substituting

Fixed bearing implants are the most common. Its name is derived from its structure, as the polyethylene component is fixed firmly to the metal base. The femoral component then rolls over the plastic. Medial, or mobile bearing implants are the most suitable for younger, more active patients, as they are generally designed for higher performance and more wear and tear. The design is very similar to fixed bearing implants, however the polyethylene component can rotate short distances in the medial portion of the tibial tray. Medial pivot implants is the most natural-feeling prosthesis for the knee, as they replicate the rotating, twisting, bending, flexion, and stability of the natural knee. Moreover, this design is arguably more stable than the natural knee, as it doesn’t allow for the roll back of the lateral side of the knee when the medial side rotates, a natural flaw in the human knee. PCL retaining implants are used when the posterior cruciate ligament is kept in place during the surgical procedure. The implant includes a notch to accommodate the ligament. On the other hand, PCL substituting implants are used when the PCL needs to be removed. The implant has a raised portion meant to compensate for the missing ligament (BoneSmart).

While total knee arthroplasty has existed in some form for over fifty years, the complexities of the knee were not fully understood until about thirty years ago. There are records dating back to 1860 of tissue arthroplasty as a form of treatment for arthritis, using skin, muscle, fascia, fat and big bladders (Palmer et al., 2014) to replace the articular cartilage. Metallic molds for the femur and tibia appeared in the 1930s (Scuderi et al., 2010), however neither mold was firmly fixed to the underlying bone. Additionally, the molds consistently loosened with movement. The fusion molds were later replaced by a hinge design that improved articulation with the patella. It was after WWII and the Korean War that the need for durable, mobile replacement joints, resistant to the corrosive organic acids in the blood (Merieb, 2013) became a necessity. As soldiers returned from the battlefield, there was a suddenly a tremendous need for strengthening and replacing body components that transplantation at the time simply could not accommodate (Hanker et al., 1988). This led to a surge in biomaterials and biomedical implants, a field of study that would impact physicists, chemists, engineers, biomedical and clinical scientists for years.

The creation of prostheses (artificial joints) for medicine (Merieb, 2013) required the employment of artificial materials to replace structural components of the body. These materials needed to be biocompatible and assume the mechanical roles required for fracture fixation (for example, weight bearing). While metals, polymers, and ceramics have been used, metals have been the most successful (Ratner, 2004). This required that the prosthesis be implanted into the surrounding tissue of the area in order to be integrated into the body. In order for the implant to
be successful, it must have surface characteristics that would promote interaction with the appropriate eukaryotic cells (Hanker et al., 2010). In other words, successfully integrate with the proper blood cells and fibroblasts, while being careful not to promote bacterial growth and adversely affect the biological environment (Ratner, 2004). In like manner, the implant count not be adversely affected by the host tissues and fluids. The prosthetics created for orthopaedic applications included a combination of cobalt and titanium alloys, tailored to any degree of hardness, flexibility and tensile strength required (Hanker). These compression molded plates could be screwed in place and then reinforced with cement or methyl methacrylate. Each year, significant advances are being made in the type and quality of the different metals, polyethylene, and ceramics being used in the prosthesis manufacturing process. Each advancement leads to improved longevity of function of the prosthesis (Simon et al., 2014). Currently, a prosthesis used for total knee arthroplasty will last about ten-fifteen years (without excessive strain), and restore 80% of the original joint function (Merieb, 2014). Most importantly, they help to reduce pain.

Aside from reproducing the mechanical roles of the knee joint, tension is also an important factor in total knee arthroplasty. During orthopaedic procedures, a handheld magnetic sensor is used to measure soft tissue tension in a similar way that the tension in a string would be measured. The sensor uses a pressure-mapping technique that allows surgeons to adjust for misalignments during surgery. Balancing the tensile tissues in the collateral and cruciate ligaments surrounding the bone leads to restoration of original limb function, and longevity of the implant (Signal et al., 2012).

The procedure is also quite adaptive, and certainly not one-size-fits-all. The total knee arthroplasty can be modified by the surgeon to fit the specific needs of the patient. In the case of arthritis, if the patient has only a mild form and only one of the condyles are affected, the surgeon can perform a unicompartmental knee arthroplasty. If both condyles have been diseased, but the patella is untouched, a bicondylar knee arthroplasty can be performed. If both condyles and the patella have been diseased, a tricompartmental knee arthroplasty can be performed. In all cases, an ultra-high molecular weight polyethylene (UHMWP) is used as a polymeric component with the metal implant (Kurtz, 2004). [See Figures 6 and 7.] Today, UHMWP is the most prevalent, widely used material for articulation with metallic components in total knee arthroplasty.

Conclusion:

As a future physical therapist, I have seen an over-abundance of total knee replacement patients in the short time that I have been studying this field. In fact, the knee arthroplasty was the single most common rehabilitation I experienced while observing physical therapists in action. I truly believe that this form of medicinal physics will continue to grow and improve upon itself as it is likely to be even more common as the years pass.

While technology in the field has come a long way, it is still relatively primitive in the grand scheme of things. The current models are not yet suitable for the young, strong, sports-active patients who are often middle aged and pushing their joints to new levels (Scuderi, 2010). The prosthesis works itself loose over time, so engineers and doctors alike are working to enhance the fit between the implant and the bone. Despite the fact that people are becoming more self-aware and are more open to non-invasive procedures to better their health, humans are
flawed and accidents still do happen. Hereditary conditions still exist, and sometimes factors such as arthritis cannot be avoided.

Computer aided technology is working to advance these designs to allow for integration of bone marrow stimulation, grafting and implantation, and even stem cell regeneration (Merieb, 2013)! Despite this, it is possible that the incredible demand for total knee arthroplasty over the past twenty years will resurface as these older generation prosthesis models wear over time and need replaced with the better, improved model. This possibility represents significant clinical and economic implications (Cram et al. 2012). Studies have also shown a tremendous growth in total knee arthroplasty in the elderly, in part by enrollment in free-for-service Medicare beneficiaries, which currently represent about 60% of all total knee arthroplasty procedures.

Another important concern that will affect this growing field of medicinal science is obesity. Growing obesity rates are of particular relevance in the United States and have been linked as a risk-factor for osteoarthritis. Studies have recently shown that obesity can have a negative impact on the outcome of knee arthroplasty (Kerkoffs et al. 2012). Considering that obesity can lead to osteoarthritis and knee damage, factors that end up leading to knee replacement, this creates a bit of a conundrum if being obese can severely impact the result of the surgery. It proposes the question in the form of the chicken and egg analogy: which came first, the obesity or the knee injury that caused a lack of movement? If an overweight patient undergoes a total knee replacement, will the prosthesis survive long enough for the person to lose the weight?

The total knee arthroplasty has revolutionized the quality of life for millions of people, reducing their debilitating pain and returning function into their everyday lives. In addition, it has unified the scientific community into collaborating with a common goal: to continue to advance the design and implementation of this prosthesis to determine what other orthopaedic applications can be improved upon. I for one am excited to see where this field of study goes, as I am sure I have only seen the beginning.
Figure 1. General structure of a synovial joint (Merieb)

Figure 2 and 3. Two-dimensional model in the sagittal plane representing the force vectors on the patellofemoral joint in the knee (Scuderi)
Figure 4. A total knee replacement prosthesis before implantation (Cram et al.)

Figure 5. Different types of total knee implants (BoneSmart)

Figures 6 and 7. Example of knee prosthesis using UHMWPE
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The Physics of Applied Scanning Electron Microscopy

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ABSTRACT

Since the 14th century, scientists have been using some form of magnification to examine objects on a smaller level. In 1938 the way that microscopes function was changed for all upcoming generations as Ernst Ruska integrated physics into his designs for a magnifying apparatus known as the scanning electron microscope. The incorporation of physical principles, such as the use of electron streams to create detailed images, has allowed advances and discoveries throughout the research community to increase without an end in sight. Due to many technological advances, the basic principles of microscopy have been enhanced by integrating key concepts of physics, allowing the field of laboratory research to improve through the use of scanning electron microscopes.

History of Scanning Electron Microscopes

The use of magnifying lenses has allowed individuals within the scientific community to examine minuscule objects for hundreds of years. Since the late 1500’s, scientists such as Robert Hook and Anton Van Leeuwenhoek began devising magnification apparatuses to help them understand the world around them in greater detail. Although rudimentary, the notion of constructing a device which enhances observable details was one of the first steps for scientific advances to come. While many other varieties of microscopes were being implemented throughout the scientific community, the introduction of the first scanning electron microscope (SEM) permitted scientists to examine new aspects of specimens which had never before been seen. Max Knoll assisted Ernst Ruska, a student being mentored by Knoll, in the construction of the first electron microscope for his Ph.D. thesis project in 1931. Through experimentation, Knolls and Ruska essentially created the first electromagnetic lens which could magnify the image using a stream of focused electrons in place of an illuminating agent. Ruska made modifications to the lens in 1930, later working with an engineer from the Siemens Company to produce the first fully functioning electron microscope. Although crude, the first version of the SEM provided scientists with a way to visualize viruses for the first time, allowing them to conduct research with much greater detail than ever before.

Throughout the years, scientific research has advanced by applying the intricate principles of physics to the scanning electron microscope. Important discoveries throughout the late 1900’s have solidified the relevance of scanning electron microscopes and the astounding concepts used to devise such advanced equipment. Specifically, events such as the identification of the first ebola outbreak in 1976, followed by the designation of the disease’s causative agent, filovirus, in 1989. Countless other identifications have been made possible by using the cutting-edge technology provided by the combination of physics and microscopy.
An electron microscope falls into the category of scientific instruments which are specifically designed to utilize a stream of electrons to create an image of a particular specimen. The electron microscope is able to achieve greater levels of magnification than a traditional optical microscope, which uses only a combination of light and varying degrees of magnifying glass lenses to achieve relatively low levels of magnification. Since electron microscopes have a much greater resolving power than a traditional microscope, scientists are able to observe their specimens in much more detail.

An ordinary scanning electron microscope runs at a high vacuum. A concentrated stream of electrons is created by a source such as a field emission gun or a tungsten filament. The beam of electrons is subjected to a high voltage typically measuring 20000 volts. After passing through the voltage, the electrons are considered to be accelerated and continue through a combination of electromagnetic lenses and chasms, producing a concentrated electron beam. Scan coils within the machine employ the electron beam to examine the surface of the specimen. Furthermore, the particle being examined emits its own electrons due to the scanning beam. The electrons are then collected by a detector, producing the results displayed on a screen to be viewed by the person operating the microscope. Moreover, the collection of data when using an SEM involves more than a simple beam of electrons. As the sample experiences the beam of electrons, those electrons remain within a specified area of the sample known as the interaction volume. The electrons scatter throughout this area, causing interactions between the specimen and the beam produces four different types of signals, including x-rays, secondary electrons, visible light due to cathodoluminescence, and backscattered electrons.

**Physics of a scanning electron microscope.**

Kinetic energy plays a large role in an SEM. The electrons in an SEM provide very detailed images through kinetic energy transfer. The electrons in an SEM are accelerated and carry an impressive amount of kinetic energy. Kinetic energy is defined as the energy an object possesses due to the motion of that object, as demonstrated by the equation in figure 1. The total value of the object’s kinetic energy and potential energy is considered to be the total mechanical energy of that object. The formula for conservation of energy seen in figure 2 shows that the energy of an isolated system may only change between forms and is not destroyed or created. Indeed, the electrons in an SEM are continually in motion, utilizing kinetic energy to do work on the particles being evaluated. The energy is dispersed as the aforementioned signals produced by the interactions between the electrons and the sample when the incident electrons are decelerated in the solid sample.
One of the ways that the kinetic energy of the electrons is dispersed is through backscattered electrons. During the process of molecular collisions between the electrons and the particle being examined, electrons scatter from the surface as they are separated from their corresponding atoms. Of the electrons which have been scattered, those which have lost relatively small amounts of energy are classified as backscattered electrons\(^4\). As the sample is scanned by the beam emanating from the SEM, the high energy electrons create inelastic interactions between themselves illustrated by figure 3, which is comparable to the interactions of the nuclei of atoms on the surface of the subject being examined\(^4\) Fig. 3. An inelastic interaction is an interaction in which part of the kinetic energy is changed to some other form of energy in the collision\(^6\). Consequently, the information regarding the backscattered electrons and their locations are transmitted to the SEM's computer and interpreted as images. The intensity of the backscattered electron signal is also affected by the composition of the sample\(^4\). When backscattered electrons emanate from an area of the sample with a greater area, they experience interactions with the sample as they escape to produce secondary electrons further away from the origin of the backscattered electrons. Consequently, the resolution of the image will be decreased due to this event. Compromised image resolution will occur less frequently in specimens with relatively high atomic numbers\(^7\).

Another factor with the potential to enhance or inhibit the quality of the images can be controlled by adjusting accelerating voltages, as resolution varies with changes in the area from which secondary electrons are produced. Correspondingly, higher accelerating voltages will increase secondary electron output from further regions of the specimen. When secondary electrons are produced further away from the original spot, the resolution of the image is reduced, decreasing fine detail. This concept is illustrated in figure 4 as the Fine detail of a diatom imaged using a lower accelerating voltage of 5kV is clearer than the image of the same diatom when observed at 20kV, a much higher voltage. The resulting image at a higher voltage reveals a notable decrease in resolution and contrast\(^8,Fig 4\).

**Secondary Electron Imaging**

Although backscattered electrons are important to the creation of intensely detailed images, secondary electrons provide the most popular route of imaging for SEMs. Of the four different signals created by the electron beam, the formation of surface structures depends mainly on the secondary electrons being produced. Uniquely, the secondary electrons exist at a lower energy level than the other signals, allowing them to be absorbed by the sample when produced within the zone of interaction, shown in figures 5 and 5.2, exists directly beneath the electron beam\(^7\), Fig.5.5.2. Because the secondary electrons are only weakly negative, they are deflected by the positive pull they experience from the Faraday Cage, an enclosure which utilizes conducting materials to block electrical fields\(^8\), which surrounds the secondary electron detector to form the resulting images. Accordingly, secondary electrons produce better results than backscattered electrons. Even though backscattered electrons can also be produced deep within
the sample, their higher energy state allows them to escape the deflection of the Faraday Cage. This feature of backscattered electrons has the potential to disrupt the results produced. However, very few of the escaping backscattered electrons interfere with the secondary electron signals.

While it is true that the scanning electron microscope is a phenomenal application of physics in the scientific community, it is important to realize that special training is required to operate the microscope. Along with specific knowledge regarding the use of the device, the microscope is very large and expensive as observed in figure 6, requiring a large storage area to protect it from unwanted vibrations and external magnetic fields9.

Physics Applied to Life Sciences

Because the scanning electron microscope is an immensely detailed instrument, it is often excluded from the category of practical scientific equipment. With this in mind, it is important to highlight the many uses of an SEM in life science settings. Within the field of science, physics is incorporated into many of the techniques used to conduct research. Markedly, micro-imaging has provided an excellent example of the possibilities for investigational and clinical work. By utilizing the aforementioned principles of physics, the simple light microscope once used to complete the imaging has progressed with the use of electron microscopy, allowing significantly higher levels of resolution to be observed. The increased levels of magnification have provided scientists with a technique that can identify features within cells, such as the recombined chromosomes of cancer cells10. Physics has the innate ability to go beyond providing the tools for scientific studies.

Notably, molecular and cellular biology demonstrates an amazing application of the complicated physics at work in an SEM. Originally introduced in 1965, the scanning electron microscope (SEM) has provided a three-dimensional perspective in biological research. An image produced by an SEM image far surpasses a sample viewed with a hand lens or a dissecting light microscope. With its great depth of field and increased resolution, the scanning electron microscope is unmatched when it comes to providing the user with a magnified image of the specimen showing immense details not visible with a light microscope8.

Comparatively, microbiology combines medicine and physics on a sub-cellular level, examining microorganisms such as bacteria and viruses. Ultimately, physics has provided many of the conceptual answers to many of the questions asked by scientists. Implementation of concepts from physics has allowed the scientific community a way to study information regarding the physical form of hereditary information and the division of cells10. One particularly impressive aspect of biology that could not exist in the absence of physics is the study of viruses. Although electron microscopy is not a relatively new viral observation method, this technique remains the standard in clinical viral studies, such as those shown in figures 7 and 8, surrounding the Avian Flu Virus Fig.7, as well as the HIV retrovirus Fig.8. Astonishingly, the scanning electron microscope has also been used to observe single virus particles invading a cell and producing countless copies of its own genome in figure 9, where a bacteriophage is shown...
actively invading an E. coli cell⁹. Numerous historical observations have been made using electron microscopy, leading to the identification of emerging viruses. A few noteworthy viral discoveries are the imaging of the poliovirus in 1952, as well as the differentiation between the chicken pox and smallpox viruses in 1948².

The Future of Scanning Electron Microscopy

Undoubtedly, the discovery of scanning electron microscopy has opened countless new doors for scientific research and discovery through the years. This radical invention has allowed scientists to visualize and compare things that were once only imagined. While it is true that seeing is often believing, there is always room for improvement when it comes to the accuracy of the images produced by this groundbreaking machine. The scientific community is continually coming together, combining and refining their abilities to provide fast, accurate results. In February 2014 the US Department of Energy’s Office of Basic Energy Sciences held a two-day workshop, bringing together many experts in differing fields to identify and discuss advances in electron microscopy that could potentially lead to new scientific discoveries¹¹. One particularly interesting topic focuses on the advantages of using electrons for microscopy, allowing this specialized field of imaging to become more widely used and accepted than it currently is. One key aspect is the interaction advantage of using electrons over neutrons and x rays. The interaction between electrons and matter is coupled 10⁵ more strongly than an x ray and can be easily focused due to their characteristic charge, an advantage which is lost with the use of an uncharged neutron¹¹.

Another area of interest for future uses of electron microscopy is complex 3D imaging, along with chemical identification of impurities using different states of electrons. This idea requires scientists to master the concept of multidimensional microscopy, in which there is a combination of momentum space, real space, energy, temperature, time and mechanical stress¹¹. This concept requires scientists to explore the unsolved mysteries of condensed matter physics, a challenge that will surely lead to more phenomenal discoveries and advances in the field of electron microscopy and consequently, the field of Biology.

Conclusion

Overall, it is my desire to see physics integrated into the field of laboratory science in a way that allows researchers to make new discoveries regarding the cellular processes and systems of viruses and bacteria. With more advanced equipment comes the ability to understand the organisms that are constantly interacting with humankind. A greater understanding of the interactions between the cells and their targeted hosts will allow researchers to develop new ways of combating these organisms before taking over our body systems. Studying the intricate details of the viruses responsible for many cases of human cancer, as well as the human immunodeficiency virus, is an excellent use of the amazing technology that has been provided through physics. The advancement of microscopy will allow the medical and scientific community to stay ahead of the fast-adapting organisms who cause these devastating illnesses.
With the ability to provide extraordinary magnification for analysis of many different compounds, the scanning electron microscope is an extremely valuable application of physics within the scientific community. A scanning electron microscope is essentially a microscope which produces an electron beam to scan back and forth across a sample. Due to the interaction between the sample and the beam, multiple signals are produced to provide the user with detailed surface structure information, differences of atomic number within the sample, and information about the elemental content. Few other scientific research tools are able to provide information spanning such a wide range of research fields. From applications in nanotechnology to examination of viral specimens, the creation of the SEM allowed the exploration of microscopic materials to expand further than was previously thought possible. Given these points, physics can clearly be accredited with countless scientific achievements. The importance of physics has been thoroughly demonstrated over the years in the field of microscopy with invention of the scanning electron microscope. The radical new ideas being investigated and studied would not be possible without the basic rules and concepts provided by physics.
Figures

Fig 1: Formula for Kinetic Energy
\[ KE = 0.5 \cdot m \cdot v^2 \]

Fig 2: Conservation of Energy.
\[ KE_i + PE_i = KE_f + PE_f \]
the energy a system starts with doesn't change at a later time or position.

Fig 3: Elastic and Inelastic Collisions

Before
\[ m_1 \]

After
\[ m_1' \]
\[ m_2 \]
\[ m_2' \]

Fig 4: Diatoms under SEM with differing voltages.

Fig 5: Faraday Cage Diagram.

Fig 5.2: Path of electron beam.
Fig 6: Scanning Electron Microscope⁷.

Fig 8: HIV retrovirus. gstatic.com

Fig 7: Avian Flu Virus. http://blogs.wayne.edu

Fig 9: T4 bacteriophages infecting a live *E. coli* bacteria cell. http://www.hyglos.de
References


Space Exploration: Complications and New Possibilities of Exploring the Cosmos

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Abstract

Space exploration requires travelling vast distances at high velocities that may or may not be attainable. Relativity is an important concept regarding interstellar exploration as it presents a notion of time travel. Time dilation can be caused by velocity and gravity, but may be negated via shortcuts through space to explore the outer regions of the universe. Researchers believe the universe operates with the properties of waves, and new understandings of communicating information between two points in space pose a possible means of travel for exploration. New technologies are developing that increase the efficiency of the next generation of space exploration vehicles.

Discussion

Exploring the unknown and searching to discover new information has always played a role in the development of human civilization. The boundaries of next generation space exploration continually push further beyond Earth’s atmosphere. Researchers examine new understandings of the functional laws of the universe, because new technologies are needed to extend interstellar exploration millions of miles beyond the current limits of space exploration. However, scientists and future explorers must have a solid foundation in known theories of physical laws in order to adapt, modify, or adjust concepts to new discoveries and acquire new possibilities to interstellar exploration.

Albert Einstein’s special relativity theory is an important concept when looking at interstellar space exploration as it discusses what is absolute and what is relative about time, space, and motion. The basic postulate of the theory of special relativity states there exists no such state of absolute rest in the real world. Relativity depends on the observer, for a glass of water sitting on a table may be to the left of one observer, but to the right of another depending on where he or she is sitting. If the glass is full of water, all observers should agree to that fact, regardless of their positions. In this example, the position of the glass would be a relative statement, whereas the glass being full of water would be an absolute fact regardless of the observer.

Velocity can be a relative concept. If a blue car speeds past a red car on the freeway, the driver of the red car would observe that the blue car was moving relative to him being at rest. However, from the blue car’s perspective, her car would be stationary relative to the red car moving in the opposite direction. The relative positions of the blue or red car, and which is moving depend on which driver’s point of view is considered. Each driver observes a different velocity of the passing car relative to his or her position.

The best understanding of the concept of time comes from Albert Einstein’s theories of relativity. Prior to these theories, time was widely considered as absolute and universal, the same for everyone no matter their physical circumstances. In his special theory of relativity, Einstein proposed that the measured interval between two events depends on how the observer is moving. Crucially, two observers who move differently will experience different durations between the same two events.
The twin paradox best describes the effect of time on two different observers. Imagine Steve and John are twins, and Steve travels at high speed by rocket ship to a nearby star, turns around and comes back to Earth. Steve’s duration of the journey could be one year, but upon returning he finds that ten years have elapsed on Earth. Steve and John are now different ages despite being born on the same day. The example illustrates a limited type of time travel and the effects of relativity on two different observers.

Both gravity and velocity affect the relative concept of time. Time dilation is an actual difference of elapsed time between two events as measured by observers either moving relative to each other, or differently situated from gravitational masses. Velocity time dilation occurs whenever two observers move relative to each other. Time dilation is proposed to be dramatic only when motion occurs close to the speed of light. Atomic clocks are accurate enough to record a shift in time that is stretched by motion during jet flights. Even if the time shift of atmospheric air flights is only nanoseconds, travel into the future is a proven fact with the technologies of today.

Gravity is another means accounting for a limited type of time travel. Gravitational time dilation occurs as time runs more slowly for matter in strong gravitational environments than objects in weak gravitational environments. Einstein’s general theory of relativity predicted that gravity slows time. A clock at the top of a mountain will run a bit faster than a clock on the beach beside the ocean, because the clock on the mountain is further away from the center of the Earth. Similarly, clocks run faster in space than on the ground. The effect of gravitational time dilation on Earth is miniscule, but it has been directly measured using accurate clocks. Sailors, cruise missiles, and cross-country travellers rely on Global Positioning Systems, which account for gravitational time-warping effects. Without accounting for gravitational time dilation, each would be many miles off its chosen course. Time dilation on Earth may have minimal effects, but can be greatly increased at the higher speeds of space vehicles, satellites, and exploration probes. Faster velocities, stronger gravitational forces, or both acting on a single object will slow the relative time of the matter experiencing the velocity, or gravitational force when compared with standard Earth time.

Time and velocity can be considered relative concepts depending on an observer’s frame of reference. However, the speed of light in empty space is an absolute value, regardless of an observer’s position, and is a constant fixed value. For any inertial observer, any light signal moves through empty space with the same constant speed, \( c = 299,792,458 \) meters per second, independent of the motion of the light source. The speed of light is the only speed that is independent of the observer and absolute; therefore, it plays the lead role in all of special relativity. Light speed defines the absolute speed limit for the transfer of energy, matter and information in the universe. According to Einstein, no object, however strong the forces acting upon it, can ever be accelerated to light speed.

The speed of light is a ubiquitous parameter in the equations of special relativity. If the relative velocity is small compared to the light constant, so are the relativistic effects. When velocity approaches the light constant, the relativistic effects become prominent. The relativistic effects of the speeds encountered in everyday life are almost unnoticeable.
as velocities are miniscule compared to the light constant. However, if explorers in space could increase their velocities near the light constant, relativistic effects would dramatically increase.

The concept of spacetime is absolute; however, space and time independently are not. Spacetime is the totality of all events, but can be broken into different snapshots of simultaneity. Different observers can come to differing conclusions about simultaneous events when looking at snapshots of events at a specific time and orientation in space\(^1\). Relativity’s region of influence on an event can be represented graphically with time measured by a single observer the vertical axis and space the horizontal axis (Figure \(^a\)). Light is a constant diagonal across each quadrant of the graph. Einstein’s famous equation \(E=mc^2\) demonstrates that every body of mass will have a total energy, where \(c\) is the constant speed of light\(^1\). The relativistic increase of mass happens in a way that makes it impossible to accelerate an object to light speed. The faster the object already is, the more difficult any further acceleration becomes. The closer the object's speed is to light speed, the greater the increase in inertial mass. To reach light speed exactly would require an infinitely strong force acting on the mass. According to Einstein the relationships enforce special relativity’s limit that no material object can be accelerated to light speed\(^1\).

Not being able to accelerate a body of mass to the speed of light presents a problem with high-speed space exploration of the future. Objects at incredibly high speeds increase inertial mass, which means an infinitely strong force is required to continue to accelerate the mass. An enormous amount of energy is required to counteract the increase. Either new sources of energy are needed for future exploration, or alternate pathways are required to span the vastness of space in a short period of time.

The effects of relativity can be demonstrated using particle accelerators. Particles that decay at a certain rate have a particular half-life. When a particle with a specific half-life is placed into an accelerator and its velocity greatly increased, the particle remains intact longer than its rate of decay specifies. In essence, time has slowed down for the particle that has been accelerated to great speeds relative to the world around it. The rate of decay has slowed down for the particle, or in other words the particle has been accelerated into the future due to a great velocity and time dilation. The concept of time dilation states a moving clock is slower than a stationary one. Particles accelerated to high speeds extend their inner clocks and decay more slowly\(^1\).

Understanding the concepts of relativity and the absoluteness of space, time and motion provide insight into space exploration. Future space vehicles that can accelerate to very high speeds will alter the concept of time for its boarded passengers. A possible solution to long distance space exploration and the effects of time dilation is utilizing short cuts through the fabric of space, also known as wormholes.

A wormhole is conceptualized as a theoretical bridge between two regions of spacetime that would be traversable, meaning people or objects could pass through from one side to the other side. Wormholes are also called Einstein-Rosen bridges since the concept was developed during Einstein’s studies on relativity. The bridge, or shortcut through space, would be a fold in the fabric of space that connects two otherwise extremely distant
regions in space. The alternate path could allow explorers the ability to travel long
distances over a short period of time, and possibly alleviate the issue of time dilation.
Special relativity’s determined speed limit would also be avoided by taking a shortcut
through two connected points through space and time (figure b).

Studying wormholes allows physicists to take theories they more or less understand like
general relativity and quantum physics, put them together, and see what breaks3.
Wormholes would be a handy device for traveling halfway across the universe in the
blink of an eye. Identifying a wormhole, or learning how to create and stabilize a
wormhole presents problems to space exploration as none are currently known to exist.
Researchers continue to study the theory and possibility of the existence of a shortcut
across space, and recently a team of physicists concluded it might be surprisingly easy to
make a wormhole traversable3.

Wormholes naturally fit into the general theory of relativity, whereby gravity warps not
only time but also space2. The theory allows analogues to alternative routes connecting
two points in space. Just as a tunnel passing under a hill can be shorter than a surface
street going over the hill, a wormhole may be shorter than the usual route through
ordinary space2. Relativity in wormholes stands up just fine. As outlandish as it might
seem to move ships and people faster than the speed of light, Einstein’s rules permit it
because wormholes provide shortcuts across the fabric of space and time3. Wormholes
circumvent the need to accelerate matter to great velocities to cover the extreme distances
between locations in space.

One problem with relativity is it shows a wormhole can be turned into a time machine,
allowing a paradox to be created if people go back in time and alter historical events.
Scientists have discovered wormholes, like black holes, have event horizons: regions
beyond which not even light can escape. The event horizon, sitting squarely in the throat
of the wormhole, would prevent a traveller from leaving the wormhole he or she had
entered3. Nature seems to have appointed a cosmic censor to prevent backward time
travel paradoxes from occurring. Wormholes might only be plausible across half of the
spacetime fabric, allowing a shortcut across distances in space, and yet not affect time.
More research is required to determine if wormholes exist, and what exact limitations
they may contain.

It seems plausible that wormholes can be used to traverse great distances through space.
To make a wormhole traversable, negative energy exotic matter would be required to get
rid of the event horizon, allowing a traveller to enter and exit the shortcut. Negative
energy has less energy than an equivalent volume of empty space3, and could banish the
event horizon to make the wormhole passable. Quantum mechanics require that such
exotic matter is forever being created and annihilated on tiny scales3, which seems
inadequate to get rid of event horizons. A large amount of negative energy would be
required for stabilization so more research is needed to better conceptualize travelling
through wormholes.

For a wormhole to be traversable it must contain exotic matter. Exotic matter is
something that generates antigravity to combat the natural tendency of a massive system
to implode into a black hole under its intense weight2. Antigravity, or gravitational
repulsion, can be generated by negative energy or pressure\(^2\). Negative energy is known to exist in certain quantum systems, suggesting that the laws of physics do not rule out exotic matter, although it is unclear whether enough anti-gravitating exotic matter can be assembled to stabilize a wormhole\(^2\). Physics allows for anti-gravitational exotic matter, and new discoveries through continued research may lead to a means of collecting large quantities for specific uses such as stabilizing wormholes through space to traverse long distances over a short period of time.

A problem that stands in the way of shortcuts through the fabric of space is the creation of the wormhole in the first place\(^2\). Space may be threaded with such structures naturally as remnants of the big bang that remain currently undetected. Alternatively, wormholes might come into existence on tiny scales, such as Planck lengths, which are about 20 factors of 10 as small as an atomic nucleus. In principle, such a minute wormhole could be stabilized by a pulse of energy and then somehow inflated to usable dimensions\(^2\) with exotic matter. New technology, or future concepts of physics yet to be discovered are required to be able to create, enlarge, and stabilize wormholes. It is conceivable that the next generation of particle accelerators will be able to create subatomic wormholes, as new discoveries continue to be realized on the quantum mechanics scale.

The laws of physics as they are widely understood today do not prohibit the existence of a negative mass of matter as once thought, but the density and the amount required to create a wormhole are breathtaking\(^4\). It is possible to conceive the idea of a wormhole existing somewhere in the universe, but the concept would require an astronomical amount of negative exotic matter to fathom even a small portal stable enough to exist through spacetime, if even only for a few seconds. From a practical point of view, the problem with wormholes of traversable dimensions is that they require exotic matter, which is real matter you can stop in a laboratory that has negative mass. The amount of negative exotic matter required parallels the need for a Jupiter mass, roughly 2\(\times\)10\(^{27}\) kg, concentrated in a region of small dimensions\(^4\). The only plausible source of a compact Jupiter mass of exotic matter must be something latent in the world as we find it that can be exposed in some way\(^4\) to harvest and compact to usable proportions of great density. If negative mass exotic matter were to be available, researchers would still need to manipulate the exotic matter in unknown ways to create a portal through two points in space.

Exotic matter would need to be compacted into an incredibly dense form to supply the necessary energy for a wormhole. Researchers have theorized the concept of negative exotic matter by stating that given some modest amount of everyday type matter, say a few hundred or thousand kilograms, all they have to do is enclose the matter within another presumably thin shell of matter wherein they can change its mass from positive to negative\(^4\). The matter would have to become sufficiently negative to null the positive mass of the initial mass of the shell and the matter it encloses. By doing so, it would screen the gravitational influence of the matter in the rest of the universe on the matter within the shell\(^4\). That, in turn, for distant observers, would render their masses negative and about 21 orders of magnitude larger than their original positive masses, which would provide the Jupiter mass of exotic matter needed to make a wormhole\(^4\). To the traveller in the wormhole everything would look normal per locally measured variants, but someone outside the wormhole would see a major spacetime distortion\(^4\) present at the wormhole.
Exotic matter is a term that refers to matter, which would somehow deviate from a normal activity and have exotic properties. The concept of exotic matter, or matter that violates certain physical laws, is not a new idea. Exotic matter is at the basis of many intriguing theoretical possibilities, such as wormholes, time machines, and even so-called cosmological doomsday models of the universe in which the universe’s energy density continually increases⁵.

Exotic matter is vital to theoretical concepts of interstellar space exploration. Exotic matter could conceptually contain new and unique properties to provide a source of renewable energy, fuel, stability to wormholes, or possibly fill gaps between developing concepts of physics in the universe. An exotic material with unique properties violating the weak energy condition in particle physics leads to a light cone with regions of positive and negative total energies. Energy in the light cone continuously transfers from the negative region to the positive region, resulting in perpetual motion⁵. However, theories involving continuous energy flows are considered inherently unstable⁵.

Perpetual motion is a theoretical concept many scientists tend to avoid as it can be considered exotic by stating that perpetual motion is movement that goes on forever. Thermodynamics cannot support perpetual motion systems, as they are impossible to sustain. The law of conservation of energy states energy can neither be created nor destroyed, however it can change forms and flow from one place to another. In other words, the total energy of an isolated system does not change. Some theories state perpetual movement would generate an endless stream of energy. An exotic material with unusual properties could cause energy to flow continuously between different regions of space, resulting in a runaway transfer of energy⁵. Advancements in technology allowing the construction of a device to capture runaway energy from a perpetual motion system would provide an endless source of energy that could be used to fuel space exploration vehicles.

Along with the concept of exotic matter is the relationship of dark matter with the universe. Dark matter, proposed decades ago as a speculative component of the universe, is now known to be the vital ingredient in the cosmos. Dark matter is six times more abundant than ordinary matter, one-quarter of the total energy density, and the component that has controlled the growth of structure in the universe⁶. Dark matter is known to exist, and carry large amounts of energy, but very little is known about the actual properties of the substance. The introduction of dark energy became prevalent to researchers when it was realized that there was not sufficient matter to explain the structure and nature of the universe⁶. Dark energy along with dark matter is present in abundance in the universe, and fills gaps of previously misunderstood notions of the universe. Neither dark matter nor dark energy emit or absorb light, so testing theories of their properties while tied to a visible light spectrum is incredibly difficult. A better understanding of dark matter is a vital component to future space exploration.

Dark energy has changed the view of the role of dark matter in the universe. According to Einstein’s general theory of relativity, in a universe composed only of matter, it is the mass density that determines the geometry, the history, and the future of the universe⁶.
With the addition of dark energy, the story of the universe is different. First, what determines the geometry of the universe is whether the total energy density equals the critical value, where now the dark energy contribution is added to the mass portion of the composition of the universe\(^6\). Secondly, the period of matter domination has given way to dark energy domination\(^6\). Dark energy coincides with the exotic material dark matter, but is now becoming the dominating component of a previously matter driven universe.

The future is determined by the nature of dark energy, which is sufficient to cause the current expansion of the universe to accelerate, and the acceleration will continue unless the dark energy should decay or change its equation of state\(^6\). New understandings of physics and the way energy and matter interact in the universe will be introduced to incorporate dark matter and dark energy into current concepts. Dark matter and its relationship with dark energy could hold a key to future space exploration, as some researchers believe the universe operates similarly to the properties of a wave.

The three media of the universe – time, space, and energy – are waved together. Each has its own properties and behaviors, but one cannot exist without the others\(^7\). Theoretically, each media has neither beginning nor end, and they are all one entity. Nevertheless, the media are changeable, depending on the different phases of energetic matter in which they appear and decay together\(^7\).

In space, different energy levels appear from time to time. The energy activity brings about the appearance of swirls, which cause changes of time and space. The appearance of energetic activity causes the appearance of space, which causes the appearance of time, which coincides to the concept of spacetime. It is impossible to describe the existence of space and time without energetic activity\(^7\).

Because energy appears in wave formations, researchers believe that time appears in the same form and oscillates together with energetic flow. Every phase has its time when energetic matter transitions to the next phase, or to space\(^7\). Energetic matter does not disappear rather it transfers like a wave, with time, to other phases or formations such as dark matter, and dark energy\(^7\). Energetic matter connects each wave to its neighboring formations, which belong to greater formations, and so on. Together, the wave formations belong to the universe, with its defined space, energy and time\(^7\). The universe is one large wave with a defined space that oscillates with energetic properties. Changes in energy levels are transmitted immediately throughout the wave, as the universe behaves according to the same laws of energetic matter that govern the smallest wave\(^7\).

In different energy formations, such as dark matter, communication across the fabric of space is by another means, not light, which allows the exchange of signals by energetic matter to be greater than the speed of light. Wormholes can be described as highly energized spaces connecting the dark matter of different galaxies in the universe\(^7\). Dark matter links together forming a communication network between distant spaces of the universe, and allows the transfer of energetic matter across the universe as a whole at an instantaneous rate. Dark matter is responsible for the rigidity of galaxy waves\(^7\). The wave theory of the universe brings the whole universe together as primarily composed of one
rigid dark matter that transfers signals immediately, as one rigid formation, conforming to Einstein’s idea of absolute time.

Dark space maintains proportions between energetic matter and time. Time, space, and energy are eternal but changeable; depending on the energetic activity in their phases, as all are maintained in constant proportions. The media appear and decay together in every phase; they protect all information of the universe, or the quanta of energy, space and time. The universe is composed of matter and its swirling movement creates different energetic formations like a wave. Dark matter, which is in continuous contact with every structure in the universe, maintains the structure with the constant interchange of energetic matter relationships. Vibrations transmit signals throughout the universe; thus, dark matter is the most important formation in the universe. The signals are transferred in extremely condensed time frames, which are irrelevant to the size of the waves in the universe.

If dark matter is the framework that connects the universe, and signals are transferred instantaneously across the universe through dark matter like a wave, then tapping into those signals provides a network to the entire universe. Dark matter could be the means by which matter or objects transfer from one location to another. The limitations of accelerating mass to high speeds would not affect travelling from one galaxy to the next through the dark matter network, as objects would be communicated near instantaneously via the dark matter as non-light formations. Hopping on the dark matter signaling network presents a possible opportunity to be transported across the fabric of the universe in an instant, but discoveries are needed on how to access the dark matter network.

Space exploration requires energy, whether that energy propels a vehicle, stabilizes a wormhole, or opens the universal wave dark matter network to use as a transportation system. New energy systems to propel rockets and vehicles through space are at the forefront of research as improvements can be made to existing systems. Electromagnetic research poses promising results as an efficient propellant of next generation spacecraft. Advancements in specially conditioned electromagnetic research may overcome obstacles preventing the breakthroughs in energy research that are desperately needed for future terrestrial and space power and propulsion.

Near-term payoffs for specially conditioned electromagnetic beams are predictable for beamed energy systems that accomplish rocket propulsion. Beamed energy systems utilize lasers or microwave beams to heat air and propellant inside vehicles. If electromagnetic energy loss could be reduced by special conditioning or modulating of the laser beam, the beam power for vehicle thrust generation could be 2 to 4 times greater than the most promising laser wavelengths. Specially conditioned lasers result in a beam that is insensitive to refractive index fluctuations caused by air density variations, and allow specific propulsion direction via very precise pointing of the beam. Aneutronic fusion reactions are an energy and propulsion breakthrough that emit no neutrons and result in no radioactivity. An enormous challenge for such fusion is limiting radiation lost from intensely excited electrons and ions before the high energies and temperatures needed for fusion of aneutronic fuels are reached. However, aneutronic fusion shows
promise as an energy source to propel craft through space without the complications of radioactivity exposure to human passengers.

Space is a vacuum with no vehicle drag, but continued efficiency in aerodynamics has benefits linked to the exploration of other atmospheric planets, as well as benefiting current commercial air travel. Plasmas formed by electromagnetic discharges that are emitted from the front end of vehicles and interact with oncoming airflow to reduce vehicle drag are of interest for future power and propulsion systems that would reduce vehicle size and the fuel and propulsive energy needed for high-speed atmospheric flight. Electromagnetic discharges of modest intensity from the front end of vehicles into air reduce vehicle supersonic drag, resulting in engine power savings of 15-20 times the electric power that must be expended for discharge generation. Plasma reduces drag by weakening shock waves and bifurcation of the atmosphere. Vehicles that utilize plasma discharges would increase fuel efficiency and speed of travel.

Another promising technology to power space vehicles utilizes photons. Mastering photon propulsion is proposed to be the key to overcoming the limit of the current propulsion technology based on conventional rocketry and potentially opening a new space era. One important factor in rocketry is the efficiency of energy transfer from the propellant energy to the spacecraft kinetic energy. Among propellants capable of delivering relativistic spacecraft velocities, photons are most promising because their high flux generation is much technologically simpler and lighter than that of particles. Photons can be recycled to amplify thrust using high reflectivity mirrors. By recycling photons, the propellant system increases efficiency, consumes less fuel, and would near a self-sustaining energy system. The emerging high power high energy efficiency laser and high power optics technologies show great promise in scaling up photon recycling propulsion to propel interstellar spacecraft.

Travelling through space at great velocities presents many challenges. The most serious problem is that humans cannot travel at the speed of light because the relativity factor becomes an imaginary number. It is very difficult to provide evidence that the possibility of travelling faster than the speed of light may or may not exist. However, there is anomalous behavior in the cosmos that provides such evidence. If the gravitational field of a black hole is so strong that even light cannot leave, jets created by the black hole itself that arise are evidence of a physical phenomenon that moves faster than light speed. The physical phenomenon ties in with the wave theory of the universe as dark matter is considered the communication backbone that instantaneously transfers non-light signals across the universe.

Regarding field equations and relativity, Einstein introduced the term cosmological constant to account for a static universe, but the effort was unsuccessful. The static universe described by the theory was unstable, and observations of distant galaxies by Hubble confirm the universe is not static but expanding. The expansion of the universe presents opportunities for dark energy, gravity, or a yet to be discovered exotic concept to open new knowledge on the universe. Learning how to manipulate gravity, transmit
matter via energetic waves through dark matter, or obtain shortcuts through points in space present opportunities to continue to explore undiscovered regions of the universe.

Conclusion

Exploring new frontiers and discovering unknown truths about the workings of the world have been central concepts in human history. As new understandings of the world around us emerge, technologies are developed that manipulate or break current understandings of the laws of physics, which lead to additional discoveries. Science fiction continues to become science fact, and space continues to be the destination for new discoveries as very little is actually known about the operation of the universe.

Earth is a tiny spec in the universe, and we have pushed beyond our atmosphere with satellites, space telescopes, and exploration craft to the moon, Mars and beyond. The Webb telescope is more powerful than Hubble and in development to eventually orbit hundreds of thousands of miles away from Earth. Many cosmic systems are such extreme distances away that a faster, or alternate mode of travel is required to open the possibility to human exploration. We can increase the velocity of vehicles through technologies and new self-sustaining fuel systems that provide the required energy, but relative limitations exist due to the vast distances between known galaxies, as well as Einstein’s relationships between energy, mass and velocity. Research will continue for alternative pathways of space exploration to alleviate the limit of accelerating matter to the speed of light. Wormholes and the universal network of dark matter are two promising alternatives to interstellar exploration. Creating and stabilizing shortcuts across space will open the possibility to explore galaxies once thought unreachable. Continuing to research dark matter and how energetic activity is communicated across the universe could lead to a new transmission medium, possibly warping matter across the universe via dark matter information waves. Both concepts show promise, but much more research is needed.

Particle accelerators open new understandings of the laws of physics on a microcosmic scale. Information gained from these experiments can be used to correlate the size of the Earth to the universe. As new experiments are created, physicists at Fermilab and CERN will make discoveries of new concepts about the creation and manipulation of currently unknown exotic particles. These applications will expand, or alter theories about the universe, and may be applied to verify the existence of wormholes, or possibility create portals on a larger scale, connecting the relationship of Earth with the universe.

In the mean time, new technologies like electromagnetic plasma pulses, or photon propulsion systems, will benefit current commercial applications such as international travel, renewable energy resources, and sustaining the natural resources of Earth through more efficient usage. Renewable fuels, self sustaining energy systems, more efficient jets, and vehicles all help utilize the resources we currently have on Earth to sustain what’s available, yet get places more quickly. Exploring space, pushing the limits of physics, and obtaining new discoveries about the workings of the universe will be applied to current systems, primarily transportation and information communication, which will increase travel efficiency, attainable speeds, and unite humans across the world.
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Figure a Relativity diagram of observed light cones in time and space.
Diagram courtesy of www.wikipedia.org

Figure b Diagram of a wormhole depicting negative energy event horizon in the throat of the wormhole used to stabilize the shortcut through space.
Courtesy of www.daviddarling.info
Another perspective of a wormhole diagram depicting a shortcut through two points in space. The beam of light traversing a path between two points in curved spacetime takes a longer journey than straight through the wormhole. Courtesy of [www.pbs.org](http://www.pbs.org)
Blasters and Battleships: A Brief Overview of Laser Technology and Modern Laser Weaponry

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Honors General Physics 111, 19002

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Abstract

Lasers are an integral part of modern science. They are the driving force behind many modern machines and technologies and are involved in most of the cutting-edge scientific research today. This paper seeks to explain the basic physical properties and mechanics of lasers, including the history of their development, their application to the real world and their use as weapons in the modern military.

Introduction

From popular culture and science fiction to the very real world of science and physics, lasers are an integral part of our daily lives. Whether geeking out to blaster battles on Tatooine, scanning items at a grocery store or simply looking at a cell phone, lasers permeate nearly every aspect of normal life, transcending fictional entertainment and practical application alike. In academia, lasers are integral to virtually every researching science and find a bounty of uses as tools, or in some cases, as weapons. But what are lasers? How do they work and how are they being used today?

What is a laser?

The word ‘laser’ stands for ‘light amplification by stimulated emission of radiation’. On its most general level, a laser is a special kind of light. Light is important to understand before talking in-detail about lasers. It is part of the electromagnetic spectrum – a fancy way of saying that light is one of many possible ways in which radiation (the transmission of energy through space or a material) behaves. Light has many different properties – it has frequency and oscillation, and thus behaves as a wave, but also obtains certain particle properties, despite having no mass. A ‘piece’ of light is called a photon. Each photon is imbued with a quantity of energy, given by

\[ E = hf = \frac{hc}{\lambda} \]

where \( h \) is Planck’s constant, equal to 6.6x10^-27, \( c \) is the speed of light (299,792,458 m/s in a vacuum) and \( \lambda \) is the wavelength of the light. As given by this equation, the energy in a photon increases as the light wave decreases (Ready 1978).

Normally, light is produced when electrons jump between energy states within their atomic orbits. Adding energy to the atom (usually in the form of heat) causes its electrons to jump into ‘levels’ of higher energy. Eventually, the electron returns to a lower level by shedding their excess energy in the form of a photon. This is how many light-giving appliances in the modern world work – light bulbs and electronic signs, for example (Miloni 1988).

But what about lasers? The above process is random and sporadic. Electrons jump between energy levels chaotically, releasing photons without discernable order and in many different directions. Lasers, on the other hand, are typically focused and precise. How does a frenzied mess of photons turn into a straight, orderly wavelength of light?
The answer lies in the last few letters of the ‘laser’ acronym: stimulated emission. As it turns out, there is another way in which electrons can emit light besides randomly releasing photons whilst hopping between energy levels. If an excited electron (excited, meaning already stimulated by some kind of energy) is hit by a traveling photon, the electron will release a second photon of identical phase, frequency, polarization and direction of travel. The electron still drops in energy state, but rather than firing off a photon at random, it releases one in perfect order with the photon that stimulated the reaction to begin with – a clone of the original photon. If one were to cluster a group of excited electrons together and fire a single photon into their midst, the resulting reaction would create a unified beam of photons, all moving in the same direction with identical properties. This is, fundamentally, how a laser works (Thyagarian, Ghatak 1981).

![Figure 1](image)

As shown here, a photon (hv) interacts stimulates an excited atom (given from its resting state in E1 to its excited state in E2). After the photon interacts with the atom, its electrons return to the atom to its resting state and an identical hv is produced alongside the original hv.

Typically, this light is refined by positioning the excited electrons (often called a laser ‘median’) between two parallel mirrors. The electrons begin jumping energy levels and releasing photons as usual, but the majority of them end up exiting the system, reflected off of the mirrors at differing angles. Only those photons that are aligned perpendicular to the parallel mirrors remain in the system, bouncing back and forth between the two reflectors in the world’s most boring game of Pong. A hole in the center of one of the mirrors allows the uniform light particles to pass through in a unified beam of focused photons – laser light at its finest.

**Kinds of lasers**

Using this basic principal, a number of different kinds of lasers arise. Though they vary according to the specifics of their mechanics, the way in which they produce light,
in their build and in the chemical composition of the fuel, all are concerned with directing specific light through stimulated emission.

Solid state lasers (SSL) use a solid lasing medium – usually a glass rod or a crystal, gem (or often a ruby) – to direct light. The first optic laser in the world was an SSL. These lasers usually pair the rod or gem (called a ‘host material’) with an active material, such as chromium, neodymium, erbium, holmium, or titanium. Neodymium is the active material in the most widespread applications. SSL’s are very easily adjusted and altered to specific variables. Achieving a desired joule-output is usually as easy as flicking a switch (Olson 2011). Electrons within the host material are stimulated by an energy source, which results in stimulated emission across two parallel mirrors. While this is one of the easier lasers to create and use, they are often not as capable of the power output that chemical lasers or gas lasers are.

A chemical laser utilizes a chemical reaction to create population inversion (a system of atoms in which more particles are in an excited energy state than a lower energy state – as explained earlier, necessary for stimulated emission) in the lasing medium. An example of this kind of laser is the Mid-Infrared Advanced Chemical Laser (MIRACL) developed in the mid-1980s. The MIRACL is a continuous-wave, mid-infrared laser. “Its operation is similar to a rocket engine in which a fuel (ethylene, C2H4) is burned with an oxidizer (nitrogen trifluoride, NF3). Free, excited fluorine atoms are among the combustion products. Just downstream from the combustor, deuterium and helium are injected into the exhaust” (Olson 2011). At this point, the deuterium is mixed with the fluorine to create high-energy deuterium fluoride molecules. All the while, helium continues to stabilize the reaction and keeps the temperature in check. The exhaust gas heats the laser median, resulting in optic energy. The rest of the system functions much like an SSL, in which the ensuing light bounces off of mirrors or a ‘host material’ into a refined beam of photons. Chemical lasers are often preferable because, depending on the chemicals used, they can produce much higher energy than most SSL’s.

Gas lasers are another type of commonly-used lasers. Technically a chemical laser, gas lasers use a pure gas or gas mixture to create a beam. “The typical gas laser contains a tube with mirrors on each end. One end transmits the beam out of the cavity. Most gas lasers use electron-collision pumping, with electric current passing through the gas” (Olson 2011). The helium neon laser is, perhaps, the most famous gas laser. It emits a bright red, continuous beam of a relatively low intensity. This is the beam used in many laser-scanners and alignment tools, such as certain price checkers and laser pointers. It has many practical applications in the modern world. University students may use them in optical training laboratories, since the lower-power output poses few health risks. The helium neon laser (HeNe laser) even has uses in helping other kinds of lasers, with some larger blasters containing HeNe inside the beam path to help correct beam alignment. They are a very cheap and rugged laser, capable of working continuously for thousands of hours (Olson).

Other kinds of popular gas lasers include carbon dioxide lasers and argon lasers. A carbon dioxide laser operates in a manner similar to the helium-neon laser. It uses an electric charge for gas pumping whilst simultaneously using a percentage of nitrogen gas as a pumping gas. This kind of laser is traditionally much more powerful than a standard HeNe laser, capable of power outputs of up to 10 kilowatts. Argon lasers can go even higher, but require considerably more energy to power (Ohanion 1989).
Here is a nice illustration of a helium-neon gas laser. As demonstrated by this figure, helium atoms are energized by electrical discharge, causing their electrons to rise in energy level. When they collide with neon atoms an energy transfer occurs and photons are produced. The depleted atoms return to more basic energy levels, and the process repeats.

**History of the laser**

The history of the laser, like most scientific breakthroughs, involves the collaborative work of many different scientists over an extensive period of time. Max Planck was perhaps the first to seriously theorize how a laser might work by theorizing about early photons. “Planck deduced the relationship between energy and the frequency of radiation, essentially saying that energy could be emitted or absorbed only in discrete chunks – which he called quanta – even if the chunks were very small” (Rose 2010). His work was closely observed and expounded upon by Albert Einstein. In 1917, Einstein proposed ‘stimulated emission’, arguing that electrons could be stimulated in such a way that light would be emitted at specific wavelengths (Billings, Charlene 2006).

The production of an actual laser begins with the invention of something else entirely – the maser. Standing for ‘microwave amplification by stimulated emission of radiation’, the maser was developed in 1954 by Charles Townes and Arthur Schawlow. Though not able to use visible light, the maser had application in enhancing radio signals. Although these two gentleman theorized that a device could be built that that would “use infrared and/or visible spectrum light” (Rose), their speculations did not come to fruition by their hands.

All of this changed in 1960 when Theordore Maiman invented the ruby laser, considered by many to be the first optical laser. This is contested through the invention of
Gordon Gould, who began building an optic laser in 1959, but failed to patent his invention. With the floodgates opened, a host of other laser archetypes came to the forefront. The first gas laser was invented by Avi Javan in 1960, also having the distinction of being the world’s first continuous-light laser. The first semiconductor injection laser was invented in 1962 by Robert Hall, and is still used in many electronics today (Rose).

![Components of the first ruby laser](image)

**Figure 3**

Given here is the first every optic laser. The power supply provides energy to the quarts flash tube, which generates a bright beam of light. These photons enter the laser median – the ruby crystal that the quartz flash tube is wrapped around – and stimulated emission begins. At this point, the photons bounce back and forth between the two reflective mirrors till only those photons with the same direction and oscillation remain. They then exit through the hole in one of the mirrors, resulting in a laser.

**Applications of lasers**

Closer to modern day, lasers found many applications in modern medicine from the 1970s and on, particularly regarding eye health. In 1970, Stephen Trokel patented the Eximer laser. Initially used to etch onto silicone computer chips, technicians and medical experts realized the potential of using a laser to cut through human skin. A laser could remove biological tissue without inflicting collateral heat damage at a precision far greater than a blade. Trokel would take this knowledge to the next level, performing the first laser surgery on a human eye in 1987, striking an interesting irony; the part of the human body most easily damaged by lasers is the eye (Ready, 1978), and yet laser retinal surgery provides an exactness and clarity in operation that is unattainable through most conventional means (Mastropasqua, Carpineto, Ciancaglini, Falconio 1999). Over the next ten years, this knowledge was refined and channeled into many other fields of medical health. Today, lasers play a role in virtually every field of medicine. They are
used to shrink and destroy cancer tumors, to remove decay within teeth and reshape damaged gums, to reconstruct arteries and repair heart defects, and much more.

In some cases, lasers are viewed as the conduit through which the most advances sciences in the world will be realized. The research of A. K. Shakhbazyan (2009), for example, maintains that laser technology is the key to make mammalian cloning possible, “that the use of laser instead of the conventional cell engineering methods for the key cloning stages may have certain advantages”. He believes that continues investment into laser research and will open new avenues of scientific exploration in the field of cellular biology. Indeed, lasers are an absolutely crucial part of the medical industry, and advancements continue. “The increasingly larger number of clinicians involved in research on laser surgery adds even more to the diversity of laboratory projects that will in turn become the basis for techniques that in the future will gain clinical acceptance” (Wolbarsht 1984).

But the use of the laser goes far beyond the medicine. They have extensive applications in nearly every field of science, including engineering, communications, measurement applications, holography and art (Billings, Charlene). Most modern day machinery, whether used in practical engineering projects or in research advancements, involve lasers in one way or another (Billings, Charlene). Lasers also play a role in the entertainment industry. Aside from sparking our imaginations in science fiction books and movies, they are an integral part of many performing arts. Laser light shows began manifesting in the early 1970s and became an archetype of psychedelic entertainment. The shows were usually paired with a live musical performance on a stage, or with pre-recorded music. Genres of all type found a place in laser shows, from the heavy metal grunge of Led Zeppelin to the softer beauty of the Electric Light Orchestra. “Just as the laser itself was seen as ‘a solution without a problem’ when newly invented, laser shows are taking visual effects to heights that could have never been imagined a few decades ago. The continuing evolution of laser technology could give rise to even more spectacular effects in the future. The brilliant, focused colors of laser beams offer endless possibilities for ongoing innovation at the intersection of art and science” (Daukantas).

Lasers as weapons

Of equal interest is the role the laser plays in the weapons industry. The late 1970s and 1980s marked an invested time period for developing lasers into weapons. All branches of the military and industry were working their hardest to master high power levels, beam manipulation, and adaptive optics. In 1999, the Department of Defense formally recognized lasers as potential tools of destruction and began researching as to how most effectively weaponize them. Perhaps the most considerable government-funded project concerning laser weapons came about in 2000, during which time the Joint Technology Office for High Energy Lasers was formed. Intended as a one-stop laser research facility, the JTO was designed to bring all laser technologies together to develop a complete laser weapon system that could be used defensively (Olson 2011).

Developing lasers as weapons has many advantages. Traditional projectile-based weapon systems are inhibited by gravity and air resistance and subject to relatively slower velocities and the potential for collateral damage. Lasers, on the other hand, is completely unaffected by gravity, causes minimal collateral damage due to its accuracy
and exactness, travels at the speed of light, precisely and exactly reaches far distances, and is capable of causing exact, predetermined amounts of damage to targets.

Generally, a laser is considered a weapon if it is used against someone and packs more than 50 kW to megawatts of power. This extreme output of energy is obviously much greater power than commercial lasers. Accordingly, they have greater support and maintenance needs. Weaponized lasers must maintain environmental standards, must meet chilling requirements, must constantly upkeep their mirrors and must be refueled constantly (Olson).

Of important note is the attitude of the rest of the world towards laser weaponry. Although the United States makes leaps and bounds in the field, they are still restricted by international law. In 1995, the UN General Assembly put strict rules down as to how laser weaponry should be approached. “The review Conference adopted on 13.10.1995 a new Protocol IV prohibiting the use and transfer of blinding laser weapons and on 3 May 1996 adopted an amended version of Protocol II on mines, booby traps and other devices” (United Nations 1980). Like chemical weaponry and nuclear/atomic armaments, laser technology is heavily scrutinized and is closely watched to ensure the protection of human rights.

So how far has laser technology in the military come in the United States? Certainly much farther than the first ruby laser. In 2015, the Department of Defense announced its intention to spend $371 million on lasers technology and research. Due in no small part to extensive research done in the 60’s and 70’s, the laser weapons currently being tested can provide extensive defensive capabilities at a much lower cost than traditional ammunition. Laser weapons “can deliver shipboard defense at $1 per shot” (Resnick 2015) through the use of existing battery packs.

But modern-day laser weaponry is more than just the stuff of prospective research. In early December of 2014, The Office of Naval Research reported that its laser weapons system, christened LaWS, had performed above expectation in a test tour aboard the transport dock USS Ponce, located in the Arabian Gulf. The LaWS underwent a series of rigorous trials from September to November that tested the weapon’s accuracy, damage capability and maintenance requirements. The laser is in active use. “We're not testing it any more. This is operational. It's on a ship in the Persian Gulf,” reported Rear Admiral Matthew Klunder. “This isn't something we've got in a box we're saving a special moment. We're using it every single day” (Lendon 2014). The Navy expects to be using laser technology en mass as soon as 2020. If these numbers prove true, Americans may expect lasers on the decks of every major naval ship very soon, possibly within the next five years (Lendon).

Although the Navy is the most prominent military institution making strides in laser technology, the US Army is also experimenting with their own laser weapons. Their High Energy Laser Mobile Demonstrator, a mobile truck mounted with a high-intensity laser, is currently being tested. In an examination in September 2014, the weapon performed exceptionally well. “The High Energy Laser Mobile Demonstrator (HEL MD) successfully took out drones and 60mm mortars under sub-optimal conditions during tests in Florida earlier this year. Fog and wind did not detract from its performance, which paves the way for the weapon to be used at sea” (Ernst 2014). The HEL MD has been on the drawing board for nearly a decade, only now seeing action and giving an opportunity to prove itself. Though not quite on the high plateau occupied by the Navy’s
LaWS, it certainly shows much promise for the army in the future, who may end up mounting them on their vehicles and trucks in much the same way the navy plans to mount their lasers on their ships. Only time will tell if these sci-fi visions of laser cannons on the sides of trucks and battleships will prove true, but the military is certainly taking an interest in them.

![Figure 4](image)

Shown above is the LaWS, the latest in laser weapon technology. A fully-operational laser cannon, the LaWS is capable of shooting down planes, ships, and even fast-moving missiles out of their air. The LaWS could revolutionize naval weapon systems, the machine being considerably less expensive than traditional guns (each blast only cost about one dollar) and far more accurate (Lendon).

But the military is not the only industry examining the capacity of lasers as weapons. Other engineering companies are also breaking boundaries on laser weaponry technology. Lockheed Martin’s latest laser weapon is a machine of astounding power. Known as ATHENA, (Advanced Test High Energy Asset), this ground-based prototype laser ate through an entire car engine in a matter of seconds. The beam was fired from more than a mile away. “Fiber-optic lasers are revolutionizing directed energy systems,” remarks Keoki Jackson, Lockheed Martin chief technology officer. “We are investing in every component of the system – from the optics and beam control to the laser itself – to drive size, weight and power efficiencies. This test represents the next step to providing lightweight and rugged laser weapon systems for military aircraft, helicopters, ships and trucks” (Fisher 2015).

What makes lasers so destructive to begin with? Their physical properties have already been examined, so what about a stream of energized photons is so deadly? On the most basic level, lasers represent a closely-ordered concentration of highly-energized particles. Shining a laser on any material distributes energy from the beam into the
surface of the target. If the energy distribution is high enough (meaning that the beam intensity is great enough), the sheer heat residue resulting from beam-contact could cause extensive damage to the target.

This, of course, raises more questions. Are certain mediums more easily damaged by laser fire than others? What does color have to do with laser intensity and destructive power? To answer these questions, a test team (coincidentally, the writer of this paper) conducted a series of experiments to test the destructive nature of lasers on different mediums. Far from the exciting realm of blowing up planets or disintegrating whole trees, the test team opted instead to examine how quickly it took a green laser to destroy different colored balloons.

**Experiment**

The following materials were used during the experiment: one green 500 mw 532 nm laser; red, green and yellow water balloons; red, brown and black sharpies.

The test team began by measuring which colors of balloons are most susceptible to green laser fire. Both green and yellow balloons sustained absolutely no damage – despite having the laser trained on the yellow balloon for nearly fifteen minutes, nothing happened. The red balloon, however, was destroyed almost instantly, perhaps a half second delay interceding from the point where the laser made contact with the balloon’s surface to the moment when it exploded.

The test team next tested whether or not the inflation level of the balloons had any affect on how quickly they were destroyed – or if it made them more easy to destroy. Three balloons of each color were inflated to different lengths, tale to head: 2”, 4” and 6”. In every scenario, there was no change in the time it took to explode them. Those that did not initially explode continued to remain intact.

At this point, it was clear that red balloons were particularly susceptible to green laser fire. This raised an interesting question. Would the other balloons explode if they were given a bit of red coloring? To test this hypothesis, the test team took some red sharpie marker and colored a yellow balloon on its edge. The balloon exploded as if it were a red balloon. The same happened to the green. The next tests involved laser stimulus with different colored sharpies. Blue had no effect, but brown and black both exploded the balloons at an even faster rate than the red had. It seems that darker color is more susceptible to laser fire than lighter.

<table>
<thead>
<tr>
<th></th>
<th>No Marker</th>
<th>Red Marker</th>
<th>Brown Marker</th>
<th>Black Marker</th>
</tr>
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<tbody>
<tr>
<td>Green Balloon</td>
<td>No Effect</td>
<td>.5</td>
<td>Near Instantaneous</td>
<td>Instantaneous</td>
</tr>
<tr>
<td>Red Balloon</td>
<td>.5</td>
<td>.5</td>
<td>Near Instantaneous</td>
<td>Instantaneous</td>
</tr>
<tr>
<td>Yellow Balloon</td>
<td>No Effect</td>
<td>.5</td>
<td>Near Instantaneous</td>
<td>Instantaneous</td>
</tr>
</tbody>
</table>

Still curious, the test team decided to color a red balloon green and see if the green laser would be unable to pierce it. They hypothesized that the green marker would protect the red balloon. Unfortunately, they were wrong. The laser penetrated the green-
colored red balloon with no lag time. It should be noted however, that the combination of red and green produced a particularly dark color. Perhaps the color of the balloon itself is not as important as the shade of the material, darker colors being more susceptible to laser fire, lighter colors being less.

The answer to this phenomenon lies in color, and the susceptibility of different colored objects to laser fire. Objects have color because their chemical composition reflects certain kinds of light. Green objects reflect green light, red objects red light, but black, reflecting none, “absorbs most of the light in the visual spectrum” (Abel 2001). Being the most energy-absorbent ‘color’, black (and all darker colors, for that matter) are far more susceptible to laser fire as opposed to lighter colors, which repel more kinds of colored light on the light spectrum.

As far as military application goes, this little fact probably does not have much relevance. Most industrial lasers are so powerful that color is not a significant factor. A military laser will burn through white and black materials alike with little lag time between.

Conclusion

Laser technology is an essential part of human life. It pioneers advancements in every scientific research field; extending the quality of human life in medicine, the efficiency of industry and the defensive capabilities of the military. Furthermore, it promises decades of continued advancement in all scientific disciplines, from engineering to cellular biology. It is, quite literally, the way of the future. Experimenting with the destructive capacities of lasers, I demonstrated that different colored mediums are more susceptible to laser damage than others and explored the role that color and reflected light plays in optic firepower.

Reviewing my paper thus far and examining the history of lasers, their application and modern development, it becomes clear to me that lasers are an extremely diverse and multifaceted phenomenon that readily lends itself to the disposal of the scientific community. In only fifty short years since lasers have begun to be used in earnest, they have revolutionized the way medicine, industry and technology operate. That said, it is clear that laser technology has taken a drastically martial turn since the first ruby laser was developed in 1960. With 371 million dollars spend on laser weapon development in 2015 alone and with a fully-operational laser cannon already in use by the US military, laser weaponry is clearly at the forefront of military technology and will continue to play an important role in the way the world’s militaries operate.

I look forward to this future with both excitement and apprehension. On one hand, I eagerly await the development of laser blasters with the giddy anticipation of a child. Lasers are cool. Realizing the dreams of science fiction and making reality out of what was previously mere fantasy would be a phenomenal achievement. With laser technology pioneering so much of the future, a part of me wishes that we waste no time in developing the machines of tomorrow, sparring nothing in our quest to advance as far as possible as quickly as can.

That said, I am also fearful. Laser research, like all sciences, is intricate and beautiful, but is ultimately assigned purpose only through the people that to use it. Science is a tool in the hands of men, an unstoppable force moving in whatever direction
we point it at. At this moment, laser technology is pointed in the direction of violence and warfare. While having the means to defend ourselves is important, we must remember that laser have continued application far outside of the realm of war. In the words of Arthur Galbston, “Nothing that you do in science is guaranteed to result in benefits for mankind. Any discovery, I believe, is morally neutral and it can be turned either to constructive ends or destructive ends.” With this in mind, we must constantly strive to steer science towards good and virtuous aims, seek to guide our inquiry in the pursuit of knowledge and the general welfare of the rest of mankind. Lasers have broken countless boundaries already, and they continue to provide unprecedented innovation into scientific inquiry. It is our responsibility to use them for good, to favor wisdom and peace over the sword, to constantly strive to use our understanding for the betterment of the entire world.
Figures

Figure 1 – An illustration of stimulated emission. (Parry-Hill 2015)

\[ E_2 - E_1 = \Delta E = h\nu \]

Figure 3 – A diagram of how a Helium-Neon laser operates. (Ohanian, 1989)
Figure 3 – An illustration of the first ruby laser (Livermore 2015).

Figure 4 – A photograph the Navy’s LaWS. (Lendon 2014)


The Life and Death of a Low Mass Star
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Spring 2015
Professor J. Weitz
There are seven phases of low mass stars. The first phase is the protostar phase. This is the phase where the star is born. The second phase is the main sequence phase where the star will spend the majority of its life. The third phase is the red giant phase where the star starts to die off. The fourth phase is the helium-burning phase. In this phase, the helium core contracts and begins to heat up. The fifth phase is called the double-shell burning red giant. In this phase the helium fuses into carbon in a shell around the carbon core. The sixth stage is called the planetary nebula phase. This is the second to last stage and is where high temperature regions drive the outer half of the star away. The seventh and final stage is where the star becomes a white dwarf. It contracts to its final size and has no more fuel left to burn.

A star begins its life as a breakdown of molecular clouds into dense clumps. This is stage is known as the protostar phase. This phase can last anywhere from 100,000 to 10 million years, depending on the size of the star. There are a couple ways this breakdown could have started, a galaxy collision, or the shockwave of a nearby supernova. It’s usually difficult to see protostars in the visible spectrum because clouds of dust surround them and block the light they emit. A protostar becomes a main sequence star when its core temperature exceeds 10 million K (Las Cumbres Observatory). This temperature is needed for hydrogen fusion to take place. The amount of time a star takes to form depends on the size. So for low mass stars, it could take up to one hundred million years to form.
The second phase is called the main sequence phase and this is where the star spends the majority of its life. Most of the stars in the universe are in the main sequence. They are powered by the fusion of hydrogen into helium. They are also in hydrostatic equilibrium, which is when the outward force balances the inward force of gravity. At the end of the main sequence phase, it becomes impossible for any nuclear reactions to occur and there becomes no source of energy to create outward pressure to balance out gravity. This is when the core starts to collapse and heat up. Temperatures increase with the contraction, eventually reaching levels where helium is able to fuse into carbon (Nola Taylor Redd). The energy that’s produced by helium fusion makes the star expand way bigger than it’s original size.

The third phase is where the star begins to die off. This is called the red giant phase. Red giant stars reach sizes of one hundred million to one billion kilometers in diameter (sixty two million to six hundred and twenty one million miles), one hundred to one thousand times the size of the sun today (Nora Taylor Redd). Since the energy is spread across a larger surface, temperatures actually become cooler, reaching four thousand to five thousand eight hundred degrees Fahrenheit. This change in temperature makes stars shine in the redder part of the spectrum, which gives them the name red giant, even though they have an orange color to them. Stars can spend a few thousand and one billion years in the red giant phase. Eventually, the helium in the core runs out and fusion stops causing the star to shrink again until a new helium shell reaches the core. Once the helium ignites, the outer layers of the star are blown off in clouds of gas and dust known as planetary nebulae (Nora Taylor Redd).

Before the planetary nebula phase, the helium core contracts and starts heating up which is called the helium burning phase. This is the fourth phase in stellar evolution. Hydrogen fusion in the shell produces more helium (Australia Telescope National Facility). This gets put onto the core, which adds to the mass and makes it heat up even more. When the core reaches one hundred million Kelvin, the helium nuclei have enough Kinetic energy to form carbon in a two-stage process. This process is the main source of carbon and oxygen found in the Universe, also in our bodies.

The fifth phase in stellar evolution is called the double shell burning red giant phase. This is where helium fuses to carbon in a shell around the carbon core. Since there is no heat left, the core begins to shrink and helium burning moves into a spherical shell around it. This is when the core becomes a mixture of carbon and oxygen and becomes over a million degrees. The heat that is made by the star is from two shells. The outer shells burns hydrogen to helium, and the inner shell burns helium to add fresh carbon and oxygen to the core. After all this, the star becomes larger and cooler again.

Planetary nebula is the sixth phase. This is the phase where high temperature regions drive the outer half of the star away. As the hydrogen burning shell begins to burn its way to the stellar surface, it begins to turn itself off. The helium shell does the exact same thing. Since there are no heat sources, the heat that leaves the star is not replaced. With nothing holding the star together, it begins to shrink and burn out. Since the star is so bright, the heat leaves rapidly and the shrinking can be fast. When the star becomes hotter than about thirty thousand degrees, ultraviolet light strips electrons from hydrogen atoms and begins to expand and surround the star.

The final stage of stellar evolution is called the white dwarf. The star has no more nuclear fuel to burn and it contracts into its final size. When a star runs out of fuel it
collapses into itself. White dwarfs contain the mass of the sun and have the radius of Earth. This makes them really dense and the gravity on the surface is three hundred and fifty thousand times that of gravity on Earth. White dwarfs reach this incredible density because they are so collapsed that their electrons are smashed together, forming what is called "degenerate matter" (Nola Taylor Redd). This means that a bigger white dwarf has a much smaller radius than its less massive counterpart. Burning stars balance themselves by the inward push of gravity with the outward push from fusion. With white dwarfs the electrons must squeeze together tightly to create that outward pressing force. After all of this, white dwarfs usually fade away. They eventually give away all of their energy to become a black dwarf.
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The Physics Behind Prosthetic Limbs

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PHY-112
Dr. Casey Durandet
November 20th, 2014
Abstract

The field of prosthetics is a very exciting and innovative field that is often overlooked by the medical and technology communities. The development and innovation revolving around prosthetics is breaking new grounds for multiple different fields. This paper will review the history of prosthetic limbs, how prosthetic limbs relate to modern physics, and where the field of prosthetics is heading in the future.

Introduction

Prosthetics and orthotics is a growing practice in the medical field. Due to a growing number of amputee patients, the demand placed on the artificial limb industry has greatly increased. In 2009, about 2 million individuals have a lost limb in the US\(^1\). That number is growing at about 185,000 amputations per year\(^1\). This demand has put in place a movement in the industry to search for new technologies as well as to refine older methods of developing prosthetics. Along with the increased rate of amputations in the US, doctors, physicians, as well as prosthetists are searching for the major causes of this increase. Some feel the rise in military forces and demands have caused an outbreak in lost limbs and medical amputations. Forced trauma that causes or leads to an amputation currently accounts for 45% of the total amputations in the US\(^1\). Some common examples of forced trauma includes burns, severe deformation or bone fractures, severed limbs, and frostbite. The biggest cause of amputation in the US is cardiovascular disease. Cardiovascular disease accounts for 54% of the total causes for amputation in the US\(^1\). This is a major issue in the US for the lack of exercise, poor diet, and obesity among the population. Lowering risk factors for amputation as well as using modern technologies to develop more effective and useful artificial limbs are the main focus of physicians today.

History of Prosthetics

When most people think of early artificial limbs and body parts, most think of a pirate with the prosthetic hooked hand and wooden leg. In reality, artificial limbs and body parts have been developing about 3000 years before the rise of pirates. In fact, the first prosthetic body part is dated back to 1500 BC of a functional wooden toe found mummified Egyptian body\(^2\). Other than artificial legs, most early prosthetics was worn for aesthetic purposes only. Like most medical procedures and technologies, functional prosthetics did not start to develop until the renaissance period. In the early 1500’s Gotz von Berlichigen of Germany had functional iron hands made that could be moved into different positions with a series of springs and movable joints\(^2\). This advancement paved the way for functional upper extremity prosthetics. In 1529, a French army surgeon named Ambroise Pare developed new and innovated ways for amputating extremities and developing better fitting prosthetics. He also developed a functional above the knee prosthetic that has devices that are still used in modern prosthetics\(^2\). Pare is now known as the father of modern amputation and prosthetics. Other early prosthetic designs that are used today include concealed joint gaps, multi-articulate leg and foot, as well as suction sockets. The rise in amputees in World War I gave birth to the American Orthotic Prosthetic Association (AOPA). At the end of World War II, the American government set forth to develop new advances in artificial limbs. Most of these advances involved
using lighter modern materials such as aluminum, plastics, and composites to create a more functional yet still strong prosthetic. Moving towards today, advances in computer technology and robotic design has given birth to modular prosthetics. Modular prosthetics gives the artificial limb robotic function to allow the patient to carry out daily tasks. Even more innovative, is the combination of modular prosthetics with neural transmission. This allows the patient to use brain function to activate movements of the artificial limb as they would if their natural limb was still there. Neurotransmission advancement and modular prosthetics can even be used for regaining function among paralyzed patients.

BodyPowered Prosthetics

The majority of prosthetic limbs found today are functional but are not modular. This means that they provide function but are not operated by a microprocessor. All the forces involved with non-modular limb movement consist mostly of mechanical physics. Besides prosthetics being classified as modular and non-modular, the can also be divided based on the location of the amputation. The four major types of limb replacements are transradial, transhumeral, transtibial, and transfemoral. Transradial prosthesis is a prosthetic limb for an arm amputation below the elbow joint. Transhumeral prosthesis is a prosthetic arm for an amputation above the elbow joint. Transtibial prosthesis is a prosthetic limb for an amputation of the leg below the knee joint. Transfemoral prosthesis is a prosthetic leg for an amputation above the knee. Each location type has different movements and forces acting upon the prosthetic. The forces that the prosthesis translates to the patient, as well as the forces the patient puts on the prosthesis varies depending on which of the four different types.

Transradial

As outlined previously, transradial prosthesis involves the amputation of the arm below the elbow. In most cases, this allows the patient to still retain the ability for elbow flexion and extension. This ability greatly increases the overall function and abilities of the patient when compared to a patient with a transhumeral amputation. Like all artificial limbs, the prosthesis attaches to the residual limb “stump” through the socket of the prosthesis. The prosthesis is usually held into the residual limb by straps or by suction. The straps counteract the force of gravity acting upon the prosthesis by wrapping around both shoulders and creating a suspension type hold. The suction device counteracts the opposing forces upon the prosthesis by creating a vacuum within the socket. As the patient slides the residual limb into the socket, a one-way valve allows air to escape the socket. This creates negative pressure between the two and the prosthesis is now attached. The most common non-modular transhumeral prosthesis is that of the cable operated. A cable operated or body-powered prosthesis uses a very similar concept to that of a mechanical brake system on a bicycle. With the prosthesis attached to the residual limb, the cable runs up the arm and attaches to both shoulders. The end of the cable on the prosthesis is attached to a clamping type device. By elevating or extending the shoulder, a pulling force is applied to the cable. This applies tension to the cable and causes it to move towards the shoulder. The clamping device attached to the other end of
the cable is then activated by this tension and therefore opens up. When the tension is released, a spring or rubber band then contracts or closes the clamp with opposing force to the cable.

Transhumeral

Transhumeral amputations provide a lot of difficulty for prosthesis function. Since the elbow is removed during amputation, elbow flexion and extension has to be mechanically imitated. This is commonly done through a modular prosthesis. In the case of a non-modular prosthesis, a joint mechanism is put in place to allow for the joint to be adjusted and locked in different positions. This includes complete flexion as well as full elbow extension. The same method of the cable system as outlined for the transradial prosthesis is also applied to the transhumeral prosthesis. See Image 1. The locking elbow mechanism combined with the body-powered clamping systems allows for a combination of tasks and movements to be performed. Along with these two mechanisms, a number of different hand attachments can be applied. These attachments include cosmetic and functional attachments to allow for multiple different activities. Outdoor activities are also possible with attachments to hold sports equipment such as bats and rackets. There are even attachments that allow the patient to grip the handlebar on a bicycle or motorcycle.

Transtibial

Transtibial amputation provides many challenges for movement as well as the forces acting upon the prosthesis. Transtibial amputations occur below the knee and allow the function of knee flexion and extension to still occur. This function is a big deal for the patient because it allows walking to be a much easier task when compared to transfemeral amputations. Like most upper extremity artificial limbs, transtibial prosthesis can either be attached by suspension, suction, or a combination of both. The difference being between upper and lower extremities, are the negative forces applied to artificial leg through standing and walking. This force helps keep the prosthesis in place. But, the alternating force vectors acting in the y-axis while walking can jar the prosthesis loose. When the prosthesis is in contact with the ground, there is a normal force and a gravitational force acting in equal but opposite directions, which holds the prosthesis to the residual limb. Then, when the foot is lifted off the ground, there is a gravitational force and tension force acting in equal and opposite directions that want to pull the prosthetic away from the residual limb. Another major concern associated with all lower extremity prosthetics is the contact point between the prosthesis and the ground. For normal everyday use, lower extremity prosthetics are outfitted with a platform that mimics the size and shape of the foot. This allows for shoes to be worn and gives more of an equal feel between the prosthesis and the opposite leg. It also gives a very stable platform for the prosthesis to provide optimal balance and control during motion.

Transtibial prosthesis can also be made for specific sports and activities just like upper extremity prosthesis. Prosthesis made for running feature a carbon fiber c-shaped spring to allow for shock absorption as well as stored energy for “push off”. See Image 2. The push off phase of running is mostly provided by the ankle and foot and is powered
by the gastrocnemius “calf muscle”. The carbon fiber spring is a simple design that captures the function of all three skeletal-muscle parts involved with this motion. Different shapes and spring rates allow for a variation of forces from different types of running. Snow ski specific prosthesis features fixed knee flexion as well as fixed dorsiflexion of the foot to give optimal balance to the skier. Motocross specific prosthetics feature pneumatic shock absorbers and a pivot point mimic the flexion and extension of the foot. There are also sport specific leg prosthetics for golf, swimming, and bicycling.

Transfemoral

Transfemoral prosthetics are replacement artificial limbs for amputations done above the knee. By losing the knee with a leg amputation, the patient loses the function of the knee along with the function of the ankle and foot. Transfemoral amputees expend 80% more energy to walk than that of a normal functioning individual. This provides many challenges for prosthetists as well as a lot of room for development and innovation. Along with the many difficulties outlined with a transtibial amputation, transfemoral prosthetics has to address the function of the knee. This function includes knee flexion as well as knee extension. The most common non-modular prosthesis that provides this function uses a mechanical knee. The function of the mechanical knee is to provide free motion. With this free motion the knee acts as a pivot point while the weight of the lower leg gives a pendulum action. Using the same concept of normal knee function, as the foot is lifted of the ground during a step, the foot swings forward due to the gravity acting upon the foot. The patient then sets the foot down when the knee is fully extended allowing for a forward stride. This is one of the more difficult functions to perform for any prosthetic user and takes a lot of practice for optimal speed and balance. The knee can also be locked in to place. Similar activity specific prosthesis as outlined with transtibial prosthetics can be applied for transfemoral prosthetics. The more sophisticated prosthetics in this class contain pneumatic shock absorbers at the knee joint for sports such as bicycling, motocross, and skiing.

Externally Powered Prosthetics

Externally powered prosthetics is the present and future of artificial limb design. There are multiple different types of externally powered prosthetics including modular, myoelectric, pneumatic, and hydraulic. A lot of this same technology has been developing in robotics for decades now. With robotic technology, neural-transmission research, and the knowledge of prosthetic functional and body movement, the prosthetic field has hit an all time high for innovation and development.

Modular prosthetics refers to an artificial limb that is externally powered by an electric source and can perform multiple movements at the same time. This idea has lead to more seamless and natural movements among externally powered arms and legs. Through the use of a microprocessor, circuit of electric motors “servos”, and gears, movement of the prosthesis takes place. For instance, in an externally powered modular hand, each finger relies on its own servo and gearbox in order to move independently from the rest of the hand. See Image 3. This allow for very intricate and precise
movement. This microprocessor circuit is powered by direct current battery, which supplies a long duration of use. The circuit can also be turned off by a circuit switch, which cuts off the voltage supply from the battery when the prosthesis is not in use. A variety of different controls are available for these systems depending on the amputation type and the function of the patient. A remote control system can be used in the soles of the feet to control the prosthetic arm or vice versa. Myoelectric control is one of the most innovative technologies used in prosthetics today.

Myoelectric prosthetics uses the natural neurotransmission of the nerve fibers from the residual limb to provide control of the prosthesis. When the brain signals for the use of muscles, a neural electric signal travels through the nervous system to the specific spot of the muscle. This neural signal transmits an electromyograph (EMG) signal to the skin, which can be picked up by electrodes that are attached on the skin. The EMG signal is very similar to that of an electro-magnetic field (EMF), which we use in everyday life at traffic lights and metal detectors. These electrodes send the signal to the microprocessor to move the specific part of the limb desired. Although this type of neural control seems very complicated, it translates to a very natural and fluid movement when it is paired with an advanced modular prosthesis.

Conclusion

Prosthetics is a medical field with a long history that dates back all the way to 1500 BC in Egypt. Some big advancement in prosthetics came during the renaissance period with more functional artificial limbs as well as advancement in amputation procedures. With the rise of the two World Wars came a demand for more functional and comfortable prosthetic limbs. The AOPA was formed and gave a base on which modern prosthetics could be developed. From this base, modern functional prosthetics as well as modular prosthetics was developed to give patients reliable and more useful function. The body-powered prosthesis uses common gravitational force as well as muscle force to give function to the prosthetic. This is done by a cable and clamping system in the upper extremities to give hand function. In the lower extremities, transtibial prosthesis uses normal and gravitational force to provide stability. Spring systems such as the c-shaped carbon fiber prosthesis provide specific athletic function to runners. Transfemoral prosthesis uses gravitational and normal forces that act upon the artificial knee joint to give function at the knee for walking. A pneumatic shock absorber system can be designed to give function for specific sports such as motocross and bicycling. Externally powered prosthetics are the most modern of functional prosthetics. This allows the patient to have functional movement through the use of electrical circuits and a microprocessor. Multiple different types of control systems can be used for these prosthetics, but the most advanced is that of the myoelectric interface. Myoelectric prosthetics uses the electromyography signal given off by neurons at specific points on the muscle. This intern allows for control of the bionic limb with a more seamless natural sense of movement.

I personally feel a lot more integrated by the field of prosthetics after completing this research paper. The development of prosthetics throughout history is very impressive and shows how much technological advancements have happened just in the last 20 years. One of the more interesting parts of prosthetic field is how physics relates
to each aspect and type of artificial limb that has been developed. Gravitational forces and body weight created a demand for the development for stronger and lighter materials. Kinetic motion and kinesiology paved the way functional body-powered prosthetics. The development of circuits, microprocessors, servos, and DC batteries has allowed for a huge advancement in externally powered artificial limbs. This concept combined with myoelectric interfaces allows for an extremely useful prosthesis. After watching and researching myoelectric prosthetics, I feel the need to try out a bionic arm myself. With all this development of artificial limbs, I do feel remorse for the reason for demand. The rise in amputees do to forced trauma and cardiovascular disease is very concerning. Finding new ways to prevent amputation and lower risks of men and women in the military should be the number one concern in this field. Losing an arm or a leg is a very traumatic life-changing event. I can’t even imagine how much my life would change with the loss of a limb. A lot of amputees experience the loss of two limbs, which makes it even more difficult. The bravery of men and women in our military to go out and risks their health and lives for our freedom cannot go unnoticed. Therefore the development of prosthetics to retain a more normal life is very important within this field.

The future of prosthetics is very hard to predict because of how far advanced we are today. The most interesting part of prosthetics today is that we are mimicking the ideas of science fiction movies such as “Terminator” and “i Robot”. I feel that the only place to go with bionic prosthetics is to fully functioning and unnoticeable designs. This will allow the prosthesis to look and feel like a normal arm and leg. It will also have very fluid and natural movements. The other advancement related to myoelectric interfaces are receptors for touch and feel. This is developed by attaching artificial sensor receptors to the patient’s nerves and give feedback similar to actual touch senses. Developments of bionic arms are now being used in paralyzed patients through neural control as well. In the future, this technology could lead to the ability of a paraplegic individual to walk again. Another advancement in the health industry is stem cell research. Although it is a controversial subject among many, there is the possibility for the regrowth of limbs among amputees. Overall, the prosthetics is a very fast growing medical field. With the advancement of medical and technological sciences, would could be seeing some pretty amazing developments in the near future.
Images

Image 1:

Image 2:

Image 3:
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Motion in the Spectral Analysis

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Motion is a huge part in physics, chemistry, biology, geology and many other sciences. This also applies to astronomy, specifically dealing with the Doppler effect and how it’s used not only here on Earth but outside of Earth in space. The Doppler effect is made up of wavelengths and shifts. We determine which shift is what while its shifts along the electromagnetic spectrum. The Electromagnetic Spectrum allows us to determine what a Redshift and a Blueshift are. These types of shifts along the spectrum are compared to sound frequencies here on Earth and work in similar ways.

When it comes to motion in Astronomy we look at the Doppler effect and how it transfers from basic physics to the cosmos. A man named, Christian Andreas Doppler during the year 1842, first introduced the Doppler effect. He was able to set the standard for motion and was able to better our knowledge in motion. The Doppler effect (also known as Doppler shift) is the change in frequency of a wave (or other periodic event) for an observer moving relative to its source. The Doppler result can be noted for each kind of wave - water wave, sound wave, light wave, etc. We are most acquainted alongside the Doppler result because of our experiences alongside sound waves. Perhaps you recall an instance in that a police car or emergency vehicle was voyaging towards you on the highway. As the car approached alongside its siren blasting, the pitch of the siren sound (a compute of the siren's frequency) was high; and next unexpectedly afterward the car bypassed by, the pitch of the siren sound was low.

The Doppler result is of intense attention to astronomers who use the data concerning the shift in frequency of electromagnetic waves produced by advancing stars in our galaxy and beyond in order to derive data concerning those stars and galaxies. The belief that the cosmos is increasing is established in portion on observations of electromagnetic waves emitted by stars in distant galaxies (Cool Cosmos, 2015). Furthermore, specific data concerning stars inside galaxies can be ambitious by request of the Doppler effect. Galaxies are clusters of stars that normally rotate concerning a little center of mass point. Electromagnetic radiation emitted by such stars in a distant galaxy should materialize to be advanced downward in frequency (a.k.a red shift) if the star is rotating in its cluster in a association that is away from the Earth (Cool Cosmos, 2015). On the supplementary hand, there is an upward shift in frequency (a blue shift) of such noted radiation if the star is rotating in an association that is towards the Earth.

The redshift or blueshift of a galaxy is the advancing of its spectral features to longer (or shorter) wavelengths chiefly due to the combination of Doppler gestures and the finished development of the Universe. Extra properly, the word radial velocity is utilized chiefly for the Doppler gestures, that are normally the consequence of gravitational contact, as redshift is kept for the cosmological results, even though it is not usually probable to distinct out cosmological development and Doppler velocities except for adjacent galaxies and those recognized to be associates of galaxy clusters.

The Doppler Redshift results from the comparative gesture of the light emitting object and the observer. If the basis of light is advancing away from you next the wavelength of the light is stretched out, i.e., the light is advanced towards the red. These results, individually shouted the blueshift, and the redshift are jointly recognized as
doppler shifts. There are different types of redshifts as well (Cool Cosmos, 2015). The Cosmological Redshift is a redshift provoked by the development of space. The wavelength of light increases as it traverses the increasing cosmos amid its point of emission and its point of detection by the same number that space has increased across the crossing time. The Gravitational Redshift is a shift in the frequency of a photon to lower power as it ascends out of a gravitational field. Redshifts are extremely beneficial to space exploration and space history.

A redshift is also determined through the electromagnetic spectrum. In the visible portion of the electromagnetic spectrum, red light has the lowest. In the visible serving of the electromagnetic spectrum, blue light has the highest frequency and red light has the lowest (Cool Cosmos, 2015). The word blueshift is utilized after visible light is advanced in the direction of higher frequencies or in the direction of the blue conclude of the spectrum, and the word redshift is utilized after light is advanced in the direction of lower frequencies or in the direction of the red conclude of the spectrum (Zavala, 2014). Today, we can discern light in countless supplementary portions of the electromagnetic spectrum such as wireless, infrared, ultraviolet, X-rays and gamma rays. Though, the words redshift and blueshift are yet utilized to delineate a Doppler shift in each portion of the spectrum. For example, if wireless waves are advanced into the ultraviolet portion of the spectrum, we yet say that the light is redshifted - advanced in the direction of lower frequencies.

The light from most objects in the Cosmos is redshifted as perceived from the Earth. Merely insufficient objects, generally innate objects like planets and a little adjacent stars, are blueshifted (Hirashita, 2014). This is because our Cosmos is expanding. The redshift of an object can be measured by scrutinizing the absorption or emission lines in its spectrum. These sets of lines are exceptional for every single atomic agent and always have the alike spacing. After an object in space moves in the direction of or away from us, the absorption or emission lines will be discovered at disparate wavelengths than whereas they should be if the object was not advancing (relative to us). Astronomers are able to determine how far away distant objects are by measuring this wavelength expansion.

The Cosmological Redshift is probably one of the biggest discoveries in Astronomy. Redshifts and The Big Bang Theory go hand in hand. This type of Redshift is known as the Cosmological Redshift. (the expansion of outer space), like I said previously (Kong, 2015). The wavelength of light increases as it traverses the expanding universe between its point of emission and its point of detection. The cosmological redshift is a redshift provoked by the development of space (Crockett, 2012). As a consequence of the Large Bang (the incredible blast that marked the commencing of our Universe), the Cosmos is increasing and most of the galaxies inside it are advancing away from every single other. Astronomers have discovered that all distant galaxies are advancing away from us and that the farther away they are, the faster they are moving. This recession of galaxies away from us reasons the light from these galaxies to be redshifted. As a consequence of this, at extremely colossal redshifts, far of the ultraviolet and visible light from distant origins is advanced into the infrared portion of the spectrum.
(Fuelner, 2006). This way that infrared studies can give us far data concerning the ultraviolet and visible spectra of extremely youthful, distant galaxies (Zhang, 2015). In 1929, Edwin Hubble announced that almost all galaxies appeared to be moving away from us. In fact, he found that the universe was expanding - with all of the galaxies moving away from each other. This phenomenon was observed as a redshift of a galaxy's spectrum. This redshift appeared to be larger for faint, presumably further, galaxies. Hence, the farther a galaxy, the faster it is receding from Earth.

Currently, the objects alongside the highest recognized redshifts are galaxies and the objects producing gamma beam bursts. The most reliable redshifts are from spectroscopic data, and the highest confirmed spectroscopic redshift of a galaxy is that of “UDFy-38135539” at a redshift of $z = 8.6$, corresponding to just 600 million years afterward the Large Bang. Three teams discovered it in September 2009 in sensitive infrared Hubble Space Telescope images. All teams independently recognized the basis probable a tremendously distant galaxy because there was no measurable light at visible wavelengths (caused by absorption of hydrogen gas alongside the line of sight). Pursuing the invention of this candidate distant galaxy, one more team targeted this object alongside ground-based spectroscopy to confirm the distance, describing a redshift of 8.6. Though, endeavors to replicate this observation powerfully counsel the early claim was in error, meaning that at the present period the galaxy merely has a photometric redshift estimate. The galaxy is placed in the constellation Fornax, and is approximated to have encompassed roughly a billion stars, even though it was merely at most one tenth of the diameter of our own galaxy, the Milky Way, and had less than 1% of the mass of the Milky Way's stars. This was just the example of the highest redshifted galaxy we understand at the moment; it is extremely humiliating invention and an outstanding accomplishment. There were countless supplementary galaxies out there and here are UDF’s older foes. The preceding record was grasped by IOK-1, at a redshift $z = 6.96$, corresponding to just 750 million years afterward the Large Bang. Somewhat less reliable are Lyman-break redshifts, the highest of that is the lensed galaxy A1689-zD1 at a redshift $z = 7.6$ and the subsequent highest being $z = 7.0$. The most distant noted gamma beam erupt was GRB 090423, that had a redshift of $z = 8.2$. Just little cool examples of real life examples!

In conclusion we see that redshifts and blueshifts are extremely important in astronomy discovering, exploration in general. This is all due to the Doppler effect and how it relates from Earth to space. Paired with the electromagnetic spectrum we are given the ability to not only seeing how far away stars are but much more like galaxies, clusters, and even solar systems. The fact that redshifts are a huge tool in showing us how old are our universe is, is essential. Cosmological redshift was revolutionary and also bonded with Hubble’s law is perfect. These findings in the last century to 2 centuries have furthered our advancement in knowledge and space knowledge as well.
Annotated Bibliography: Motion in the Spectral Analysis


Exoplanets

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Introduction to Solar System Astronomy

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In the distant universe too far for us to truly see, there are thousands of planets that orbit two or more stars not including the sun, these are called exoplanets or extrasolar planets. Thousands of exoplanets are consistently being discovered, analyzed, have theories developed around them, are proven, and disproven of their “exoplanet” title. The following information will discuss what an exoplanet is, its characteristics, how they can be found, their structure, the history of them, and other possible theories surrounding the exoplanet.

What makes exoplanets so interesting compared to the main planets in our solar system is that they do not orbit the sun. They can potentially be examined as “super earths” very similar to our own planet but have more unique and differing characteristics. The Jet Propulsion Laboratory managed for NASA by the California Institute for Another Earth used the analogy that exoplanets are growing up with “more than one parent star” (“Four-fathers: New exoplanet discovery part of a quadruple-star system”). Since the first discovery of an exoplanet in 1988, astronomers have come up with theories and found some to be true about exoplanets since then (Stromberg). There are many different sizes of exoplanets from ones that are “bigger than Jupiter to smaller than Earth” (“What are exoplanets?”).

With an incredible amount of discoveries of exoplanets, astronomers have to categorize them into which exoplanets are confirmed and which are candidates. Confirmed exoplanets are planets that the astronomers have proven with confidence from multiple observations and resources. Candidate exoplanets derive from the Kepler mission. Since the German mathematician, Johannes Kepler, stated that thousands of exoplanets have been found they need to be proven using other data which takes a great deal of time. They are possible “candidates” to officially be called “exoplanets” and not “fake positives” that are other cosmic phenomena (“What's the difference between a candidate and a confirmed exoplanet?”).

Astronomers rely on many different ways to find exoplanets since we cannot see most of them with the naked eye. One way is by the Transit method where most often the Kepler space telescope is used to detect the planet passing in between you and its star. It briefly blocks the incoming starlight where dimming occurs. In “enough frequency you might be able to infer the presence of the planet” (Stromberg, 2014). Basically, scientists can measure the brightness of a star for a long period of time in search for when the brightness decreases. This is caused by a planet passing in front of it (“What are exoplanets?”).

Stars that are within the galaxy but smaller than the sun are called “M Dwarfs”. The reason M Dwarfs are important is because if scientists can learn more about them they will further their knowledge about exoplanets. In the Harvard-Smithsonian Center for Astrophysics in Cambridge, Massachusetts they developed the MEarth Project which is an astronomical survey using robotic telescopes to “observe nearby M dwarf stars in search of new Earth-like exoplanets”. By using these telescopes, these scientists can determine how closely a planet with an appropriate temperature to sustain life is orbiting an M dwarf which makes it likely for the planet to pass in front of it (“What are exoplanets?”).

Another option is by Orbital Brightness where an exoplanet “orbiting its star causes the amount of light reaching Earth to rise, rather than dip….it's [then] heated to the degree that it emits detectable amounts of thermal radiation”. Radial Velocity is another main method where if the exoplanet has enough mass it could potentially pull the star toward it. This causes “periodic, predictable yet still minute shifts in the star's position [which] can be used to infer the presence of a large planet near that star” (Stromberg, 2014). Evidence shows that drawing an exoplanet closer to certain stars could be influenced by a change in the planets' orbits. An example is the "hot Jupiters" which are “planets around the mass of Jupiter that whip closely around their stars..."
in just days... might be gently nudged closer to their primary parent star by the gravitational hand of a stellar companion” (“Four-fathers: New exoplanet discovery part of a quadruple-star system”).

Rarely astronomers can simply see the exoplanets with Direct Imaging. The astronomer would need to detect that the planet is far enough away from its star but also if it is too far away then the exoplanet won’t be able to “reflect enough of the stars light to be visible”. The final method is Gravitational Lensing where very rarely “astronomers have been able to infer the presence of distant planets by the way that they magnify the light of even more distant stars” (Stomberg, 2014).

Scientists are not only searching for the exoplanets themselves but once they find them they intend to learn more about their interiors and what the individual exoplanets are made of as well. What is most difficult about determining an exoplanets interior composition is that only their mass and radius can be found with no other information even regarding the interior. Scientists hope to learn from more exoplanets they find to compare the data and come to a few more simple conclusions (Seager, “Sara Seager Planetary Scientist/Astrophysicist: Research”).

Exoplanets are constantly being discovered almost every day. But determining whether they are confirmed or candidates as stated previously takes time. Looking recently at April 20, 2015, there were exactly 1,912 planets, 1,208 planetary systems, and 480 multiple planetary systems, according to the official Exoplanet catalogue. This catalogue lists the thousands of exoplanets discovered with their names, mass, radius, period (day), AU, e, degree, distance, arsec, year of discovery, and the date the knowledge about them was later updated (Zolotukhin, 2015).

NASA has developed a “Exoplanet Exploration Program” (ExEP) over the past few years to understand planetary systems around nearby stars. This means that the ExEP looks at the project with long-term intentions to “chart out a strategic timeline of missions and instruments”. First the program must determine how many and what different kinds of these other planetary systems in the universe exist. The ExEP uses albedo, infrared, and spectroscopy measurements that has scientists “make temperature maps and calculate general atmospheric compositions... [so they] will be able to record increasingly detailed and accurate data and pictures, a feat that would have been unthinkable just a few decades ago” (Rodriquez).

In March 2013, astrologists discovered some of the gases that make up the atmosphere of four exoplanets in a nearby galaxy orbiting one star. In order to find such exoplanets, planet hunters need to study light from its host star. The Astronomers were able to determine and measure “the gases surrounding the exoplanets by looking at how they affect light that leaves them”. This exoplanet is known as “HR 8799c” and was studied by Quinn Konopacky of the University of Toronto, Canada, and her team. The infrared light being emitted from HR 8799c was detected showing that the exoplanet’s atmosphere contains not only water but carbon monoxide. Because the exoplanet is about 130 light-years away it would take 130 years to reach Earth (Ornes, 2013).

An exoplanet in January 2015 was discovered by a Dutch-U.S. team which was said to have a ring system 200 times larger than Saturn’s rings. “Researchers say there are probably more than 30 rings, each measuring tens of millions of kilometres in diameter”, the BBC article stated. In photoshopped photos of our day and night sky, professionals placed the extremely large rings of the exoplanet around where Saturn is in the sky and they was clearly visible. If the astrologists are correct of it’s existence, the exoplanet would be the first to obtain such a structure around our planet that is beyond our solar system. "You could think of it as kind of a
super Saturn,” said Prof Eric Mamajek, from the University of Rochester in the U.S. Astronomers first saw a complex series of eclipses that were lasting up to 56 days, which is quite abnormal. The astrologists think this could be caused by this planet “with a giant ring system blocking out light as it passes in front of the star J1407” (Rincon, 2015).

In 2004, an exoplanet with “super-earth qualities” being twice as large in diameter and over eight times more massive than Earth was discovered. The planet’s name is 55 Cancri e and is 40 light-years away. 55 Cancri e was found with uninhabitable living conditions of 3,900 degrees Fahrenheit combined with a high amount of carbon creating a unique place to find large amounts of diamonds. (Fazekas, 2012). Researchers estimated that at least a third of the exoplanet’s mass, equally the mass of three Earth’s, could be diamonds. The difference between these possible diamond encrusted exoplanets and regular planets is that their planetary system is made up of mostly carbon and a planet like Earth’s atmosphere is mostly oxygen and silicates. This planet would be made of mostly graphite and diamond rather than like Earth with granite and water.

In 2012, the exoplanet was found to be the “first Earth-sized exoplanet whose light was directly observed via the infrared capabilities of NASA’s Spitzer Space Telescope” (Major, 2012). “55 Cancri e’s” host star, “55 Cancri”, is one of the first known stars to host an extrasolar planet and it is the first time a ‘smaller’ exoplanet’s light has been detected directly” (Atkinson, 2012).

An illustration of 55 Cancri e shows a surface of mostly graphite surrounding a thick layer of diamond. Illustration courtesy Haven Giguere, Yale.

According to planet hunter, astrophysicist and planetary scientist at MIT, Sarah Seager, believes that further knowledge and development with exoplanets can spark the potential of finding other life in the universe. The first planet found orbiting a star other than the sun, exoplanets, was about 20 years ago. Since then about 1,000 exoplanets are known to exist and are being recorded on their activities every day. Most “candidates” are still being analyzed which means that astrologists believe there are many more Earth-size planets out there in habitable zones of small stars. A “habitable zone is a region around a star where a planet can have surface temperatures consistent with the presence of liquid water”. As we already know, water is required for all life so searching for a planet inhabiting water as a liquid could very well be the next “super Earth”. A common mantra in astrobiology is to “follow the water”, meaning to find planets that have water on them too. But when it comes to a certain habitable zone, “there is no universal habitable zone applicable to all exoplanets”. In a perfect universe, there would be which would make “super Earth” hunting much easier (Seager, “Exoplanet Habitability”, 577-580).

Exoplanets are one of the rare aspects of the universe that are usually too far for astrologists to touch but with our advancing technology, close enough to study them.
Astrologists are actually able to speculate and determine information regarding their host star that they orbit, if they inhabit water, how large or small they are, and their core interior. I never knew that there were so many other planets we actually have found that are not in our own universe. Exoplanets are a very fascinating mystery that we will continue to explore leading us to many other realizations about the universe.
Sources Cited


This website article was useful because it gave me information about diamond planets. This was a specific topic I wanted to cover and not much is known about them yet. Diamond planets are seen as mainly a number of theories but this article used facts and many interesting reasons and statements.


This journal article found on the PVCC library database explains the “productive... methods to probe exoplanet atmospheres”. This is a useful source in that the writer also explains the results obtained and suggest broad topics where work could be furthered. I need at few more database sources so this is a very reliable one.


This magazine article was especially reliable since it was from one of the best nature and culture professional magazines, the National Geographic. Being one of the only other sources I found on exoplanets, it explained a specific one found. The article not only analyzed how the planet was found by its interesting interior chemistry.


This article was especially useful in that it was the first one I found. Not knowing what an exoplanet is the article went into depth of a quadruple star system which was easier to learn about that an exoplanet. So from this knowledge I was able to comprehend easier how an exoplanet was found and what they have to do with multiple stars.


This peer reviewed journal article will be useful for my research in the study of the properties of exoplanets. This is a reliable source because the authors are university professors, physicist, and other types of researchers. The authors state in the article the multiple sources used to discover exoplanets and the history of them which would be an interesting background.


This article interested me because it was the most useful one I found regarding exoplanets containing a diamond interior. There really was not a lot of information about diamond planets
probably because they are so unknown. But this article described their importance as exoplanets, rare composition, and certain exoplanets thought to contain the diamond interior.


This main website was useful in a few different ways. But for this article it described the overall aspect of exoplanets describing light travel, what makes an exoplanet different from regular planets, and a few examples of exoplanets. I could use this to describe the basis in simple terms of what an exoplanet is and what goes into discovering one.


This recent newspaper article from the BBC online website was published on January 27, 2015. It explained that an exoplanet was just found to have a giant ring system “200 times larger than that around the sun”. This is relevant because it is a very recent and easy to understand about an interesting and different exoplanet.


This article was found on NASA’s website specifically on exoplanets. It was useful for my research because the website gave background information on how astronomers have explored the extrasolar planets. I used the example of the program and what they do in the further researching of these planets.


The peer-reviewed journal article I found on PVCC’s library database will be useful to my research and paper. Seager’s article promises in the future more knowledge on the search for exoplanets and all the mysteries behind them. An important point made is that some of these planets might even be sustainable enough for life like Earth so this would be a breakthrough in discovery.


This website was created by Professor Sara Seager, a planetary scientist and astrophysicist at Massachusetts Institution of Technology. She is a known planet hunter and does a lot of her work solely on exoplanets. The website has her biography, list of books she has written, research on exoplanets with their description and atmospheres, and other published work of hers on this topic. This will help me understand the basic premise of what exoplanets are and what they are like to start out in my research.


This magazine article I probably used the most throughout my research paper. This is because it discussed certain topics like finding exoplanets, who finds them, and the different methods used in doing so. I discussed this a few different times in my article and the source was very reliable being an article from the Smithsonian Institute.

This was a useful article on my topic because it explained a way scientists and astrologists can learn even more about exoplanets. It had a set of important vocabulary words at the bottom of the article helping the reader. There was also an extensive list of suggest reading which could possibly help in my research with exoplanets.


On this website I found more information simplified as to what an exoplanet is and how they can be found. I also came across a team at Harvard University, which is where the site is from, that explains finding exoplanets through their MEarth Project.


This was one of the first sources I found for my topic. This website from NASA introduced me to the types of exoplanets in which ones are confirmed as being exoplanets and which are still being questioned. Most of the knowledge of exoplanets is made up from theories and slight observations.


This catalog was found on an educational website dedicated to exoplanets. This specific page is a catalogue of all the exoplanets discovered going back to 2005 and stops at 2015. It lists the exoplanets’ names, mass, radius, period, AU, e, degree, arsec, status, and year discovered. This could be useful in mentioning how many there are and the most recent one discovered with how common they are.
Telescopes in Orbit
Cory J. Gomes
Paradise Valley Community College
Introduction to Solar System Astronomy
2015
Professor J. Weitz
This paper will cover various space telescopes that are currently in use. Each telescope has a unique set of instruments designed to fulfill its unique purpose. These instruments give scientists the ability to detect the different wavelengths of the electromagnetic spectrum and allow them to see what would otherwise be invisible to the naked eye. I will go over some basic facts about each telescope and briefly cover each telescopes collection capabilities and equipment.

The Spitzer Space Telescope was deployed on August 25th, 2003 and is used to detect the infrared bandwidth. Manufactured by Ball Aerospace and Technologies and Lockheed Martin the project cost a total of 720 million USD. This space telescope orbits the Earth at a height of 353 miles and is used primarily to detect distant and nearby objects. The Spitzer Space Telescope has three main instruments, these are the Infrared Array Camera (IRAC), Infrared Spectrograph (IRS), and the Multiband Imaging Photometer for Spitzer (MIPS). Each of these instruments allow scientists to see well into the world of the infrared and various levels of its spectrum.

The Hubble Space Telescope is one that is much more commonly known and has been orbiting the Earth for 25 years now. This telescope was manufactured by Perkin-Elmer and Lockheed and cost a total of 2.5 billion USD. Launched on April 24th, 1990 the telescope orbits the Earth at 347 miles high and studies deep space object. The Hubble Space Telescope has equipment that gives it the capability to detect visible light as well as ultraviolet and near-infrared radiation. This space telescope has five main instruments that serve various purposes. The ACS or Advanced Camera for Surveys is the Hubble Space Telescope’s third generation imaging camera. The ACS is one of Hubble’s primary instruments and used to perform surveys and broad imaging campaigns. The Cosmic Origins Spectrograph (COS) is a UV sensor on Hubble, it can operate on two channels allowing for far-UV and near-UV data collection. NICMOS is another tool used by this space telescope, its acronym stands for Near IR Camera/Multi-Object Spectrometer. NICMOS is Hubble’s only instrument that allows for near-infrared data collection. Another instrument utilized by Hubble is the Space Telescope Imaging Spectrograph or STIS. This Instrument is used to obtain high resolution spectra of Hubble’s resolved objects. Finally, the last instrument is the Wide Field Camera 3 (WFC3), this system allows the wide field images produced by Hubble’s various data collection sensors.

A more recently developed space telescope is the Kepler Space Telescope. This telescope was deployed on March 7th, 2009 and was manufactured by Ball Aerospace & Technologies Corporation. One interesting aspect of the Kepler Space Telescope is that it does not orbit our planet like most satellites that we know, it in fact orbits our Sun with us and has an orbital period of 372.5 days. This 600 million USD project is used by detecting visible light to find and study exosolar planets. This telescope is composed primarily of two elements: the Photometer and the Primary mirror. The Photometer is an array of 42 CCDs (charge coupled devices). Each 50x25 mm CCD has 2200 x 1024 pixels. The CCDs are actually not used to take pictures, the images collected are intentionally defocused to improve the photometric precision. As for the Primary mirror it is 4.6 ft in diameter, and the largest mirror located outside Earth orbit. Using ultra-low expansion (ULE) glass, the mirror is specifically designed to have a mass only of only 14% of a mirror of the same size. Kepler utilizes these instruments to study extrasolar planets through what is known as the Transit Method. Transits by terrestrial planets produce a small change in a star's brightness of about 1/10,000, lasting for 2 to 16 hours. This change must be absolutely periodic if it is caused by a planet. In addition, all transits produced by the same planet must be
of the same change in brightness and last the same amount of time, thus providing a highly repeatable signal and robust detection method.

Though the Kepler Space telescope has only been deployed for a relatively short amount of time it has had quite the eventful journey. In July 2012, one of Kepler’s four reaction wheels (wheel 2) failed. On May 11, 2013, a second wheel (wheel 4) failed, threatening the continuation of the mission, as three wheels are necessary for its planet hunting. Scientists and engineers were able to resolve the issue by devising a way to keep Kepler’s stability by keeping balance between radiation pressure from our sun, active force from Kepler’s thrusters, and utilizing the two remaining reaction wheels.

A major contributor of x-ray images of our universe is the Chandra X-ray Observatory. This space telescope was manufactured by Northrop Grumman and TRW inc and cost 1.65 billion USD. It was deployed on July 23rd, 1999 and orbits out to a whopping 65,438 miles above Earth to avoid losing any x-ray radiation through Earth’s atmosphere. On board Chandra has four major instruments, these instruments include the High Resolution Camera (HRC), the Chandra Advanced CCD Imaging Spectrometer (ACIS), the High Energy Transmission Grating Spectrometer (HETGS) and the Low Energy Transmission Grating Spectrometer (LETGS). The incoming X-rays collected by Chandra are focused by mirrors to a tiny spot about half as wide as a human hair on the focal plane, over a distance of about 30 feet. The HRC is one of two instruments used at the focus of Chandra, where it detects x-rays reflected from the assembly of eight mirrors. The ACIS is an array of CCD's, it can make x-ray images, and at the same time, measure the energy of each incoming X-ray. The LETGS and HETGS are two instruments aboard Chandra dedicated to high resolution spectroscopy. Each spectrometer is activated by swinging an assembly into position behind the mirrors to collect data.

For the gamma-ray range of the electromagnetic spectrum of our universe there is no better tool used by astronomers than the Fermi Gamma-Ray Space Telescope (FGST). This space telescope was manufactured by General Dynamics and cost 500 million USD. Deployed on June 11th, 2008 it orbits the Earth at a height of 340 miles. Gamma rays are the highest end of the electromagnetic spectrum, having the shortest wavelength and the most energy. This type if radiation is emitted from the hottest and most energetic objects of our universe. To detect gamma rays the Fermi Gamma-Ray Telescope has two main instruments it utilizes, the Large Area Telescope (LAT) and the Gamma-ray Burst Monitor (GBM). The LAT is FGST's main instrument, within its 1.8-meter cube housing, the LAT uses 880,000 silicon strips to detect high-energy gamma rays. Much like it’s name implies, the Gamma-ray Burst Monitor detects gamma-ray bursts and other transient phenomena. Together with the LAT, the GBM enables FGST to make gamma-ray burst observations spanning a factor of ten million in energy.

So whats in the works for the future of space telescopes? Well the James Webb Space Telescope is the planned successor instrument to the Hubble Space Telescope and the Spitzer Space Telescope. It is scheduled to launch in October 2018 with the program capped at $8 billion USD. The the James Webb Space Telescope will be located near the second Lagrange point of the Earth-Sun system, 930,000 mi from Earth. The primary mirror is a 6.5-meter-diameter gold-coated reflector with a collecting area of 25 m² composed of 18 hexagonal segments. This configuration will allow the James Webb to offer unprecedented resolution and sensitivity from long-wavelength visible to the mid-infrared.


Deep Brain Stimulation - Mechanism, Use and Treatment

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PHY 112
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ABSTRACT
This paper reviews deep brain stimulation (DBS), its mechanism and as an effective treatment for neurological disorders like Parkinson’s disease, dystonia and essential tremor. The effectiveness of DBS is thought to be due to the blocking of pathological electrical signals coming from the brain by the two DBS electrodes implanted in the thalamus region of the brain. Components of DBS include the electrical lead, the wire extension and the implantable pulse generator, which acts on the brain like a pacemaker acts on the heart. Though the exact mechanism of how DBS works is unknown, the future of DBS is promising, especially as neurological specialists gain expanded understanding of how electric stimulation affects various parts of the brain. Future research is needed to understand the mechanisms for deep brain stimulation.

DISCUSSION
Deep brain stimulation (DBS), high-frequency electrostimulation in the brain, has been successfully used for treating Parkinson’s Disease (PD) and movement disorders, such as essential tremor and dystonia, with widespread use beginning in the 1990s, and it is also being explored alternatively to standard of care treatment for behavioral and psychiatric disorders such as obsessive-compulsive disorder. DBS has been helpful in reducing dyskinesias, or involuntary muscle movements, including tremor, as well as increased and decreased body movements, known as hyperkinesias and hypokinesias, respectively. The exact mechanism for why DBS treatment is successful continues to be researched. Structure and electricity of the brain will be reviewed so that components, mechanism and use of DBS can be explored. Previously brain lesions were surgically ablated to destroy brain tissue that was creating the movement disorders; however this is an irreversible procedure. DBS, on the other hand, is seen as an adjustable, reversible procedure with significant benefit to quality of life and activities of daily living. DBS research has also revealed adverse psychological, neurocognitive and surgical consequences.

Important in understanding deep brain stimulation is an understanding of some of the structures of the brain and anatomical target(s) of DBS. In a brain without a movement disorder important structures for conscious and subconscious control of movement include: the premotor and primary motor cortexes/areas of the cerebrum, thalamus, subthalamus, and the basal ganglia and associated structures.

The premotor cortex of the cerebrum is a memory bank and initiator of complex, sequential learned activities. The premotor cortex allows a person to write his/her name, for example, by sending impulses to specific muscles in sequence. The primary motor area, deals with complex, skilled and fine movements, like those of the hands. The thalamus consists of multiple nuclei and is the “relay station” for sensory signals from the spinal cord and brain stem.

The basal ganglia is composed of three bodies, or nuclei: the globus pallidus, putamen and the caudate nucleus. The basal ganglia receives signals from cerebral cortex and sends signals back out to the motor areas of the cortex by way of the neural pathways through the thalamus and is functionally associated to the substantia nigra and the subthalamic nucleus. The primary and premotor areas provide the primary input to the basal ganglia which send out excitatory signals to the striatum, which also receives signals from the dopamine producing neurons of the substantia nigra. Neurons that run from the substantia nigra to the basal ganglia release...
dopamine to control subconscious muscle activities, and disordered signals from these dopaminergic cells lead to neurologic and movement disorders\textsuperscript{8,9}. The globus pallidus externa is responsible for sending signals to inhibit movement to the subthalamic nucleus\textsuperscript{9}. The globus pallidus interna in turn sends signals to the ventral anterior and ventral lateral nuclei of the thalamus with direct and indirect dopamine pathways\textsuperscript{k}.

Understanding of the function of the basal ganglia and how the region interacts with the thalamus has steadily evolved as it has been studied further. Originally, basal ganglia circuitry was thought to control movement exclusively, however the understanding of these structures now show that the basal ganglia have high level behavioral control as well\textsuperscript{10}.

The subthalamus contains the subthalamic nucleus (STN), the red nucleus and the substantia nigra\textsuperscript{9}. As research is continuing to uncover more detailed connection of the STN to conditions like tremor and hypokinesia, studies have detailed clearly that the STN is excitatory in its effect and has come to be known as the “clock of the basal ganglia,” and has had definite effects on behavior\textsuperscript{5}. The STN is located centrally in the basal ganglia and was previously thought to be a relay station for the basal ganglia, but now is understood to be one of the main regulators of motor function\textsuperscript{5}. The STN is now known as an excitatory nucleus for brain signals, instead of inhibitory nucleus as previously understood\textsuperscript{5}.

Deep brain stimulation is a surgical procedure that consists of implanting a thin, insulated wire electrode, the DBS lead, into the brain. The lead consists of four thin, coiled wires that have a 1.5mm electrode on each end\textsuperscript{11d}. The surgeon will make sure to place the tip of the electrode in the specific area of the brain s/he wants to target\textsuperscript{7}, with the goal that the electricity from the electrodes block signals initiated in the brain causing abnormal movements. Another insulated wire, called the extension, runs under the skin to the neurotransmitter implanted near the clavicle, which is the device, almost like a pacemaker, that delivers the electric impulses to the lead and electrodes within the brain\textsuperscript{11d}.

Components of the deep brain stimulation system include the lead, the extension and the implantable pulse generator\textsuperscript{7d}. A thin, insulated wire electrode, called the lead, is the component that delivers the electrical signal directly to the brain that the surgeon wants to target. The surgeon will insert the lead through a small opening in the skull. Areas of the brain usually targeted include the thalamus, subthalamic nucleus and globus pallidus\textsuperscript{6e}.

From the electrode, a second thin, insulated wire – the extension – is passed under the skin of the head, neck and shoulder connected to the part of the device that generates the electrical impulses, called the implantable pulse generator\textsuperscript{7d}.

The IPG is a battery operated pacemaker and provides electrical impulses that travel down the extension to the lead to block nerve signals causing unwanted movement or symptoms\textsuperscript{7d}. Surgeons use various imaging techniques, including stereotactic MRI, ventriculography and microelectrode recording to target the implantation point\textsuperscript{15}. Electrostimulation settings were adjusted according to the effects for each patient with the standard pulse setting to be 60 x10\textsuperscript{6}sec at 130Hz and voltage adjusted to each patient’s stimulation\textsuperscript{15}. In Deusch’s study of effects of neurostimulation vs. medical management on quality of life for Parkinson’s patients\textsuperscript{17}, the lead
was implanted in the traditional subthalamic nucleus location and anatomically coordinated with appropriate imaging techniques to be located 0-3mm behind the midcommissural point, 4-6mm below the intercommissural line and 11-13mm lateral to the midplane of the third ventricle\textsuperscript{17,e}.

The brain contains 100 billion neurons, which translate a stimulus into an electric signal, called an action potential, which are then conducted to other neurons or muscles\textsuperscript{12,13}. Action potentials travel due to changing concentrations of ions inside and outside of the neuron cell membrane, such as Na\textsuperscript{+}, Ca\textsuperscript{+} and K\textsuperscript{+} and negative anions, such as ATP and large proteins\textsuperscript{13}. Neurons have a negative resting potential of -70mV, with the inside of the plasma membrane more negative than the outside\textsuperscript{13f}. With stimulus, in order to generate an action potential or electrical signal, the neurons must charge to the threshold of -55mV\textsuperscript{13,g}.

The neural system of the brain is an electrical and chemical circuit, where the neurons can be compared to capacitors that need charging\textsuperscript{14}. Simplifying the brain’s circuits, like an AC or DC circuit, for example, the membrane of the neuron has a capacitance (C\textsubscript{m}) and the sodium and potassium ion gradients create currents (I\textsubscript{Na} and I\textsubscript{K}), and can be labeled and used in developing equations, accordingly\textsuperscript{14}. Kang used a schematic diagram\textsuperscript{h} and assumed that activating neural fibers delivering signals away from the motor cortex, to the STN was the dominant mechanism of success in DBS\textsuperscript{14}. Equations with variables similar to those in Ohm’s Law equations used in electrical circuits, were used to create a mathematical models that can aid in clarifying the mechanisms by which DBS works in the STN and basal ganglia areas\textsuperscript{14}. For example, equations to calculate potential in the interneurons, cortical neurons, STN neurons and globus pallidus externa (GPe) neurons were developed\textsuperscript{14}, see Figure 9 for the cortical neuron example\textsuperscript{i}.

The difference between an AC or DC circuit and the electricity of the brain is that the brain relies on electrical and chemical mechanisms to send signals\textsuperscript{13}. Unlike a capacitor in a circuit, neurons in the brain are always polarized have an internal potential, external potential and a potential at rest\textsuperscript{13} represented by V\textsubscript{i}, V\textsubscript{e} and V\textsubscript{rest}, that must be considered in the equations\textsuperscript{14,i}. When one neuron fires due to stimulus, it fires and then depolarizes, which stimulates the chemical mechanism to open ion channels\textsuperscript{13}, for example, the sodium (Na\textsuperscript{+}) ion channel\textsuperscript{j}. It is the change in ion gradient that creates the current in the body\textsuperscript{13}.

The scientific understanding of how the electrical signals spread in the brain with DBS, or its mechanism of action, is still unclear\textsuperscript{3,4,5}. There are some general hypotheses, including: depolarization blockage, synaptic inhibition, synaptic depression, and stimulation-induced modulation of pathologic network activity\textsuperscript{2}.

The first hypothesis for mechanism of DBS electrode placement is that electrostimulation from the electrodes reduces neuron activity by blocking depolarizations across the neurons\textsuperscript{2,5}. The electricity from the electrodes suppress voltage-gated Na\textsuperscript{+} as well as Ca\textsuperscript{+} currents, which hinders activity in the neuron\textsuperscript{5}.

With a disease like Parkinson’s Disease (PD) involving disordered movement, neural activity is synchronized in the basal ganglia network and DBS signals are thought to suppresses this abnormal activity by adding electric stimulation at the nucleus against this neural activity which to suppress excess movement\textsuperscript{5,16}. To further confirm the blockage hypothesis, oscillating neural
activity in the STN of monkeys was seen at a frequency range of 13-30Hz, as well as the globus pallidus interna (GPi) of Parkinson’s disease patients, and when DBS treatment was applied to the STN, this oscillation diminished nearly completely\textsuperscript{14}.

Conversely, an increase in electrical signals at the nucleus being stimulated with DBS have been recorded\textsuperscript{5} and movement stimulated when a patient was experiencing rigidity or hypo/bradykinesia\textsuperscript{16}. An effect of both stimulation and suppression has also been recorded at high frequencies of DBS\textsuperscript{5}, which confirms why the mechanism still needs to be studied rigorously.

The subthalamic nucleus (STN) is often targeted for electrostimulation in the brain in patients with disordered movement\textsuperscript{17}. The latter hypothesis suggests that when DBS is applied in subthalamic nucleus (STN), it generates excitatory effects on neurons around the DBS electrode, and this high frequency stimulation overrides the neural activity that creates the unwanted movements\textsuperscript{5}. The volume of tissue stimulated has been studied and, in their 2009 study, Maks, et. al. discovered that it was the most beneficial to have electrodes positioned near the border of the subthalamic nucleus so that the volume of tissue activated by the electrostimulation would extend outside of the STN\textsuperscript{3}. Those patients who did not cite improvement were found to have the majority of the volume of tissue activated stay within the borders of the STN\textsuperscript{3}.

Likewise, deep brain stimulation of the ventral lateral nucleus (VL) of the thalamus\textsuperscript{c,k} was shown to eliminate tremors, and further understanding of the circuitry of the basal ganglia has continued to narrow the focus of the surgery\textsuperscript{6}.

According to P. Silberstein\textsuperscript{18}, “DBS may be employed in the management of medication-refractory tremor or treatment-related motor complications, and may benefit between 4.5% and 20% of patients at some stage of their disease course.” When management of Parkinson’s disease by medication and physical therapy is no longer effective, long term deep brain stimulation treatment has also been shown to have lasting effects, including effects on the brain’s ability to adapt and change its structure and function based on treatment and experiences\textsuperscript{18}. In dystonia patients for whom DBS was successful, physiological status was measured when DBS was turned off for two days and no change was measured\textsuperscript{19}. This result suggests that DBS may actually reorganize neurons due to long-term treatment\textsuperscript{19}. Deep brain stimulation has not been studied against the effectiveness of medical therapy (medication, and physical/occupational/speech therapy) in 99% of trials\textsuperscript{1}, leaving effectiveness to be observed on a case by case basis.

In the first of its kind trial with advanced Parkinson’s disease (PD) patients, DBS was compared to medical therapy, and was found to be effective in 71% of cases, vs 32% of medical therapy cases\textsuperscript{1}. Medical therapy used in this double blind, randomized study included, medications and non-pharma therapy which included, speech, occupational and physical therapy. DBS, however, also showed an increase in consequences as compared to medical therapy, including reduction in information processing abilities, psychological consequences, surgical adverse events and cerebral hemorrhage (1 case in 121 patients)\textsuperscript{1}.
Many DBS studies only focus on changes in motor function, however this does not take into account other quality of life measurements that may be impacted with treatment, e.g. cognition, mood/depression, behavior, complications from surgery or medication\textsuperscript{17}. Depression, for example, overshadows motor benefits of treatment and has been found to increase for some patients undergoing deep brain stimulation\textsuperscript{5,17}. Gunther’s unblinded study\textsuperscript{17}, focused on quality of life measures and compared DBS treatment to medical management in patients under 75 years old with advanced PD and severe fluctuations in mobility and dyskinesia. Results indicated that DBS was “superior to medical treatment alone\textsuperscript{17}.” The effects of DBS electrostimulation on motor function was also measured, as it relates to clinical measures and in conjunction with quality of life measures and patient diaries, revealed that the length and severity of periods of immobility and dyskinesias was decreased in DBS patients\textsuperscript{17}. These same patients did not rate any side effects from surgery or electrostimulation as decreasing their quality of life\textsuperscript{17}. This is an important distinction because, instead of only reporting the presence of side effects from surgery or electrostimulation in isolation, results were tied to the subjects’ experience of quality of life, and indicated what the subject portrayed as a true benefit of DBS in their life.

CONCLUSIONS

Research data from a physics standpoint, was extremely difficult to find on deep brain stimulation due to scientist’s lack of understanding of the mechanism and why DBS is an effective treatment. I was genuinely interested in understanding the physics of the electricity of the brain and the mechanism and was surprised again and again as each article researched continued to have multiple sentences directed at stating that the mechanism of DBS was still unknown, needed further researched, or that the mechanism needed further research to clarify why results were happening. The articles would most often then move into comparing outcomes of DBS treatment in various kinds of patients with various symptoms or conditions. Understandably, the medical community is interested in outcomes, and my assumption would be that the grant money researchers receive is from industry or government that is interested in outcomes of DBS treatment, however the lack of understanding our scientific and medical community appears to have as it relates to the mechanism of DBS and the mechanism of which parts of the brain actually impact behavior movement, is astounding. This lack of understanding made writing a physics paper on the subject next to impossible, yet through the many weeks of researching, I remained optimistic that there would be articles with the mechanism or possible mechanism of DBS. This was not the case.

I could only begin to wrap my mind around the possible mechanisms of DBS presented in some of the research when thinking about neural membranes having a capacitance and the ion gradients as creating current, as well as the myelin sheath around the axon of the nerve as increasing resistance so the current does not leave the axon. This reminded me of studying circuits in PHY 112 and this understanding allowed me to use a crude comparison of the nervous system to an AC circuit with resistor and capacitor, where the voltage leads the current. In my comparison, once the capacitors were “charged” to the threshold of -55mV, then a current actually flow down the axon in the form of an action potential.

In the end, the research left me excited for neuroscience and those who are currently researching in the field. It appears that there is an “unknown frontier” in this studying how and why the brain
is able to do what it does. I am also curious as to what else we don’t understand about the functioning of the human body.

Overall, I believe the future of deep brain stimulation is promising, especially as neurological specialists gain expanded understanding of how electric stimulation affects various parts of the brain. Continued research and understanding of the brain and the effects of electrostimulation on the functioning of structures like the basal ganglia, STN, GP and related structures will continue to play a critical role in designing more advanced treatment and DBS systems. As was seen in long-term DBS effects in patients with dystonia, DBS had the potential to be tailored to the individual to increase the brain’s ability to adapt and modify its functioning to retain and build on DBS benefits. The possible influence of DBS on brain plasticity is extremely exciting and has potential to positively impact treatment and quality of life for patients in the future.

Additionally, as DBS is used for treating conditions over the long-term, understanding long-term physical, emotional and quality of life consequences is important to research, so that benefits to the lives of people are maximized.
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bFigure 2 Basal ganglia and associated structures. Available from: http://en.wikipedia.org/wiki/Basal_ganglia_disease

cFigure 3 Anatomy of the thalamus. Rughani, AI. Available from http://emedicine.medscape.com/article/189830-overview#aw2aab6b3
**Figure 4** Components of Deep Brain Stimulation. Brown University. http://biomed.brown.edu/Courses/BI108/BI108_2008_Groups/group07/Parkinsons.html

**Figure 5** Sagittal section of the brain, 12 mm lateral to midline, demonstrating subthalamic nucleus (STN; lavender). *Tagliati, M. Available from http://emedicine.medscape.com/article/1153743-overview.*

**Figure 8** Deep brain stimulation schematic diagram of connections between STN, GPe and applied DBS. “+” signs indicate excitatory pathways while “−” signs represent inhibitory pathways. AIS: axon initial segment, GPe: globus pallidus externa. Kang. Available from http://www.ncbi.nlm.nih.gov/pmc/articles/PMC3958751/figure/F1/  

\[ \frac{C_{mi} V_{ni}}{dt} = G_{ai} (V_{n} - 1_i - 2V_{ni} + V_{e,n} - 1_i - 2V_{e,ni} + V_{e,n} + 1_i) - I_{c,ni} \]

**Figure 9** Cortical neuron equation to calculate applied DBS current, conductance and extracellular potential. \( C_{mi} = \) membrane capacitance, \( G_{ai} = \) axon conductance, \( V_{e,ni} = \) extracellular potential of the \( n \)th compartment and \( I_{c,ni} = \) ionic current passing through the \( n \)th compartment. \( V_{ni} = \) reduced membrane; \( V_{i,n} = \) internal potential and \( V_{rest} = \) internal resting potential; therefore \( V_{ni} = V_{i,n} - V_{e,n} + V_{rest} \). Kang. Available from http://www.ncbi.nlm.nih.gov/pmc/articles/PMC3958751/#B16
Figure 10  Model of voltage gated sodium channels. Eijkelkamp, N. Available from http://brain.oxfordjournals.org/content/135/9/2585
**Figure 11** Schematic diagram of basal ganglia circuitry. Red arrows indicate inhibitory signals and green arrow indicate excitatory signals between motor cortex, putamen, globus pallidus externa (GPe) and globus pallidus interna (GPi), subthalamic nucleus (STN), substantia nigra reticulata (SNr) and substantia nigra compacta (SNC), and ventrolateral thalamus (VL) are represented. D1 and D2 indicate direct and indirect pathways, respectively. 6Tagliati, M.
Seeing the Light:
The History of Lasers and Their Impact in Dermatology

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ABSTRACT

This paper traces the invention of the laser and its use in the medical field, particularly the use of carbon dioxide lasers in dermatology. To understand how lasers actually work, the process of electrons in various energy states are discussed. The first laser and gas laser will be introduced, including their introduction in surgical and other medical procedures. The use of gas lasers in resurfacing treatments is explored as well as a prediction for what their future holds.

INTRODUCTION

The advancement of laser technology through the past 55 years is astonishing. From their various uses to the different fields they have been used in, lasers are one of the more fascinating inventions of the 20th century. To fully appreciate the topic and understand how lasers work, the behavior of stimulated electrons must be explored.

ELECTROMAGNETIC WAVES AND THE PHOTOELECTRIC EFFECT

In 1865, electromagnetic waves were theorized by Scottish physicist James Clerk Maxwell. He determined that oscillating accelerated charges could produce changing magnetic and electric fields. This, in turn, produced waves. He calculated that these waves were equivalent to the speed of light: \( c = 3 \times 10^8 \text{ m/s} \). However, Maxwell was unable to prove this theory experimentally.

In 1887, building on Maxwell’s theory, German physicist Heinrich Rudolf Hertz used LC circuits to create and detect electromagnetic waves. Hertz created two circuits. The first consisted of an induction coil connected to two metal spheres. The second consisted of a single loop of wire connected to two metal spheres, illustrated in Figure 1a. Short voltage pulses were sent through the coil and into the spheres in the first circuit, creating oscillations that gave each sphere an opposite charge, causing a spark between the two. This circuit was called the transmitter. The second circuit was called the receiver because it received the energy in the form of waves, producing electromagnetic waves.

Hertz was also able to determine the wavelength of these waves by moving the receiver around and using the equation: \( v = \lambda f \), where \( v \) represents the velocity of the wavelength, \( \lambda \) (Greek symbol lambda) represents wavelength, and \( f \) represents the frequency of the wavelength. The waves were radio waves and their wavelengths were a million times greater than visible light wavelengths. Hertz also determined that if he used ultraviolet light on the negatively charged side of the spheres, attached to the single loop of wire, the spark could jump the gap between the spheres. However, he was not sure what to make of this discovery. Little did he know, he had discovered the photoelectric effect.

The photoelectric effect was explained by German-American physicist Albert Einstein in 1905. He theorized that light consisted of small bundles of energy particles called photons. When
a single photon hits a metal surface, an electron is emitted. These electrons became known as photoelectrons. The energy of a photon is calculated by the following formula: \( E = hf \), where \( h \) represents Planck’s constant, which is the connection between a particle’s frequency and its total energy\(^4\). The wavelength of a photon determines how energetic it is\(^4\). Einstein also calculated the maximum kinetic energy of a photoelectron using: \( KE_{\text{max}} = hf - \phi \), where \( \phi \) (Greek phi symbol) is the work function of the metal being used and represents the minimum energy of the electron bound to the metal\(^5\). Danish physicist Niels Bohr further studied electrons and their energy states\(^7\).

**ATOMIC ELECTRON TRANSITIONS**

In 1913, Niels Bohr created an atomic structure model of the hydrogen atom. He stated that electrons were in specific orbits and were capable of making instantaneous quantum leaps to another allowed orbit. The innermost orbit is the most stable and has the lowest energy; this is also known as the ground state. The farther orbits are less stable and electrons gain potential energy; this is known as the excited state\(^7\).

An electron can move from a ground state to an excited state when the atom absorbs a photon. This absorption can only happen if the energy of the photon equals the energy separation of the orbits that the electron moves between\(^7\). This process is also known as a stimulated absorption process\(^8\) and is represented in Figure 2\(^b\). It is highly probable that the electron will return to the ground state and the atom will emit a photon of energy in a process known as spontaneous emission, shown in Figure 3\(^c\). A third process, known as stimulated emission, can also occur. It was first described by Einstein in 1917, and is a process critical to the functioning of a laser\(^8\).

**THEORY OF STIMULATED EMISSION, THE MASER, AND THE RUBY LASER**

Einstein’s theory of stimulated emission involves an incoming photon while an atom is already in the excited state. This photon increases the likelihood of the atom to release a second photon when it travels back to the ground state, depicted in Figure 4\(^d\). The released photon has the same energy as the incoming photon in the form of radiation; therefore, the incident and emitted photon are in phase. These photons stimulate other atoms to do the same in a continuous process\(^8\). A device did not exist to demonstrate these features until almost 40 years later\(^9\).

After World War II, radar scientists began collaborating with physicists to generate shorter wavelengths on the leftover radar equipment\(^10\). Charles Townes, a physicist at Columbia University, was already familiar with microwaves so he worked with his students to come up with a new device. In 1954, they built the maser (microwave amplification by stimulated emission of radiation), seen in Figure 5\(^e\), and it operated successfully using ammonia:

A stream of excited ammonia molecules was sent through a cylindrical “foouser” which allowed only those molecules in the high-energy state to pass through. This
high-energy stream was guided into a resonant cavity where, under the influence of an electrical field, amplification by stimulated emission of radiation was achieved, producing a microwave output at to the resonant frequency of ammonia. A year later, in Moscow, physicists Nicolai G. Basov and Aleksandr M. Prokhorov independently built a maser as well. Charles Townes believed the operation was possible in certain solids, like crystals, which became apparent with a ruby.

Before the ruby was used, Townes and his brother-in-law, Arthur Schawlow, discussed how they could emit light with the maser at infrared and optical wavelengths at Bell Laboratories. They explained that if they put two mirrors, one reflective and one transparent, at each end of a long narrow cavity, the stimulated atoms could travel back and forth, which would stimulate more atoms to release photons. Some of the rays created during this process could then be released through the transparent mirror. Townes and Schawlow published their findings in 1958. During this time, graduate student at Columbia University, Gordon Gould, discussed this same issue with Townes. Gould recorded his ideas and continued his research. He filed a patent on behalf of his employer, TRG Corporation, in 1959. However, several months earlier, Townes and Schawlow had filed a patent on behalf of Bell Laboratories and it was accepted. As physicists began to understand how they could possibly emit light through these processes, everyone was working frantically to be the first discoverers.

Within the year of Townes and Schawlow’s patent, teams of physicists at several laboratories built their own masers, using either gases with an electric discharge, or crystals with tungsten lamps. But, none were capable of emitting light. It was not until physicist Theodore Maiman of Hughes Laboratories experimented with a pulse of light, rather than a continuous beam of light. This pulse of light caused more atoms to be in the excited state, rather than the ground state, which is known as population inversion. He placed a cylindrical ruby through the inner space of a helical flash lamp, as seen in Figure 6f, and on May 16, 1960, pulses of red light were observed. The first laser (light amplification by stimulated emission of radiation) was constructed. Now that it was known a flash bulb provided the solution, the various teams of physicists began making their own lasers. However, another team at Bell Laboratories wanted to perfect the use of a gas medium.

THE HELIUM-NEON LASER AND THE CO2 LASER

It was later that year that Iranian-American physicist Ali Javan and American physicist William Bennett created the first gas laser. They used an electrical discharge across a long quartz tube, with a reflector on each end, filled with both helium and neon. This discharge excited the helium atoms. The helium atoms then collided with the neon atoms, transferring their energy to them. As this process continued and the neon atoms released photons, the laser emitted a continuous beam of coherent light, as seen in Figure 7g. The highest wavelength they recorded
was in the near infrared range at 1,153 nanometers. A nanometer is a unit of length for electromagnetic waves. One nanometer is equivalent to 1.0x10^-9 meters.

Physicists at Bell Laboratories continued to research different types of gas lasers. One in particular was C. Kumar N. Patel. In late 1961, after he made his own helium-neon laser, Patel began to experiment with the concentrations of each gas. Consequently, he discovered that the neon laser worked without helium; a pure gas laser was possible. With this in mind, he continued his experiments through the next few years using noble gases, then diatomic gases, and, finally, triatomic gases. The triatomic gas that he chose was carbon dioxide because of its stability. In late 1964, he reported the first working infrared carbon dioxide laser with two wavelengths near 10,000 nanometers. Now that a more powerful gas laser existed, applications for its use were explored. Even before the carbon dioxide laser’s introduction, lasers had made their debut in the medical field.

LASERS IN THE MEDICAL FIELD-RUBY AND ARGON

One year after the ruby laser was invented, American dermatologist Dr. Leon Goldman, demonstrated that the laser could be used to remove pigment lesions from the skin or melanomas. Because of the wavelength in which the ruby laser operates, and the absorption band the pigmented skin can absorb, the pigment can be removed. He began experimenting on himself and on laboratory animals, and later determined that the laser could remove tattoos as well. In the same year, American ophthalmologist Dr. Charles J. Campbell, used a ruby laser to destroy a detached retina that was caused by a tumor. Laser eye surgery continued to improve with the use of the argon laser, which emits green-wavelength light. This light is absorbed by the pigment in red blood cells, hemoglobin, and stops the blood vessels from bleeding. The argon laser was also used to treat vascular lesions. However, it caused hypopigmentation (lightening of the skin) and scarring.

CO2 LASERS IN THE MEDICAL FIELD

After the invention of the carbon dioxide laser in 1964, a medically-developed version was devised in 1965. It was used surgically in the larynx and middle ear. In the early 1970s, Dr. Goldman used it to treat skin cancer lesions. What made the carbon dioxide laser unique, compared to either the ruby or argon lasers, was that it was absorbed by water in the body. Pigmentation or vascularity were not factors in the absorption. The laser was much more efficient than the use of a scalpel as it did not leave frayed endings, resulting in less edema and pain.

However, the carbon dioxide laser had its drawbacks. Since infrared light is invisible, the laser had to be mounted to an operating microscope. The surgeon followed the beam of the microscope, which provided guidance for the invisible beam. Later improvements to the laser included articulating arms for easier use, as shown in Figure 8.
The other concern was that the laser’s continuous coherent beam was too thermally damaging. To prevent injury, the beam’s energy needed to be less than the thermal relaxation time of the skin, which is one millisecond. This is the time it takes for the heat absorbed by the skin to cool down to half the amount that was originally applied. So, laser manufacturers went to work on pulsed lasers. Developments in ultra-pulse lasers provided a pulse duration in one millisecond.

During this time, solid and gas lasers continued to be used in various surgeries and medical procedures. They were becoming so prevalent that Dr. Goldman founded the American Society for Lasers in Medicine and Surgery (ASLMS) in 1981. This group consisted of doctors, surgeons, and physicists who collaborated on research and education about lasers. As the years and advancements progressed, lasers were used in various specialties, including dentistry, cardiology, gynecology, urology, oncology, pulmonology, and many more. Dermatology continued to be a leader amongst laser usage, specifically for resurfacing techniques, especially with the use of ablative carbon dioxide lasers.

ABLATIVE CO2 LASERS

Ablative carbon dioxide lasers ablate and vaporize the outer surface of the epidermis, an approach that is aggressive and results in wounding of the skin. These lasers emit light at 10,600 nanometers of wavelength and are primarily used in resurfacing treatments of damaged or aged skin, including sun damage, acne scars, and wrinkles. The skin’s absorption of the heat promotes collagen formation. There are two types of ablative carbon dioxide lasers, traditional (unfractionated) and fractionated.

Unfractionated carbon dioxide lasers were the first type to be used in resurfacing treatments. ‘Unfractionated’ means that the whole area being treated, say a few millimeters of skin, receives the full effect of the laser. This creates more wounding of the skin and possibly involves more risk. The risks involved with these lasers are scarring, hypopigmentation, hyperpigmentation (darkening of the skin), and infections. Although these lasers require a week or two of down time for one service, they are the most effective in resurfacing treatments.

The concept of a fractional laser was introduced by American dermatologist Rox Anderson in 2001. With further developments, the laser became available in 2004. The fractional carbon dioxide laser works by using microscopic thermal columns into the skin. Hence, only a fraction of the skin receives the thermal waves. The keratinocytes (cells in the skin responsible for immune response) in the uninjured skin migrate to the affected area and remove the necrotic debris from the skin. This process promotes faster healing and less chance of scarring. However, the same risks involved with unfractionated lasers are still possible with the fractionated lasers. Yet, this invention proved to be an enormous advantage to resurfacing treatments, as seen in figure 9.
CONCLUSION

Altogether, lasers have been an incredible invention. With the help of Maxwell, Hertz, Einstein, and Bohr, atomic energies provided a basis for the first amplification by stimulated emission of radiation. From the maser to the first sign of light in the ruby laser, physicists proved that light could be manipulated. The solid mediums led to experiments with gas mediums, which brought on the helium-neon laser and carbon dioxide laser. Both the ruby and carbon dioxide lasers quickly made their way into the medical field shortly after their discoveries. Although the carbon dioxide laser initially had it’s setbacks, advancements made a big impact in the field of dermatology.

I personally think the laser is an amazing invention. To know that physicists had the makings of a laser in the late 1800s and early 1900s, it’s a little astounding that it was created decades later. But, the laser’s introduction into medicine is what intrigues me the most, especially in surgery and dermatology, because I aspire to be a physician assistant. Having had laser resurfacing done on my own skin, I know that the affects can improve damaged skin drastically.

I think fractional ablative carbon dioxide lasers will continue to improve. I think researchers will find ways to make them less aggressive on the skin, and healing processes will shorten. Although advancements in fractional nonablative carbon dioxide lasers have become favored, I do not think they will completely take over the ablative lasers. The ablative lasers provide significant results and they will continue to improve in skincare treatments.
FIGURES

Figure 1

In this depiction of Hertz’s two circuits, the transmitter is located at the top with two metal spheres connected to the induction coil and the bottom circuit is the receiver connected to the single loop of wire. Sparks are shown between the spheres.


Figure 2

This diagram represents an electron being transferred from the ground state to the excited state by an incoming photon. This is the process of stimulated absorption.


Figure 3

This diagram represents spontaneous emission. If the atom is already in the excited state, it can fall back to the ground state. When this happens, the atom emits a photon of energy.

This diagram represents an electron in the excited state dropping down to the ground state. It becomes stimulated by an incoming photon, which causes it to emit a second photon of energy. This is the process of stimulated emission.


1954 photograph of Charles Townes adjusting the first maser.


Photograph of the ruby laser inside of the flash lamp.

Figure 7g


Bell Labs/Alcatel-Lucent USA Inc., courtesy AIP Emilio Segrè Visual Archives, Hecht Collection

Figure 8h

These are images of various carbon dioxide lasers in the 80s with articulate arms.


Figure 9i

Before and after photos of a patient treated with the fractional carbon dioxide laser on their acne scars.

REFERENCES


Illuminating Sports

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Abstract

This literature review discusses the physics and history of lighting a sports arena. From the point of discovery, to the inner workings and application of lighting. This review also discusses different lamps used in lighting, such as, high-intensity discharge (HID) lamps and light emitting diode (LED) lamps. Both are utilized in different applications, with sports stadiums using them on a grand scale. Over the years, there were many important discoveries in the transfer of electricity and changes to the type of bulbs used for lighting. These findings paved the way for audiences to thoroughly enjoy athletic performances in a stadium. Each style of lamp proved to have certain advantages and disadvantages.

There is something to be said about individual and team sports. They grab the attention of spectators from all over the globe. From Jesse Owens winning 4 gold medals in a single track-and-field competition to Michael Jordan scoring the most points in a single National Basketball Association (NBA) playoff game, sports are exciting, heart-warming and nail biting all rolled into one. Society would not be able to truly experience these emotions without grand scale over the top stadiums.

The first type of stadiums built were strictly to allow large groups of people to watch a live sporting event. These arenas were very basic facilities. Different levels were created by concrete tiers or an arrangement of dirt mounds to allow for standing room only. Those who were considered important observers received a designated seating area, which was small and sometimes with an overhang structure to protect them from the elements. These beginning stadiums took on different formats until the end of the 1950s. Due to the abrupt drop in people attending sporting events, there needed to be a change to make stadiums more comfortable. Televised sporting events became more desirable than attending the live event because spectators could enjoy the game in the comforts of their home. In addition to these comforts, the images on the television gave for better visibility instead of being at the stadium. Many renovations took place including the addition of lighting for better reflectivity and to guarantee broadcasting could take place at night. These additions allowed onlookers to encounter the most optimal sports entertainment experience.

Starting in the 1820s and 1830s, electricity and lighting became more than just scientific interest. Three physicists made important discoveries related to electrical current and magnetism – Hans Christian Oersted, Andre Marie Ampere and Francois Arago. They established a path for later physicist Michael Faraday to discover electromagnetic rotation and induction. These discoveries led to the development of electric generators, electric motors and transformers.

Hans Christian Oersted first discovered electromagnetism in 1820 by accident. He noticed the needle of a compass readjusted each time it was brought near a wire actively carrying a current. If the current reversed directions, the compass needle automatically reversed direction. From this reaction, Oersted was able to determine that a magnetic field circles the wire where current is flowing. A magnetic field is represented by a series of field lines forming closed loops that are perpendicular to the current. This is where the right-hand rule began. The right-hand rule allows one to predict the direction of the magnetic field lines when the direction of an electron in motion is known. The thumb points in the direction of the electron’s motion, while the
fingers indicate the direction of the magnetic field lines. A compass needle will then align tangent to the magnetic field lines. Oersted’s innovation was then applied to create the first true telegraph. A telegraph is a system that transmits messages through an electrical current across long distances.

Andre Marie Ampere took Oersted’s discovery and went on to further explain the idea of magnetism. He showed that the magnet in Oersted’s original experiment can be changed to a different current-carrying wire. In doing this, he found that current-carrying wires parallel to each other will experience an attraction to one another when current is flowing in the same direction. When the current is going in opposite directions, the wires will repel each other. Ampere’s famous formula was born from this discovery and remains as an integral part of electrodynamics. His formula stated: the force is proportional to the product of the currents in the two wires and to the length of the wires, and inversely proportional to the square of the distance between the two.

Francois Arago was the last influential physicist to Michael Faraday. He expanded on Oersted’s work when he discovered that an electric current will pass through cylindrical copper wire to attract iron filings. When the current is turned off, the filings are no longer magnetized and will fall off.

Faraday began his experiments in 1821. His first discovery was electromagnetic rotation – an electrical current and magnetic force will rotate about each other. His second discovery occurred in 1831 when he found electromagnetic induction – when a magnetic force rotates a copper wire, an electric current is induced. From these findings came Faraday’s law of magnetic induction, which states: If a circuit contains a certain number of tightly wound loops and the magnetic flux through each loop changes during a certain time interval, the average electromotive force induced in the circuit is the number of turns in a coil multiplied by the change in flux divided by the change in time.

A great application of Faraday’s law is the transformer. A transformer changes an alternating current (AC) from one level to another. It consists of two sets of insulated wiring which are coiled around an iron core. When electrical power is applied to the first pair of coiled wires, it experiences a magnetic force which is transferred to the second set of wiring. This setup effectively changes AC voltage level and isolates two electrical circuits. The electrical conversion created by a transformer allows power to be altered multiple times as it goes through different transmission lines from a power station.

There are two types of transformers – the step-up transformer and the step-down transformer. These differences are based on voltage intensity on one of the coils. If the secondary coil experiences a voltage greater than the primary coil, it is considered a step-up transformer. If the primary coil experiences the greater voltage, then it is called a step-down transformer. Even though these voltage changes take place on the coils, the product of current, voltage or power remain constant.

Transformers are an important element for lighting a sports arena. This is because of the extensive amount of lamps hung throughout, and the electromotive force (or voltage) needed to illuminate them. The evolution of these lamps have brought greater efficiency and are now more environmentally friendly. Most sports arenas are already utilizing this new technology, while others are still in the process of switching.

The life of a light bulb goes back to the early 1800s when chemist Humphry Davy discovered an electric arc. This occurs when an electrical current jumps from one
electrode to another when there is a gap in a circuit. From this discovery, he created the arc lamp. The arc lamp produces light from an arc of electric current passing through ionized gas between two electrodes. Contained within the bulb could be any of the following elements: argon, metal halide, mercury, neon, sodium or xenon. The downside of using an arc lamp was its requirement of an expensive battery or generator in order to run. As a result, the general public were not so interested in using arc lamps as a way of lighting. It was in 1871 that the solution was created; the dynamo was invented.

The dynamo was a generator which produced direct current power to arc lamps for relatively cheap. This invention sparked public interest in using arc lamps again, and they were installed any place where lighting was needed to cover a large open space e.g. stadiums, factories, etc. The generator’s origin goes back to 1831 when physicists Michael Faraday and Joseph Henry made the discovery. A generator contains a magnet which establishes a magnetic field. There is a looped wire rotating within the magnetic field. With each rotation, the magnetic field strength changes through the wiring. This generates a force that pushes an electric charge around the wire. The force changes the route of the current after one revolution. These changes create an alternating current.

There are two categories of generators – alternating current (AC) generators and direct current (DC) generators. The AC generator works exactly as previously described. With a fixed magnetic field, coiled wiring rotates creating a force that carries an electric charge. The charge travels one direction for a period of time. After a certain length of time, the electric charge will then reverse its’ direction of travel. A DC generator uses slip rings which are cut in half and insulated; this is called a commutator. The commutator has brushes attached to each piece which are arranged in a way to maintain the current in a unidirectional pattern. As soon as the current changes directions, the coil carrying the current will slip into an external circuit making it continue to travel in the same direction.

It was in the 1880s when a historical feud between AC generators and DC generators began. The disagreement was between Croatian electrical engineer Nikola Tesla and American inventor Thomas Edison. Tesla believed electrical current should switch flow direction multiple times a day, whereas Edison was promoting the direct current approach – electric current flowing continuously in one direction. Edison’s idea was already implemented at this time and working successfully, so the idea of making the switch to AC generators was not received well. This resulted in Tesla’s bankruptcy and his switch from invention to manual laborer. His time spent as a manual laborer was cut short when he met George Westinghouse.

George Westinghouse was intrigued by Tesla’s AC current approach as it could eliminate the need for multiple DC generators. Westinghouse established his own electrical grid, Westinghouse Electric Company, in Great Barrington, Massachusetts around 1886. He purchased the rights of the previously developed transformer and rebuilt it. He incorporated Tesla’s AC generator into this electrical grid. This new system proved to be more beneficial. Its’ signature ability was to reach more people without any noticeable difference in lighting intensity when the current switched. The transformer and generator are powerful sources contributing to the bright lights and neon signs utilized in sports arenas.
As time shifted to the twentieth century, different alterations of the arc lamp were created. It was within this time period that electrical engineer Peter Hewitt invented the first discharge lamp. This discharge lamp was also called a mercury vapor lamp because of the use of mercury vapor to create light. The light illuminates once it encounters a certain strength of voltage, causing the gas to give off a blue hue. If the blue light is not desirable, it can be fixed by coating the container with different phosphors. Phosphors transform ultraviolet output from the mercury vapor into observable light by fluorescence. The different mixtures can be rearranged to accommodate the lighting necessary for activities taking place by its’ tenants. These types of lamps provide optimal lighting for large surface areas.

The competitor to the arc lamp is the metal halide lamp. These lamps are widely used in sports stadiums because of the benefits for television broadcasting. The amount of illumination given off by metal halide lamps allows for crisp imagery to fans watching in front of a color television set. The setup of a metal halide lamp is what makes its’ lighting efficiency perfect for watching sports. The lamp uses mercury vapor similar to an arc lamp, but it also contains a group of halide salts. There are a total of five halogens included to enhance the glowing color. These halogens are fluorine, chlorine, bromine, iodine, and astatine. All of these elements are contained in a quartz tube that has been fused. They are condensed inside when the lamp is off. Once the lamp is turned on, an electrical current passes through a starting electrode comprised of tungsten. The current then jumps to a second electrode with tungsten coiled around it. This jump occurs from the assistance of argon gas. The electrical current is not received straight from an AC generator power source. Instead, the lamp is connected to a resistor which is then connected to a ballast.

A ballast is important to the metal halide lighting system because of the current regulation. Without this system in place, the lamp would draw on the high voltage power source very rapidly. This pull of electrical power at an uncontrolled rate causes the lamp to overheat and burn out quickly. When the lamp is first turned on, the ballast provides a short burst of high voltage to create an arc between the two electrodes. Once this is complete, the ballast will then reduce the voltage supply and maintain a steady input of electric current, thus providing a steady light output.

Once a current reaches the metal halide arch discharge tube, the mercury enclosed is vaporized. As the current moves through the mercury vapor it experiences a resistance. This resistance slowly dissipates as the gas becomes more ionized. This is when optimal travel conditions of the current is made. The current reaches the second electrode with halide salts and they are quickly vaporized. Immediately following this reaction, metal atoms move to cooler areas within the lamp and recombine with the halogen producing its’ perfect white glow.

The metal halide lamp was created by Charles Proteus Steinmetz. Steinmetz was an electrical engineer and socialist in Germany until the need to escape to the United States came about. He began working with then established General Electric Company (GE) in 1893. On May 7, 1912, nineteen years into his employment with GE, he was granted a patent for the metal halide lamp. His idea was to enhance the glow of mercury vapor lamps. The approach, add a grouping of metal halide salts. He reached success at improving the color of the light, but he ran into problems in maintaining a consistent arc.
Steinmetz’ and colleagues’ inventions led to the first baseball game to be played at night. It was May 24, 1935 at Crosley Field in Cincinnati, Ohio. The teams going against each other were the Cincinnati Reds and the Philadelphia Phillies. GE was requested to design how the lights would be setup, including amount of lights to use, the towers needed, and the desired height. GE worked together with the Cincinnati Gas and Electric Company to make this lighting system better than the best lit minor league baseball park. The lights used for this grand spectacle GE named Mazda lights – which are also called incandescent lights. An incandescent light glows when an object is heated to a certain temperature. In order to attain white light, the object has to reach 727 degrees Celsius.

For the history of the incandescent lamp, attention returns to Sir Humphry Davy. In 1802 he found that electricity running through strips of metal will conduct heat which is hot enough to produce a light. His discovery established the general principle and inner workings of incandescent lamps. Fast forward to the 1870s brings Thomas Edison into the picture. Recognized as the light bulb inventor, he was one of many contributors to researching and designing incandescent lamps. He changed the internal metal strips to high-resistance carbon filament made of bamboo. These materials were placed inside a vacuum sealed glass container. Creating this vacuum seal is what increased the glowing life of the carbon filament. Edison’s light bulb was above the rest because of the increased electrical resistance. Light bulbs prior to Edison’s went up to a resistance of four ohms. Edison’s bulb was far superior as it went up to an astonishing two thousand ohms. With resistance exponentially larger, the copper wiring that carried a current could extend to greater lengths, thus reaching more people and lighting more places.

Nine years later, in England, Joseph Swan borrowed Edison’s idea for the light bulb and added platinum lead wires. In February of 1879, the English chemist demonstrated this incandescent bulb with grand success. Swan also changed the type of carbon filament from bamboo to cellulose; a material found in the cell wall of plants. Swan’s design was not void of any faults. When the carbon was activated by an electrical current, soot would release and cover the inside of the bulb. This reoccurring problem made the bulb a poor light source. From the year 1898 to 1924, many alterations were made to the incandescent bulb, including the addition of argon gas, the replacement of carbon filaments with tungsten, and the silicon coating inside. The switch to tungsten increased bulb efficiency to eight lumens per watt (lpw) from the previous three lpw. By adding argon gas, the effectiveness doubled and it reduced the rate at which tungsten would evaporate; the culprit causing the bulb to become black. The silicon coating inside the bulb provided a protective barrier, resulting in a reduced rate of breakage.

Incandescent bulbs, mercury vapor lamps and metal halide lamps paved the way for millions of people to light their homes, and enjoy countless activities after their work day has ended. When the addition of lighting sports arenas came about, the sports played were quickly revived. Attendance count for day games reported to be approximately 2,000 to 3,000 fans at the Cincinnati Reds’ Crosley Stadium. Most individuals had trouble attending day games because of their work schedule. On the evening of the first night game on May 24, 1935, attendance grew to an astonishing 20,000 fans.

The newest member of the light bulb family is the light-emitting diode (LED) lamp. LED bulbs convert electrical energy to light by acting as a semiconductor, where the electrical conductivity is great when it is hot and low when cooled. This
semiconductor is broken up into two regions – the p-region and the n-region\(^\text{20}\). The p-region encompasses positive electric charges, and the n-region contains negative electric charges\(^\text{20}\). Separating these regions is a junction made up of alloyed crystal called a depletion layer\(^\text{10}\). The crystal destabilizes the p-region and n-region to accurately affect conductivity\(^\text{10}\). This makes the electrons move unidirectional when a charged current flows through the semiconductor held between two electrodes\(^\text{10}\). The electrons then shed any energy picked up from the electrical current and create photons which emit a white glow\(^\text{10}\).

An LED bulb has demonstrated to be a better light option than the classic incandescent bulbs, mercury vapor lamps and metal halide lamps. These bulbs use no mercury, argon gas, carbon filament, and are much sturdier than a glass bulb\(^\text{20}\). Additional benefits of using LED bulbs include lifespan, color quality, and lumen maintenance\(^\text{20}\). The lifespan of an LED bulb is at least 50,000 hours, the white light illuminating from them is clear and concise, and the output of light remains constant\(^\text{20}\). Another advantage of LED bulbs is their quick response to produce light when receiving a voltage\(^\text{20}\).

Many sports arenas are concerned about switching their metal halide lamps to LED lamps. One major concern was that the LED technology is not quite ready to take on the amount of light needed for football fields or baseball parks\(^\text{13}\). Two professional football teams went ahead with the switch in 2012 – the Houston Texans and the Arizona Cardinals\(^\text{13}\). University of Phoenix stadium, home of the Arizona Cardinals, replaced 780 of their metal halide lamps with 312 LED lights\(^\text{13}\). This important change made the field appear brighter, which in turn improves the quality of broadcasting games on television, and improves the live fan experience\(^\text{13}\).

**Conclusion**

In conclusion, the research supports the advancements in lighting technology, purposefully and by accident, are invaluable. Each change to a light bulb created a whole new era of innovation. I personally enjoy the excitement of watching sports, whether at home or in a sports arena. It is difficult for me to imagine what it was like trying to watch sports in a stadium that didn’t have advanced comforts and great field visibility. With the new age approach of going green, there has been a trend to switch from the traditional metal halide lamps to LED lighting. The advantages surely outweigh any disadvantages, as indicated by the research. I believe the future will be much brighter as more professional sports teams follow University of Phoenix Stadium and make the switch to LED bulbs.
Diagrams/Figures/Pictures

When a magnet is moved toward a loop of wire, the ammeter registers a current.

When the magnet is stationary, no current is induced.

When the magnet is moved away from the wire loop, the ammeter registers a current in the opposite direction.

(Faraday's law of magnetic induction (Fig 20.4) – photo courtesy of College Physics, Volume 2, 10th Ed. – section 20.2, pg 703)
The mechanical energy input to a generator turns the coil in the magnetic field.

A voltage proportional to the rate of change of the area facing the magnetic field is generated in the coil. This is an example of Faraday's law.
Metal Halide Lamp Schematic

(Metal Halide Lamp Schematic – photo courtesy of www.edisontechcenter.org)

Metal Halide Lamp - Arc Discharge Tube

(Metal Halide Lamp – Arc Discharge Tube. Photo courtesy of www.edisontechcenter.org)
The first Metal Halide Lamp. This is a copy of the submission for a patent by Charles Steinmetz – photo courtesy of www.americanhistory.si.edu/lighting/history/patents/stein1.htm
Crews test the system at University of Phoenix Stadium.
Photo by: Ephesus Lighting

(Crew for Ephesus Lighting testing the new LED system at University of Phoenix Stadium – photo courtesy of Ephesus Lighting, as cited in www.sportsbusinessdaily.com/journal/issues/2014/09/29/In-Depth/LED-lead.aspx)

(Photo and schematics of an LED bulb – photos courtesy of www.ies.org/lighting/sources/led.cfm)
References


Fluid Dynamics and the Body

Katie Hawk

Physics 112
Dr. Casey Durandet
April 18, 2015
ABSTRACT

This paper is about the fluid dynamics of blood within the human circulatory system and the appropriate method of fluid resuscitation in patients who have experienced traumatic injury resulting in hemorrhage and a hypovolemic state. Appropriate fluid resuscitation will be most effective when administered via a large bore intravenous catheter with a large cross sectional area (A) and a short length (L). This is supported by Poiseuille’s Law and the equation \( \frac{\Delta V}{\Delta t} = \frac{\pi R^4 (P_1 - P_2)}{8 \eta L} \) (Serway and Vuille 2012). Bernoulli’s equation, \( P + \frac{1}{2} \rho v^2 + \rho g y \), will also be discussed as this is applied to describe flow through the large arteries and veins in the vascular system, such as the aorta (carrying blood away from the heart) and superior and inferior vena cava (bringing blood back to the heart).

INTRODUCTION

Jean Leonard Marie Poiseuille, born on April 22, 1799, was a scientist who completed the majority of his studies and experiments between the years of 1828-1868 (Skalak and Sutera 1993). Poiseuille has been a large contributor to the medical community as it pertains to the flow of blood through the vascular system. There is little documentation providing information on how his experiments were completed or where the financial support came from. The first presentation of his work regarding the flow of fluids through small diameter tubes was presented in 1838. Shortly after that his work was published by the Academy of Sciences in 1841. Initially, he used distilled water as the medium for flow measurements, but upon acquisition of more data he began to compare other liquids such as acids, salt solutions, mineral waters, plant extracts and many others for comparison. This allowed the effect of viscosity on flow to be indirectly studied. During this time Poiseuille and another colleague developed a tool to measure pressure, the hemodynamometer, which was measured in, millimeters of mercury (mm Hg). This is still the unit of measurement used to record blood pressures in medical practice today.

Poiseuille’s studies specifically focused on the microvasculature of the circulatory system, such as arterioles, capillaries and venules (Skalak and Sutera 1993). The mechanism and behavior of this flow through small diameter tubes was studied in the 15 different experiments he reported on throughout his scientific career. The goal was to identify the relationship between the following variables: volume efflux rate from a tube (Q), the pressure difference between the opposite ends of the tube (\( P_1 - P_2 \) or \( \Delta P \)), the length of the tube (L), and the radius of the tube (D). Poiseuille did not initially account for the viscosity, the friction inside of the fluid that occurs naturally, of the fluids he worked with, but as he recorded data at different temperatures and yielded very precise results, the original formula, \( Q = \frac{K'' PD^4}{L} \), was utilized with a constant (K) he derived for distilled water (\( K''=2495.224 \)). The more recognized equation, \( Q = \frac{\pi R^4 (P_1 - P_2)}{8 \eta L} \), calculates Q with a known value for viscosity, \( \eta \) (Serway and Vuille 2012). In addition, the radius (R) is used in place of D (Figure 9).

During his experiment, Poiseuille concluded that as the radius (raised to the fourth power) increased, the rate of flow increased as well. In addition, the change in pressure also affected flow rate the same way. Conversely, if the fluid that is flowing through the tube
becomes more viscous, there is more resistance and the flow rate decreases. The length of the tube increases the resistance as well in such a way that if the tube is longer, flow rate will decrease. An example of this as it pertains to blood flow in the microcirculatory system would be for a comparison of a person who is adequately hydrated to one who is not adequately hydrated. A person who has drank enough water will have more dilute blood such that there are less blood cells per mm$^3$ and therefore a less viscous fluid is circulating. This person will have a higher flow rate than the person who has more viscous blood assuming that the driving force of pressure (heart beat) is held constant (Fu and Sugihara-Seki 2004). This describes the inverse relationship between fluid viscosity and flow rate.

In order to utilize Poiseuille’s Law for blood flow calculation and fluid flow through tubing during resuscitation, there a several assumptions that must be made (Criss et al. 1985). First, viscosity is held constant. Additionally, the flow is laminar and non-turbulent in a rigid and uniform cylindrical tube. Laminar flow describes the movement of particles in the stream as staying a constant distance from the wall of the tube and flow will be predictable. In addition, contents in the blood stream also tend to be attracted toward the center of the artery or vein decreasing the amount of potential resistance and slip along the walls of the vasculature (Grobenlik 2008). This also explains why fluid velocity is greatest in the center of the pipe (Serway and Vuille 2012) and zero at the wall (Grobelnik 2008). When the flow of the fluid develops eddies (such as in the presence of an aneurysm along the vascular wall) or becomes erratic and unpredictable, it is no longer laminar flow (Finck ND). The flow of blood through the human circulatory system is however, not always predictable and steady, but according to Grobelnik (2008), the flow models that have been demonstrated give good insight into this fluid movement. As it pertains to rapid fluid resuscitation, the flow is considered representative of models in the large and proximal arteries and veins, making it accurate and comparable (van Dongen and van de Vosse 1998).

The blood stream is made up of plasma, a serum of proteins, electrolytes and platelets for a total plasma volume of 55% and a second whole cell mixture, totaling 45%, made up red blood cells and white blood cells (Grobelnik 2008). In large vessels, the viscosity is not considered to be variable, but in smaller vessels, (less than 0.5mm) the viscosity becomes variable. Bernoulli’s application and equation, \( P + \frac{1}{2} \rho v^2 + \rho gy = \text{constant} \), accounts for the ability of blood to continue flowing through the circulatory system to smaller and smaller vasculature while maintaining a constant movement of mass. This prevents constriction due to stenosis from interfering with perfusion. In addition, as branching occurs, which is virtually the only mechanism for changes in lumen size assuming vasoconstriction and relaxation is held constant at any moment in a healthy individual, the total cross sectional area is actually increasing as circulation moves distally from the heart (Grobelnik 2008). According to the Grobelnik (2008), the total cross sectional area of the vasculature at the capillary level is actually approximately 300-400 times larger than that of the aorta, the largest vessel in the body. Consequently, the overall flow rate is significantly less than when it first exits the left ventricle of the heart and into the aorta.

Utilizing the equation, \( A_1 V_1 = A_2 V_2 \), and conservation of mass (Figure 5), as the area decreases, velocity increases. Bernoulli’s equation, allows a constant flow of blood to be maintained despite interferences from narrowing vessels, stenosis or other pressure changes.
Another aspect of blood flow must be considered when comparing the arterial system to the venous system. The arterial system is under higher pressure and blood flows at a faster velocity. The movement of blood flow in arteries is active and has kinetic energy (actual flow of blood) and potential energy (pressure from the veins driving it forward) that force it further into the microvasculature and distal circulatory systems such as the capillary networks (see Figure 6). When blood is returning to the heart, there is no independent kinetic energy forces present and the movement of blood is the potential energy from the pressure of the venous walls and contraction of skeletal muscle being converted to kinetic energy as one walks or moves (Boundless 2014). This mechanism drives the venous blood back into the larger (but a total cross sectional area that is less, thus facilitating increased velocity as blood flow becomes more proximal to the heart) veins such as the superior and inferior vena cava that empty blood back in the first chamber of the heart, the right atrium. In addition, there are valves present in the venous system to prevent back flow of blood due to the effects of gravity. For example, when one is standing, blood is able to move upward in the legs toward the heart because of valves. These valves are present in large quantities in the extremities compared to the trunk of the human body (Badeer 2001).

There are several tools that are utilized that follow the same principles that guide fluid resuscitation that is rapid and effective in the patient who has sustained traumatic injury resulting in hemorrhage and hypovolemia. This is necessary due to the fact that with inadequate blood flow, the major organs necessary to support life, specifically the heart and the brain, will not survive and will result in cardiac failure or brain death. Rapid intravenous access is key with initial attempts beginning with peripheral intravenous access in the most proximal locations (Bhananker 2012). For example, the antecubital vein in the crook of the elbow in either the left or right upper extremity is typically large enough to support cannulazation with an 18g needle (Figure 1). A certain type of 18g needle and catheter, B|Braun, has a diameter of 1.3mm and a length of 32mm, with a flow rate from gravity of 105 ml/min (see Figure 1). In the study by Morrison et al. the 18g catheter has a length of 45mm, therefore the pictures of the B|Braun catheters in Figure 1 will actually have a faster flow rate than those listed in Figure 7 because the shorter length will have less resistance.

According to the table in Figure 8 from Hyperphysics.edu (2005), the averaged viscosity of blood in a patient with red blood cells counts that are within normal limits is 4.0, compared to water, which is approximately 1.0 or less. Poiseuille’s Law demonstrates the significance viscosity will have on flow rate: the flow rate of blood will be 25% that of water (or normal saline) because viscosity in inversely proportional to flow rate. In the average 90kg man, the total blood volume can be estimated to be 6,750ml of blood (Medscape 2011). With minimal ideal access, two peripheral large bore (18-gauge) intravenous access devices, and without any pressure infusing devices, total blood volume would take 32.13 minutes to replace with normal saline, an isotonic fluid used as the standard for fluid replacement with a viscosity comparable from water. See following calculations:

\[
(1000ml \text{ per bag})/(105 \text{ ml/min}) = 9.52 \text{ minutes per bag}; (9.52 \text{ minutes})(6.750L \text{ average blood volume}) = 64.26 \text{ minutes to replace 6.75L of blood with Normal Saline. Divide this time interval between two points of access: requires 32.13 minutes to replace the total estimated blood volume.}
\]
Normal saline is a less viscous solution than packed red blood cells, which is the standard that is used for blood product replacement. Therefore, the flow rate would be lower and consequently result in a longer time period for total volume replacement by blood products. In a person who is actively hemorrhaging, 32 minutes is already insufficient and will likely results in cardiac failure secondary to hypovolemic shock; or if resuscitation occurs, there is potential for a secondary anoxic brain injury or end-organ damage (i.e. kidneys, liver, etc.) from poor perfusion related to inadequate volume.

On page XXXX, Figure 7, from the article Intravenous fluid resuscitation: was Poiseuille right? (Morrison et al. 2011), demonstrates the flow rate of devices with normal saline from gravity, with a pressure infuser, with a needleless access valve (Figure 4), and the percent increase and percent decrease with pressure infuser and needleless access valve. The 22g needle has the shortest length, (see Figure 2), but due to the smaller radius, it will still have a lesser flow rate, Q value, due to the exponential change in flow rate secondary to the change in radius (raised to the fourth power) in comparison to the change in flow rate from a change in lumen length, L. This introduces the topic of alternative tools that’s are available to combat prolonged times for fluid resuscitation that can be used independently or in conjunction with one another to administer fluid and blood products as rapidly and as efficiently as possible for the best possible patient outcomes.

In situations in which peripheral IV access is unobtainable or inadequate, a second mode of access is available: the central venous catheter (CVC) better known as a central line. In this scenario, a large catheter is introduced into a vein that is not in an extremity, such as the internal jugular (IJ) vein in the neck, the subclavian vein on either the left or right side of the upper thorax, or the right or left femoral vein in the groin. In this case the most distal tip of the catheter from the insertion site will terminate in the superior vena cava from the IJ or subclavian access or the inferior vena cave from femoral access. There are benefits to CVC access device as the large bore catheter is made up of multiple lumens, often three (See figure 3). It allows administration of multiple products that may not be compatible together, such as certain antibiotics with sedation or analgesic medications, as well as the acceptable length of time this device can remain in the body. The peripheral IV has a lifespan of 3-4 days, where the CVC is up to 1 week before concern for infection begins to rise. In the plan of care following rapid resuscitation of the patient, this access device will be ideal. However, as it pertains to rapid administration of fluid or blood products this access is not beneficial.

According to Figure 3, the large catheter is made up of three separate lumens, one 16-gauge lumen, and two 18-gauge lumens. Initially, one would consider the 18-gauge proximal port of the central line to be more efficient than an 18-gauge peripheral IV access point, but the length of the catheter of a certain central line can be up to 34 cm (opposed to 45 mm). When comparing the rate of infusion with gravity of this device after considering the decreased rate in flow secondary to the length of the lumen the 18-gauge peripheral IV is significantly better: the peripheral IV infused at 98.1 ml/min compared to the central lines infusion rate of 29.7 ml/min.

Other tools exist in this setting to aid in rapid fluid administration, such as the Ranger (see Figure 11), better known as a rapid infuser (3M, 2015). This device supplies up to 300mm
Hg pressure ($P_1$) behind the infusing fluid to increase the pressure gradient ($P_1-P_2$) in the equation of Poiseuille’s Law, which therefore increases the value of $Q$. In addition, the tubing for this device is large bore tubing, which consequently results in an exponentially increased $Q$ rate secondary to $R$ being raised to the fourth power in the formula. Another pressure device that is available is a pressure bag. This device utilizes a hand held pump that can be pumped up to a pressure of 300 mm Hg; however, this is not a continuous pressure and as the fluid inside the bag that is pressurized flows out, the pressure in return decreases because the pressure is not constantly applied. This pressure bag was tested in conjunction with the previously mentioned access devices: the 18-gauge peripheral IV and the lumen of the proximal port of the 18-gauge CVC. According to Figure 7 from Morrison et al. (2011), the rapid infuser had a significantly greater impact when used in sequence with the central line than with the peripheral IV. The central line flow rate increased from 29.7 ml/min to 79.3 ml/min, a 167% increase; the peripheral IV increased from 98.1 ml/min to 153.1 ml/min, a 56% increase. Due to the fact that these values are a result of the pressure bag over the course of infusion time, the pressure steadily decreased and is not an accurate representation of a constant pressure of 300 mm Hg. In this case, the rapid infuser would be more efficient and would result in even faster resuscitation.

At the bottom of Figure 11, a second device is illustrated, which is a warming device that can be set to a defined temperature, typically 41°C. According to Serway and Vuille (2012), the viscosity of water (a comparison that is considered equivalent to normal saline) is less at higher temperatures. This was also demonstrated by Poiseuille’s original experiments in 1838-1841 when he identified $K''$ as his constant for various temperatures when calculating flow rate, $Q$ (Skalak and Sutera 1993). At 20°C, viscosity of water is equal to $1.0 \times 10^{-3} \text{ N}\cdot\text{s}/\text{m}^2$; at 100°C, viscosity of water is equal to $0.3 \times 10^{-3} \text{ N}\cdot\text{s}/\text{m}^2$. The average body temperature is 37°C. Additionally, prior to administration blood is kept refrigerated, which results in an increased viscosity if administered without the warmer. Published data by 3M (2015) for the ranger rapid infuser and warming systems guarantees efficient warming for rates from 5 ml/min to 500 ml/min.

For the patient who is still prehospital, such as during transport from the field scene of incident, to the hospital, by emergency medical services, the peripheral IV access is the first attempt. When this is unsuccessful within 3 attempts or 90 seconds, it is recommended to attempt another method (Bhananker 2012). Paramedics no longer insert central lines in the field and now utilize the intraosseous route. This method demonstrates that 93% of the time, access is successful; furthermore success is achieved 82% of the time on the first attempt. The IO device comes in 3 sizes and even though there is not a clear published maximum infusion flow rate, as much as 204 ml/min has been recorded (with the addition of a pressure bag). This mechanism for access is not the first choice due to the discomfort associated with insertion as a drill is used to insert the cannula into the bone, but in a life-threatening situation it is very applicable. Areas on the body that this is effective include the proximal and distal tibia, the humeral head and the superior sternum in adults. Surprisingly, the $P_2$ value is not too high to prevent adequate flow due to the fact that the tip of the cannula terminates in the soft inner portion of the bone, which is also highly vascularized and during cardiopulmonary resuscitation, medication can be circulated in under 2 minutes (Bhananker 2012). When comparing the gravity flow rates of access in the humeral head via IO access to a central venous infusion in an adult, there is little difference. This still leaves peripheral IV access as the first choice in any hemorrhagic scenario.
The last option to consider is the use of large bore intravenous tubing to connect the intravenous access device to the reservoir of fluid that is infusing (Criss). While administered, the larger tubing demonstrated a 90% increase with the pressure bag and 126% increased when administered from gravity. Additionally, there an access device referred to as a cordis, which is the introducer for the central line (opens the skin and vessel enough to allow threading of the multi-lumen catheter) (Dutky et al. 1989). The cordis is only one lumen, often times a 10-gauge (Hansbrough et al. 1983). Without multiple lumens, the single lumen has one large radius have an exponential increase on flow rate and is thus the prime choice for administration of rapid transfusion in the most severely hemorrhaging patients. According to Hansbrough et al. (1983), flow rates through the cordis with large bore tubing and pressure infusion were measured at a maximum infusion rate of 1200 ml/min.

CONCLUSION

In conclusion, it can be determined that fluid dynamics within the body are variable depending on many factors such as where pressures are being measured in the body, if it is venous or arterial, or the viscosity of the blood (how hydrated the patient is). However, the flow through the vasculature of the human body is considered to be equitable to Poiseuille’s Law and Bernoulli’s equation due to the fact that it can be assumed that the simple rules are followed: blood is a Newtonian fluid, the vasculature in cylindrical, firm and consistent throughout, and the fluid in non-compressible with a constant viscosity. In addition, these same principles can be applied when considering rapid fluid administration for the resuscitation of an actively hemorrhaging patient due to traumatic injury. Based on the results of the reviewed studies, it can be concluded that multiple large bore (18-gauge or greater) peripheral intravenous access devices is the best method for resuscitation due to the short length of the catheter and larger diameter of the lumen (raised to the fourth power in Poiseuille’s Law). As a nurse, in this field I can actively utilize this information in practice as well as disseminate it to fellow colleagues in hope to improve the outcomes of the patients I care for the Emergency Department and Intensive Care Unit setting.
FIGURES

Figure 1: In order: 16g, 18g, 20g, and 22g needles and cannulas

Figure 2: Cannulas.

Figure 3: Cross section of CVC (Arrowintl. 2015).

Figure 4: Needleless access valve. (Morrison et al. 2011)

Figure 5: Conservation of Mass

Figure 6: Bernoulli’s Equation and conservation of energy
Figure 7: Flow rates of devices (Morrison et al. 2011)

<table>
<thead>
<tr>
<th>Intravenous catheter</th>
<th>Rate of flow with gravity (ml/min)</th>
<th>Rate of flow with pressure (ml/min)</th>
<th>Rate of flow with Bionector (ml/min)</th>
<th>Percentage increase with pressure</th>
<th>Percentage decrease with Bionector</th>
</tr>
</thead>
<tbody>
<tr>
<td>14G 50 mm cannula</td>
<td>236.1</td>
<td>384.2</td>
<td>138.3</td>
<td>62.7%</td>
<td>-41.4%</td>
</tr>
<tr>
<td>14G 14 cm Abbocath</td>
<td>197</td>
<td>366</td>
<td>131.3</td>
<td>85.8%</td>
<td>-33.4%</td>
</tr>
<tr>
<td>16G 50 mm cannula</td>
<td>154.7</td>
<td>334.4</td>
<td>109.5</td>
<td>116.2%</td>
<td>-29.2%</td>
</tr>
<tr>
<td>14G 15 cm Leadercath</td>
<td>117.3</td>
<td>211.1</td>
<td>101.1</td>
<td>80%</td>
<td>-13.8%</td>
</tr>
<tr>
<td>18G 45 mm cannula</td>
<td>98.1</td>
<td>153.1</td>
<td>80.3</td>
<td>56%</td>
<td>-18.1%</td>
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<tr>
<td>16G distal port triple lumen central line</td>
<td>69.4</td>
<td>116.1</td>
<td>67.4</td>
<td>67.3%</td>
<td>-2.88%</td>
</tr>
<tr>
<td>20G 33 mm cannula</td>
<td>64.4</td>
<td>105.1</td>
<td>58.5</td>
<td>63.2%</td>
<td>-9.17%</td>
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<tr>
<td>22G 25 mm cannula</td>
<td>35.7</td>
<td>71.4</td>
<td>24.7</td>
<td>100%</td>
<td>-2.80%</td>
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<tr>
<td>18G proximal port triple lumen central line</td>
<td>29.7</td>
<td>79.3</td>
<td>28.7</td>
<td>167%</td>
<td>-3.37%</td>
</tr>
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</table>

Figure 8: Viscosity Values (hyperphysics.edu 2005)

<table>
<thead>
<tr>
<th>Liquids</th>
<th>Viscosity (Poise)</th>
</tr>
</thead>
<tbody>
<tr>
<td>Acetone</td>
<td>0.0032</td>
</tr>
<tr>
<td>Alcohol(ethyl)</td>
<td>0.012</td>
</tr>
<tr>
<td>Blood (whole)</td>
<td>0.04</td>
</tr>
<tr>
<td>Blood plasma</td>
<td>0.015</td>
</tr>
<tr>
<td>Gasoline</td>
<td>0.006</td>
</tr>
<tr>
<td>Glycerine</td>
<td>14.9</td>
</tr>
<tr>
<td>Mercury</td>
<td>0.016</td>
</tr>
<tr>
<td>Oil (light)</td>
<td>1.1</td>
</tr>
<tr>
<td>Oil (heavy)</td>
<td>6.6</td>
</tr>
<tr>
<td>Water</td>
<td>0.01</td>
</tr>
</tbody>
</table>

Figure 9: Image depicting some of the variables present in Poiseuille’s Law
### Figure 10: Intravenous Access measurements and descriptions (Morrison et al. 2011)

<table>
<thead>
<tr>
<th>Catheter</th>
<th>Description</th>
<th>Size</th>
</tr>
</thead>
<tbody>
<tr>
<td>Braun* vasofix safety cannula</td>
<td>Standard over the needle cannula</td>
<td>22G 25 mm</td>
</tr>
<tr>
<td>Braun vasofix safety cannula</td>
<td>Standard over the needle cannula</td>
<td>20G 33 mm</td>
</tr>
<tr>
<td>Braun vasofix safety cannula</td>
<td>Standard over the needle cannula</td>
<td>18G 45 mm</td>
</tr>
<tr>
<td>Braun vasofix safety cannula</td>
<td>Standard over the needle cannula</td>
<td>16G 50 mm</td>
</tr>
<tr>
<td>Braun vasofix safety cannula</td>
<td>Standard over the needle cannula</td>
<td>14G 50 mm</td>
</tr>
<tr>
<td>Hospira † ‘Abbocath-T’</td>
<td>Over the needle rapid access catheter</td>
<td>14G 140 mm</td>
</tr>
<tr>
<td>Vygon ‡ ‘Leadercath’</td>
<td>Single lumen Seldinger technique central venous catheter</td>
<td>14G 150 mm</td>
</tr>
<tr>
<td>15 cm triple lumen central venous catheter</td>
<td>Triple lumen Seldinger technique central venous catheter</td>
<td>16G 34 cm distal port</td>
</tr>
<tr>
<td>15 cm triple lumen central venous line</td>
<td>Triple lumen Seldinger technique central venous catheter</td>
<td>18G 35 cm proximal port</td>
</tr>
</tbody>
</table>

### Figure 11: The Ranger: Rapid Infuser (top) and fluid/blood warmer (bottom) (3M 2015)
CITED REFERENCES


Sustainable Batteries

Robert Heil
11/20/2014
Physics 112

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Abstract

The prime objective in this research paper is to design and fabricate a direct current crystal power cell with an infinite lifespan from consumer available products. The research paper will contain an explanation of the physical science as well as a how to guide to duplicate the process. The main objective is to define the physics of a “crystal power cell” fabricate a prototype, measure it’s electrical potential and then allocate what type of residential, commercial, and government applications this type of battery would be useful for.

Sustainable Batteries

Most scientists will agree that renewable sources of energy are the necessary components for the survival of Earth and its inhabitants. Can something as small as a battery be the quintessential factor for Earth’s sustainability in the generations to come? When I first graduated high school in 1992 our disposable society managed to use over 2.2 billion alkaline batteries annually in consumer America alone. That equated to more than 100,000 thousand tons of worthless corroded metal and battery acid after the power cell depleted and the consumer tossed it into the trash. It is also common knowledge that the alkaline battery depletes their electrical potential energy very rapidly even if there just sitting on the shelf and not being used. In reaction to these limited resources everyone has had to replace an alkaline battery or two. In some case a consumer might even have to hire a professional to replace the battery. Either way it is a cyclical night mare just like a repeating decimal that never ends. Engineers swiftly created a renewable battery that could be recharged and reused again with hope for consumer America. Unfortunately these rechargeable batteries such as NiCad or lithium ion batteries have a finite number of recharges and end up in the same location as the alkaline batteries a majority of the time. Now let’s use some imagination and visualize a power cell that never has to be charged, never has any power loss, and never ends up in the ground. “The original inventor of this was Jhon Huchison, a self-taught theoretical physicist who in his spare time conducts home experiments using surplus navy equipment” (The Infinite Battery, 2014). Huchison entitled the power cells “crystal cells” when he first invented this type of battery in the late 1970’s while recreating some of Nicolis Tesla’s work. How does a crystal power cell work and can it realistically be used in lieu of next generation batteries in residential, commercial applications and possibly even NASA?

To completely understand the fundamentals of the crystal battery a simple explanation of the crystal power cells dynamics as well as the basic fundamentals of a standard battery, rechargeable or not, are vital to the comprehension of the physics to this type of battery. Standard batteries must contain three parts, an anode, a cathode and the electrolyte. The cathode and anode are connected up to a simple electrical circuit where the chemical reactions in a standard battery build up electrons at the anode causing an electrical variance between the two. This unstable electrochemical reaction causes the electrons to rearrange to expel the difference in electrical forces. Therefore when the anode and cathode are connected in a closed circuit the electrical potential immediately want to dissipate through the closed circuit.
Unfortunately an unlimited number of electrons ultimately cause the death of the power cell over its short life span.

A crystal power cell has similar electrons that want to escape however; its electrical potential is based on a completely different reaction. First there are two simultaneous basic reactions that power the crystal power cell that should be defined as follows. The first reaction is a physical reaction termed the piezoelectric effect which requires actual physical mechanical forces acting upon the crystals to create an electrical potential. For example when two quartz crystals are squeeze or compress together and electrical reaction occurs and an electrical potential exists. This is why quartz watches never have to be rewound because they are always experiencing an electrical potential caused by the mechanical reaction of the two crystals being compressed together. The secondary reaction that resides in the crystal power cell is technically a chemical reaction which is a variable reaction based upon heat, called the thermoelectric effect. Basically an electrical potential is created specifically by heat being applied to the power cell directly. The thermoelectric reaction will vary because it proportional to how much heat is applied. Subsequently, the two reactions in the expansion of the crystals creates a small amount of thermo energy independently, however if an outside source heats the power cell a positive electrical potential gain will exist even more. These factors could make the crystal power cells more favorable in areas that do not have heat restrictions. All of these physical attributes are the defining characteristics of the crystal power cell and explain the physics behind the battery.

Now that an idea of how the battery produces an electrical potential exist, let’s discuss the design production process more in detail. Remember, sometimes in fabrication stage, modifications might have to be made based on consumer availability and affordability factors. In this particular experiment the process from several engineers was taken into consideration and then a realistic approach for fabrication was implemented. The fabrications number one priority is to be cost effective, however obtainability supersedes.

The initial step in fabricating the crystal power cell is the design process. In this stage locating the raw materials and a calculation of the cost factors are critical before getting started. The diameters of the crucial components of the crystal power cell such as the magnesium core and the copper shell play very important roles in the development of the electrical potential. The perimeter space between the magnesium core and the copper shell is where the piezoelectric effect takes place and electrical potential is created. This space must be adequate enough for your crystal mixture to flow freely during the final phase of fabrication. Technically the greater the area for the chemical compound the greater the area for expansion of the crystals after they have formed. This effect is the same mechanical reaction that happens in the quartzs watch example however its effect is in reverse. Instead of contracting the crystals to produce the electrical potential, now the crystals are expanding against the outer shell generating the electrical potential.

Obtaining the raw materials takes three different merchants to complete the scavenger hunt. The first merchant is available online by the name of Rotometals. The solid magnesium
core can be purchased at a reasonable price while having several different diameters to choose from. A 1.005 inch diameter core that is twelve inches long cost around $15 USD and can be shipped to your doorstep. The 1.0005 diameter magnesium core compliments the 1.5 inch diameter copper caps that were located at Lowes home improvement store because it creates a .25 inch perimeter space around the core for the crystal chemical compound to reside. The 1.5 inch copper shells are located in the plumbing department at almost every hardware store, just be sure to choose the ones that have no corrosion or they will have to be cleaned thoroughly before fabrication. Different inside diameters are available in the copper shells to coincide with the different diameter Magnesium cores but be careful to create a healthy perimeter space between the Magnesium core and the copper shell or problems will occur.

A simple paint stir sick in the paint department makes a great non conducive insulator that will sit snug at the bottom of the copper shell creating a floating effect that is crucial to the fabrication as well. The Magnesium core must be suspended into the copper shell therefore without the insulator this experiment will not work properly. The orange marking on the schematic clearly display that a non conducive substrate must be used as a spacer between the Magnesium core and the copper shell. Plastic cannot be substituted in this instance, because it will melt when the heat is applied in the final step. Once again the object is to suspend the copper shell into the magnesium shell while the crystal compound solidifies around the core. The final merchant with all the missing ingredients for the crystal compound is Wal-Mart. The chemical crystal compound is an equal four way ratio of KAI (SO4)2*12H20, Mg (SO4), KCl (SO4)2*12H20, and Na2B407*10H20. The chemical compounds are commonly known as Alum Powder, Epsom Salt, No-Salt salt substitute, and Borax, the 20 MULE brand. The Alum powder and the No-Salt salt substitute is located in the spice aisle while the Epsom salt is located in the pharmacy section and the Borax is located in the cleaning department with the bleach and the other everyday household cleaning supplies.

The next step is to take the magnesium core to a fabrication shop, display them your schematic and have the core prepped for the anode (-) screw in the top center as well as the final length of the core. This is necessary because the core is extremely ridged and requires precision machine shop equipment to make a precise fine threaded female insert perpendicular to the magnesium medium. There usually will be an initial shop fee for one hour, so use the time wisely and have as many cores prepped for the money being spent. With a female insert the anode (-) terminal screw will be removable for future design features. Two nuts were added onto the terminal for two reasons. First the screw can be secured to the core like a lock washer and the second nut can serve as a tie down to suppress the wire and hold it securely into place. In the end, Vic’s Machine Shop in El Mirage Arizona was able to segment two pieces off the magnesium core measuring 11/16 in length, each with an individual 5/16 female fine tread tap ready to be equipped with a screw with two nuts to serve as the anode in the power cell. Once again, use what is available for the anode screw; just try to use a metric fine tread for maximum grip.
The final phase of fabrication will be mostly chemistry and requires delicate handling due to the high temperatures that will be applied to the crystal battery. First the proper mixture of all four compounds in equal parts is necessary for the solidification process. Use a coffee grinder or a small blender to thoroughly mix the compounds together. Start with one compound, usually Mg (So4) and grind the material in the finest granules possible. Add the next compound, the KAI (SO4)2*12H2O and do the same making sure the compounds are mixed though and a free from chunks. Next add the remaining equal parts compounds of KCI (SO4)2*12H2O and Na2B4O7*10H2O and watch the coagulation begin. The compounds should become thick, coagulated and stick to the spoon almost like peanut butter. Be careful not to seize the motor while mixing the compound because the coagulation of the four compounds restricts the blades from turning properly. Now that are compound is mixed meticulously, the crystal battery is ready for the chemical process to activate the electrical potential in the power cell.

Now the crystal compound must be inserted into the perimeter space between the magnesium core and copper shell. Use the compound conservatively trying not to use too much. In fact try to leave a 1/4” to 3/16” space from the top of the copper shell. Turn the stove or burner to low-medium heat and place a flat pan or a 2 quart sauce pan on to the surface. Next place one of the shells that have been prepped with a magnesium core and crystal compound making sure that the SKU sticker has been removed from the bottom of the copper shell. As the heat dissipates through the shell it starts to heat the crystal compound and it starts to liquefy. Use a toothpick and carefully stir the flux as the melted compound is heated and liquefied. Next, turn off the stove or burner and use thongs to remove the battery from the pan. Place it an area that will allow it to cool where testing can be started.

Use a digital multi-meter to measure the electrical potential across the battery carefully touching the black lead to the anode (-) and the red lead to the cathode (+). In this instance the screw is the cathode and the outer encasement or the lip of the copper shell. The digital multi-meter reads 1.30 volts for the first battery and 1.28v for the second battery. Now use a wire as a jumper and connect the two power cell together in a series fashion. When connected in series the multi-meter measures 2.51v. This is equivalent to any standard rechargeable 1.5 volt AA, AAA, C or D cells on the market today as well as any standard sized 1.5 volt alkaline battery in the same referenced sizes. Only time will tell if the crystal battery will remain with an infinite electrical potential but after a month the observer will know right away by the voltage readouts if the cell is sustainable.

After comprehending a portion of the physical science behind the power cell, the discussion now lies on where to use these types of power cells in residential, commercial, and government applications. Imagine the consumer having the opportunity to purchase an electronic device with a power source that never had to be changed and never had to be recharged with an outside power source. Firstly, prioritizing the electronic devices that need these types of power cells is essential. Let’s be realistic, research and development would not want to spend millions of dollars on a battery for a children’s toy that would discarded after someone spilt juice on it. In a disposable society, an engineer’s battle must be chosen wisely
for the crystal power cells proper development into mainstream society. If the right devise was equipped with an infinite power cell and can give consumers dollars and sense back as well as slowing down the volatile trash build up on our planet, then grounds lie for more research and development. Putting the environment aside, it is common sense for consumers that the crystal battery will never meet mainstream production if it is not cost effective.

Primarily, the devise that initially comes to mind is the simple 9 volt smoke detector or equivalent Co2 detector. Both devises possess the same technology, except they possess different internal sensors that register different types of smoke. Smoke detectors and Co2 detectors are both wired in series and require a battery backup to properly function in case of total power loss. This type of devise constantly needs the power cell changed out before the devise need to be changed out. In fact most commercial building will require a fire protection agent to resolve the issue for the owner. Crystal power cell will save the environment trash as well as saving labor cost for service technicians. Indeed the smoke detector and the Co2 detector are excellent apparatuses for infinite power cells for today’s construction and home improvement markets and would serve as an a great test market to introduce this type of power cell.

Secondly third world countries throughout the globe are in dire need of humanitarian assistance in isolated remote locations. Due to infectious disease, like Ebola, wild carnivorous animals and local insurgents in these unpredictable remote areas, the requirement for delicate and precise timings when establishing these types of wireless communications are desired. These low voltage routers and repeaters allow for transmission of communications for doctors, military intelligence, and humanitarian efforts. Once these hot spots are initialized the technicians do not want to come back just to change its power cell in 45 days. The solar power option remains obsolete because of the absences of light 10 to twelve hours a day and the fiber optics require the cable to be buried because it is too fragile. A wireless repeater with a crystal power cell remains the prime candidate for wireless communication in remote locations for more reliable data transmission and less technical support.

Last but not least there is the government NASA application for the sustainable battery. The international space station is a self-sustaining entity that solely relies on the sun for power. The ISS’s 160 volt electrical system requires the use of direct current collectors for when the space station reaches voids of light in its orbit. Basically in simple terms the ISS switches to battery backup when it is not directly in view of the sun. All of these computer and electrical systems run off of one central computer. The computer is the central intelligent system that drives the internal workings for the entire ISS. The ISS’s computer system is a prime candidate for a power cell with infinite electrical potential because of this next factor. It takes minor amounts of voltage to secure the memory for the motherboard on this type of system and with the right capacitors and a dozen of the crystal cells in series, the crystal power cell could power a computer main frame system to keep the ISS afloat on power cell that never needs recharging. Surely NASA has it covered by now; nonetheless the crystal battery fits the model perfectly in this sci-fi scenario like a missing puzzle piece.
In conclusion discovering different methods of infinite electrical potential will definitely be a commodity in our futuristic world for our children and generations to come. Every day as we know it scientist and engineers are working on the perfect power cell that could change our world forever. “More than 2,000 organizations throughout the world are actively involved in fuel cell development” (The Future Battery, 2003). Using the crystal battery model described throughout this research paper is an excellent leading step in achieving our goals as a sustainable society that never wastes what it needs for everyday existence. Reinventing the battery will not change the way we live as a cultural society but it just might possibly change the outcome of our planet. Realistically the crystal power cell prototype designed and fabricated in this experiment is bulky and too big to fit into most of our everyday electronic devices. Subsequently, this deduction does not mean that the crystal power cell is obsolete; this just means that more intelligent engineers need to be working on this type of project and perhaps a scientific breakthrough could be made. Meanwhile, we can all use NiCad or lithium ion rechargeable batteries in lieu of alkaline batteries while we wait for the research and development of the crystal lifelong power cell that will retrofit into our life as we know it.

Schematic of Prototype

The core is solid magnesium with a solid corrosion free copper shell that serves as the cathode (+). The spacer is wood and can be substituted with plastic. The anode (-) is a metric M7 fine threaded screw that can be adjusted in length appropriately. The double nuts were added to secure the screw to the magnesium core plus a wire when connect into a circuit. The crystal compound mixture that fills the copper shell are four equal parts of KAI (SO4)2*12H20, Mg (So4), KCl (So4)2*12H20 and Na2B407*10H20, common household products branded Alum powder, Epsom Salt, No Salt substitute, and Borax…20 MULE brand.
Schematic is not to scale.

Raw Materials With Fabricated Core
Live Prototype I
References

Chemistry, Cavities and Crowns:
Studying the chemical processes and materials in oral health

By

Matthew Jones

April 27th, 2015
Chemistry 152
Dr. Scott Massey
Abstract: The study of chemistry helps provide insight about the processes that take place in our mouth. Dental enamel itself is an inorganic compound that is only produced once during a lifetime and can decay in acidic environments. Chemistry also helps to select biocompatible materials used on oral tissue to treat decay and abrasion that occurs on teeth. This paper explores the interdisciplinary relationship between chemistry and dentistry.

Chemistry has many practical applications to dentistry. It is used to show how soft drinks and acidic foods can erode healthy tooth enamel. The chemical importance of fluoride was studied when countries began fluoridating the public water supply to prevent tooth decay. Restorative materials have been carefully selected based on their physical and chemical properties. Mercury amalgam is one material that has come under controversy for its possible toxic effects. Noble metals have been used for since the turn of the twentieth century as restorative materials for cavities and prostheses. Dentists and scientists had to use chemistry to come up with solutions for the treatment of cavities and tooth loss. This paper examines the role of chemistry in the dental profession by examining the basic chemical processes that dentists use in educating and treating patients.

Chemical Composition

The physical and chemical composition of human teeth and how chemical changes can erode enamel are important to explain how cavities are formed. A tooth is divided into two parts, the crown and the root. The crown is the exposed part of the tooth that is covered with enamel. Tooth enamel is an inorganic compound made up of hydroxyapatite, \( \text{Ca}_10(\text{PO}_4)_6(\text{OH})_2 \) or \( \text{Ca}_5(\text{PO}_4)_3\text{OH} \) (Aoba, 1997). Enamel the hardest structure in the human body. Dentin is found underneath the enamel layer and is makes up the bulk structure of the tooth. Gingiva is the pink tissue found between the root and crown structures of the tooth. The gingiva acts as a barrier to prevent bacteria from invading the space below the surface.

Solubility

The tooth surface is continuously changing as the concentration of calcium and phosphate ions are altered. These minerals are found naturally in our mouth and in our saliva. A decrease in the concentration of these ions will eventually demineralize the surface of the tooth. The demineralization causes the structure to weaken and decay which then leads to cavities. This process can be prevented by maintaining the concentration of these ions in the mouth and in the saliva. The precipitation and dissolution of hydroxyapatite can be expressed as:

\[
\text{Solid Forms} \quad \text{Enamel Dissolves}
\begin{align*}
\text{Ca}_10(\text{PO}_4)_6(\text{OH})_2 & \iff 10 \text{ Ca}^{2+} + 6 \text{ PO}_4^{3-} + 2 \text{ OH}^- \\
\end{align*}
\]

These two opposing reactions occur simultaneously and they reach chemical equilibrium when the rate that the minerals dissolve is equal to the rate they precipitate.
The solubility product, or $K_{sp}$, for the saturated enamel is $K_{sp} = [\text{Ca}^{10+}] [\text{PO}_4^{6-}] [\text{OH}^2]$. The solubility product value for enamel is $6.4 \times 10^{-58}$ (Kirkham, 2000).

Tooth enamel is saturated at a pH of 5.5 (Dawes, 2003). Acidic solutions will decrease the concentration of hydroxide and phosphate ions in the aqueous solution. When the concentrations of these ions are less than the $K_{sp}$ value, the solution is unsaturated. In a study of enamel mineral loss, West and Joiner show that a decrease in pH will increase the hydroxyapatite dissolution rate (West, 2014). This process is expressed in the equilibrium equation:

$$\text{Ca}_{10} (\text{PO}_4)_6 (\text{OH})_2 (s) + 8 \text{H}^+ (aq) \leftrightarrow 10 \text{Ca}^{2+} (aq) + 6 \text{HPO}_4^{2-} (aq) + 2 \text{H}_2 \text{O}$$

Dental plaque has a neutral pH and contains low concentrations of phosphate and calcium ions. Acids are produced from bacteria in plaque as sugary foods are metabolized. This process creates a more acidic environment in the mouth as these acids interact with the hydroxyl group of hydroxyapatite. Le Chatelier’s Principle helps to explain that acids (H+) will consume the products causing the structure to dissolve as calcium and phosphate ions are released.

Soft drinks and other acidic foods are often linked to cavities and tooth decay. A study from the British Journal of Nutrition examined the harmful components of soft drinks (Luzzi, 2012)

<table>
<thead>
<tr>
<th>Acidic Food</th>
<th>pH</th>
<th>[Ca] (mmol/l)</th>
<th>[P] (mmol/l)</th>
<th>[F] (mmol/l)</th>
</tr>
</thead>
<tbody>
<tr>
<td>Coca-Cola</td>
<td>2.45</td>
<td>1.08</td>
<td>5.04</td>
<td>0.22</td>
</tr>
<tr>
<td>Pepsi</td>
<td>2.39</td>
<td>0.33</td>
<td>4.93</td>
<td>0.04</td>
</tr>
<tr>
<td>Gatorade</td>
<td>3.17</td>
<td>0.13</td>
<td>2.98</td>
<td>0.05</td>
</tr>
<tr>
<td>Red Bull</td>
<td>3.30</td>
<td>1.94</td>
<td>&lt;0.01</td>
<td>0.11</td>
</tr>
<tr>
<td>Red Wine</td>
<td>3.43</td>
<td>1.25</td>
<td>4.69</td>
<td>0.07</td>
</tr>
<tr>
<td>Yogurt</td>
<td>4.03</td>
<td>56.33</td>
<td>38.74</td>
<td>0.03</td>
</tr>
<tr>
<td>Beer</td>
<td>4.20</td>
<td>0.74</td>
<td>5.65</td>
<td>0.74</td>
</tr>
<tr>
<td>Coffee</td>
<td>5.82</td>
<td>0.69</td>
<td>0.63</td>
<td>0.07</td>
</tr>
<tr>
<td>Black Tea</td>
<td>6.59</td>
<td>1.10</td>
<td>0.27</td>
<td>1.63</td>
</tr>
</tbody>
</table>

Soft drinks and energy drinks have pH values lower than tooth enamel’s critical pH of 5.5. As the pH drops, the concentrations of calcium, phosphate and fluoride ions also decline. When these substances come in contact with teeth they place a stress on the equilibrium balance causing the enamel to dissolve into ions. Yogurt is seen as less harmful than carbonated beverages because it contains a high concentration of calcium to maintain equilibrium. Coffee has a pH of 5.82, which is higher than the saturated pH of hydroxyapatite so it’s not considered erosive to enamel.

Acidic foods and drinks are not the only substances that can damage teeth. Gastric acids are composed of very acidic gastric juices. These juices are produced in the cells of the stomach during digestion. One of the main components of gastric juice is hydrochloric acid (approx. pH = 1.0). The backflow of these juices into the mouth can cause dental erosion. Gastro-esophageal reflux disease (GERD) and bulimia are conditions associated with the backflow of these juices to the mouth. Patients suffering
from bulimia are more likely to experience dental erosion as compared to non-bulimic patients (Ren, 2011). These acids come in contact with the posterior side of the teeth and are more likely to erode that surface first.

**Fluoride**

In the 1930s, differences in the number of cavities were noticed among communities that had different levels of fluoride in their drinking water (Featherstone, 1991). Since then countries have been adding fluoride to the water supply as one way to prevent cavities. Fluoride is the anion of the chemical element fluorine, expressed as F\(^{-}\). It takes the place of the hydroxide ion (OH\(^{-}\)) in calcium hydroxyapatite, Ca\(_5\)(PO\(_4\))\(_3\)OH, to form fluorapatite, Ca\(_5\)(PO\(_4\))\(_3\)F (Kirkham, 2000).

The incorporation of fluoride lowers the solubility product to \(K_{sp} = 1.0 \times 10^{-60}\). This occurs as the fluorine attracts adjacent ions to create a denser crystalline structure. This chemical change makes hydroxyapatite more resistant to dissolving and releasing ions despite experiments that show fluoride doesn’t fully integrate into the mineral structure of enamel (Aoba, 1997). The chemical equilibrium of fluorapatite can be expressed as:

\[
\text{Ca}_5(\text{PO}_4)_3\text{F} \leftrightarrow 5 \text{Ca}^{2+} + 3 \text{PO}_4^{3-} + \text{F}^{-}
\]

Fluoride is still considered the cornerstone of cavity prevent in spite of the sophisticated advances in dental materials. For example, sodium fluoride is still a central ingredient in toothpaste and mouthwash. Fluoride has also been added to chewing gum so it is released into the saliva during chewing. Now scientists are finding ways to incorporate fluoride into dental cements and restorative appliances.

**Restorations**

Dental restorations were introduced to treat the cavities and decay resulting from acidic chemical reactions harmful to teeth. The bacterium, *Streptococcus Mutans*, is associated with the release of acid in the mouth and the formation of dental plaque. Dental plaque can stick in the grooves between the gums and the tooth surface to destroy the dentin layer underneath the enamel. This bacterium reacts with sucrose to produce lactic acid.

There are two types of indirect restorations that can be created in a dental lab. Full dentures and partial dentures are examples of removable restorations. Crowns and bridges are referred to as fixed restorations. They are fixed because a dentist will cement the prostheses over the existing tooth structure. Base metals are used to create the framework for dentures but their presence is mostly noticed in fixed restorations.

Dentists have two types of restorations at their dispose to treat dental caries. A direct restoration involves a filling material being applied to the surface of the tooth to combat superficial decay. The most common filling materials are mercury amalgam and composite resin. Indirect restorations are created outside the dental office at the dental lab. Dental technicians use impressions to build a model of the patient’s mouth in order to construct prostheses to replace missing or damaged teeth. These types of restorations are more common when significant decay has occurred.
Mercury Amalgam Fillings

Mercury amalgam dental fillings have been widely used by dentists for over a hundred years. The most common amalgam is formed with silver, tin and mercury, which is liquid at room temperature. This mixture creates a putty used to repair superficial cavities. Once it has been applied then it transforms into a hard solid as it’s cured. Mercury amalgam fillings can be identified in the mouth from their silvery metallic color. Mercury amalgam has the advantages of being strong, durable and inexpensive.

The use of mercury has been questioned in recent decades because of concerns that mercury will leak into the body causing harmful toxic effects. In 2007, the city of Philadelphia passed a law requiring that patients be provided fact sheets showing the health risks associated with mercury amalgam (McGrath 2013). They have also been banned in Norway and Denmark. In 2009, the FDA reclassified mercury from a class I (least risk) device to class II (more disk) device (Williams, 2010).

The FDA still considers dental amalgam fillings safe for adults and children ages 6 and above. Their evidence shows a low amount of mercury in the bodies of people with these fillings. This low amount of mercury is not considered to have a harmful effect. Mayo Clinic, American Cancer Society, Alzheimer’s Association and other outside organizations have endorsed the use of mercury in dental amalgam because they have not seen evidence of its adverse health effects. These positions may change in the future as more long term research is still being conducted.

Investment and Casting

The process of making metal dental crowns takes its roots from metalworking, but before a crown can be cast a model is created from a mold of a patient’s mouth. Hydrocolloids and synthetic elastomeric polymers are materials found in elastic impression materials. Alginate impression material is used because the change from a more liquid gel to a solid is irreversible. Powder and water are mixed together in proper ratio to create a paste that can record the teeth and gum tissue. The alginate powder is composed of calcium sulfate dihydrate, soluble alginate, and sodium phosphate (Powers, 2006). The following chemical reaction occurs when water is added to the powder:

$$\text{Ca}^{2+}_{(aq)} + \text{Alginate}_{(aq)} \leftrightarrow \text{Ca} - \text{Alginate}^+$$

Alginate impressions lose water by evaporation and therefore distort easier than the newer synthetic materials composed of polysulfides, silicones or polyethers. Polysulfide impression materials are supplied as two pastes in collapsible tubes. One tube contains a base component such as polysulfide polymer and the other contains a catalyst like lead dioxide. This exothermic reaction produces a rubber like material once it sets after approximately 10 minutes (Powers, 2006). A dental plaster is then poured into the impression material to create a model of a patient mouth and occlusion.

Calcium sulfate dihydrate (CaSO₄·2H₂O), gypsum, is usually used to make a cast of the oral cavity from the dental impression. Gypsum products are supplied as a powder material with similar characteristics to plaster. The setting of gypsum involves the chemical reaction of 1 g mol plaster with 1.5 g mol of water to produce 1 g mol of gypsum material (Powers, 2006). This reaction can also be expressed in grams where 100
grams of plaster is mixed with 20 grams of water to produce 120 grams of calcium sulfate dihydrate.

Wax is then applied to the gypsum model once it has been mounted with an accurate occlusion. Plant and animal waxes used in dentistry contain acids, alcohols and hydrocarbons. Synthetic waxes consisting of glycerol and fatty alcohols are refined through chemical processes in a laboratory for use in dentistry. The melted wax is applied to the area of the model where the crown or bridge will be placed.

Once a wax pattern of the restoration has been created then it is invested in a ceramic material. Investing involves surrounding the wax pattern that can duplicate is shape. The next step is to remove the wax from the investment by burning it out. This creates a mold for the molten metal to flow. Investment materials are composed of gypsum and phosphate materials because they can withstand the heat of the liquid metal. Casting machines are used to process the molten metal into the mold. The type of casting machine used depends upon the type of alloy being cast as different metals have different melting properties.

Crows

Noble metal alloys have been used to cast restorations since the investing and casting procedures were developed in the early 1900s. Alloys of noble metals contain gold, platinum and palladium. Noble metals can sometimes be referred to as precious metals but the two terms have different meanings. The term “noble” refers to the chemical behavior of the metal. The term “precious” refers to the cost of the metal. Gold and platinum are both noble and precious metals. Palladium is noble, but less expensive and is therefore not considered precious (Shillingburg, 1997) The American Dental Association classifies alloys based on their noble metal content.

<table>
<thead>
<tr>
<th>ADA Classification System for Alloys for Fixed Prosthodontics</th>
</tr>
</thead>
<tbody>
<tr>
<td>High Noble Alloys</td>
</tr>
<tr>
<td>Noble Alloys</td>
</tr>
<tr>
<td>Base Metals</td>
</tr>
</tbody>
</table>

*NThe platinum group contains the following transition metals: ruthenium, rhodium, palladium, osmium, iridium, and platinum.*

Noble metals alloy restorations were common from the 1934 to 1968. It was during this time that the price of gold was set at $35 per ounce. The price of gold began to increase after this price control ended which meant that noble metal crowns became a more expensive treatment option. Gold alloys are strong, resistant to corrosion, resistant to corrosion, and biocompatible. Silver-palladium alloys melt at temperatures above 1090°C. The disadvantage of gold alloys is the cost. The properties of noble alloys are similar to high noble alloys. However, lower noble metal content is associated with greater corrosion. Base metal alloys have been used to counter the rise in cost of noble metals.
Base metals such as nickel, beryllium, copper and silicon are used in alloys to create less expensive restorative options. Base metals were used in the metal framework for dentures before they began being used for single tooth restorations. Nickel-chromium alloys are the most popular base metal crowns. They have the advantages of high strength and low cost. There are some biocompatibility concerns with their use as some patients experience an allergic reaction upon exposure to nickel. Beryllium is a possible carcinogen that can be hazardous to dental laboratory technicians when inhaled. These alloys have melting temperatures that range from 1260 to 1430 °C (Shillingburg, 1997).

Titanium would be a great material to use for crowns because of its biocompatibility. It’s currently used in dentistry for dental implants. Titanium is also used in other medical procedures like hip and knee replacements. Titanium melts at temperatures above 1668°C. This high melting point makes it less desirable, as expensive casting equipment must be purchased.

All porcelain crowns are being used more frequently as a response to the high cost of gold alloys. These ceramic crowns and bridges share the physical properties similar to glass so they contain no metal. The appearance is very similar to a natural tooth because there is no metal structure to limit their translucency. Ceramic crowns are primarily composed of zirconium oxide (ZrO2). This material is not as strong as the metal-based restorations which means they can fracture. To prevent against fracture, the ceramic crowns must be made thicker than their metal alternatives. This means that more of the underlying tooth structure must be reduced. This is a concern because enamel should be preserved since it is the harder and more durable than the ceramic.

Ceramic crowns will become more common in future years because they are easily produced and inexpensive. Ceramic crowns are milled by machines previously used to manufacture auto parts. This replaces the detailed process of casting alloys from impressions and models. Intraoral scanners are used to take digital impressions. The impression file is then sent to a lab technician who designs the crown using 3D computer models. These digital impressions are more accurate because they don’t distort or expand like the elastic impression materials.

Conclusion

It’s impossible to completely avoid contact with all the acidic substances that can damage our teeth. The inorganic composition of dental enamel highlights the important role that chemistry plays to maintain healthy teeth and gums. It’s the responsibility of the dental team to educate patients on how to prevent acidic chemicals from destroying their teeth. This means that dentists and other dental professionals need to be aware of science behind concepts like solubility and pH. I believe that patients that are educated about the chemical reactions that take place in their mouth are more likely to keep their natural teeth. Processes have been developed to create prosthetics with similar physical characteristics to our natural teeth. These prosthetics are valued for their biocompatibility in the oral environment, but so far nothing has come to completely replace all the characteristics of our natural enamel.


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The Electromagnetic Spectrum

An Analysis of Different Light Energies

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The electromagnetic spectrum encompasses all invisible and visible light in the universe. Light can be categorized into seven different wavelengths: radio waves, microwaves, infrared light, visible light, ultraviolet light, x-rays, and gamma rays. Each wavelength has special properties and characteristics that allow researchers to learn more about distant celestial bodies.

The universe is unfathomably vast. It contains countless stars and galaxies, remains in constant motion, and houses incredible nuclear events. Unfortunately, most of this activity is invisible to the human eye, which can only see a miniscule fraction of existing light. It is difficult to believe that there is something beyond human perception, but the electromagnetic spectrum has a unique way of playing tricks with the human mind, which relies on vision to affirm an object’s presence. In order to conduct research on different phenomena, astronomers have had to turn to special instruments that allow the visual construction of celestial bodies. Technologies such as satellites and telescopes capture focused images of certain locations in the universe and give invaluable information about the composition, temperature, speed, and formation of other worlds. The electromagnetic spectrum, which can be categorized into seven types of light, provide scientists with detailed data that increases understanding of the universe.

To many people, the concept of light is limited to visible wavelengths, such as the colors green, orange, yellow, and blue. In fact, visible light is only a small piece of the entire electromagnetic spectrum, which also includes radio waves, microwaves, infrared light, ultraviolet light, x-rays, and gamma rays. Although most of the spectrum is not perceivable, it surrounds every free space in the universe. Light is everywhere, but the problem lies not with human eyes, but rather the composition of Earth’s atmosphere. According to the National Aeronautics and Space Administration, gases such as carbon dioxide, ozone, and water vapor absorb or reflect most electromagnetic waves (NASA). Visible light, radio waves, and some ultraviolet, infrared light, and microwaves can penetrate the Earth’s atmosphere, and are referred to as atmospheric windows (NASA). It is for this reason that special instruments are created and launched into space; light that is blocked by the earth’s atmosphere can be detected and studied in outer space. In order to understand how the electromagnetic spectrum helps scientists predict the structure and molecular makeup of different objects and galaxies, it’s important to first understand the nature of each individual light energy.

Dr. Liew Soo Chin, principal research scientist and head of the Research Centre for Remote Imaging, Sensing and Processing (CRISP) at the National University of Singapore, describes light wavelengths as “energy transported through space in the form of periodic disturbances of electric and magnetic fields” (Soo Chin). All light travels at the same speed, $2.99792458 \times 10^8$ m/s, commonly referred to as “the speed of light,” so it is not speed which separates the seven kinds of energies (Soo Chin). The main characteristic that distinguishes the wavelengths is frequency, or the number of waves that pass during a given amount of time. According to Soo Chin, “frequency... depends on its source. There is a wide range of frequency encountered in our physical world, ranging from the low frequency of the [electromagnetic] waves generated by the power transmission lines to the very high frequency of the gamma rays originating from the atomic nuclei,” (Soo Chin). It is frequency that makes the electromagnetic spectrum a spectrum of waves.
Radio waves have the lowest frequency and wavelength in the electromagnetic spectrum. Wave length can range from 1 millimeter to 100 kilometers, and require very large radio telescopes-- expanding many meter across-- to capture clear radio images (Soo Chin). Although radio waves are most commonly associated with broadcasting communication, they also have a variety of other uses. In the field of astronomy, radio waves can be utilized to record the activity of magnetic fields, which often emit energy in the form radio waves, from objects such as the Sun and Jupiter (NASA).

The second type of wave in the electromagnetic spectrum is microwaves. Due to the similar nature between microwaves and radio waves, some scientists prefer categorizing microwaves as a type of high-frequency radio waves. According to Paul Baumann, a geography professor at the State University of New York, microwaves are especially useful because they can penetrate through clouds, fog, snow, and rain (Baumann). As a result of this special characteristic, microwaves are great tools for capturing images of Earth’s meteorological and geological processes. There are two types of instruments that can give astronomers a clearer view of weather and land masses: active remote sensing and passive remote sensing. The only difference between the two technologies is the function of microwaves (Baumann). Whereas active remote sensing emits microwaves to gather information, passive remote sensing only collects emitted microwaves (Baumann). Despite their differences, both instruments provide great amounts of information about Earth’s fickle, and ever-changing nature.

Infrared light, the energy with the third lowest frequency in the electromagnetic spectrum, is unique in its ability to sense light and take images that would otherwise be obstructed by gases or dust (NASA). As stated by Chris Woodford, author of the widely popular book series “Cool Stuff and How it Works,” infrared light help astronomers analyze temperature, composition, and structure of different celestial objects, such as the cloudy Carina Nebula (Woodford). Metaphorically, infrared light can be seen as the vacuum cleaner of the universe; it helps clean up the skies so that scientists may learn more about the universe.

Visible light is the most studied and observed type of electromagnetic energy; light is simply the most accessible, least dangerous, and explored wavelength. Different colors can be explained by the length of each visible light wave: Red: 610 - 700 nm, Orange: 590 - 610 nm, Yellow: 570 - 590 nm, Green: 500 - 570 nm, Blue: 450 - 500 nm, Indigo: 430 - 450 nm, Violet: 400 - 430 nm (Soo Chin).

Past visible light, as the frequency of each wavelength rises and the electromagnetic spectrum becomes dangerous to humans and other life forms on Earth. Most ultraviolet light is blocked or absorbed by the Earth’s atmosphere, but a small percentage still manages to pass (Yale). According to the British Broadcasting Corporation’s science website “GCSE Bitesize,” “Ultraviolet radiation is found naturally in sunlight. We cannot see or feel ultraviolet radiation, but our skin responds to it by turning darker. This happens as our bodies attempt to reduce the amount of ultraviolet radiation reaching deeper skin tissues. Darker skins absorb more ultraviolet light, so less ultraviolet radiation reaches the deeper tissues. This is important, because ultraviolet radiation can cause normal cells to become cancerous” (BBC). Despite these effects, from an astronomical perspective, ultraviolet radiation is not so bad because it helps researchers detect young stars such as galaxy M81, an interstellar nursery.
Due to their high frequency, X-ray and Gamma rays are the most dangerous forms of light energy in the universe. According to professor S. Farooq, a geology department faculty member from the Aligarh Muslim University, the type of objects that reflect these two kinds of wavelengths possess incredibly high temperatures and can be relatively unstable (S. Farooq). Some examples include the Sun’s corona, supernova a nuclear explosions, and neutron stars (NASA).

The electromagnetic spectrum provides a boundless amount of information about the universe and every system, galaxy, and star that makes a small part of the whole. Although it may seem that there is little left to be discovered about the electromagnetic spectrum, recent advances in technology has allowed for the growth of various sectors in astronomy. Increased interest and a desire to continue exploring the numerous possible uses of the electromagnetic spectrum is sure to nurture ingenuity and usher in a new era of astronomical research.
References:


http://www.crisp.nus.edu.sg/~research/tutorial/em.htm

The following source, featured on the official website of The Centre for Remote Imaging, Sensing and Processing at the National University of Singapore, elaborates the primary concepts related to the electromagnetic spectrum. Numerical data is given to differentiate the various types of waves. Definitions for electromagnetic spectrum terminology such as frequency and wavelength, are provided.


Paul Baumann, professor of geography at the State University of New York, provides a brief description of the relevance of the electromagnetic spectrum to the development and use modern scientific instruments. This source primarily focuses on different technology and applications, but addresses the significance of understanding the types of wavelength bands and how they relate to capturing images.


Retrieved from Aligarh Muslim University website: http://www.geol-amu.org/notes/m1r-1-1.htm
This source was retrieved from a website hosted by S. Farooq, geology department professor at Aligarh Muslim University in India. The website contains lectures notes from a select number of classes which he teaches. One lesson expands understanding of the electromagnetic spectrum, remote sensing, and includes a table for categorizing electromagnetic radiation (wavelength and frequency included in inconsistent units).


This source was published on Yale University’s official website in 2008. It contains thorough descriptions of the properties of each type of wavelength in the electromagnetic spectrum. This research focuses on remote sensing and geographic information systems, but also offers information on the physical and supported numeral differences between various electromagnetic waves. A list of common terminology is listed and important distinctions made, such as for spectral reflectance and spectral resolution.


The following source was published Chris Woodford, the award-winning British science writer of the highly acclaimed how-it-works series: Cool Stuff and How It Works, Cool Stuff 2.0, and Cool Stuff Exploded. Woodford describes the electromagnetic spectrum in
terms of daily relevance, wave variations and possible dangers. This source also includes a condensed history/discovery of the electromagnetic spectrum.

British Broadcasting Corporation, GCSE Bitesize. (2014). *The Electromagnetic Spectrum*

Retrieved January 27, 2015, GCSE Bitesize:


BBC’s GCSE (General Certificate of Secondary Education) Bitesize official website offers a five page unit on the electromagnetic spectrum. What sets this source apart from the rest is that it provides a list of the hazardous effects of radiation as well as the various types of ionising radiation. Also, each electromagnetic wave description contains modern uses/ everyday points of contact.


*Introduction to The Electromagnetic Spectrum.* Retrieved from NASA Mission:Science:

http://missionscience.nasa.gov/ems/01_intro.html

The official website of the National Aeronautics and Space Administration, Mission: Science, offers a thorough, topic-by-topic guide to building a fundamental understanding of the electromagnetic spectrum. Apart from offering a visual and interactive unit on each type of electromagnetic wavelength, this source also offers the following complimentary units: “Wave Behavior,” “Anatomy of Electromagnetic Waves,” and “The Earth’s Radiation Budget.”

This article, written by the cited researchers from the Institute of Environmental Medicine in Stockholm, Sweden, was published in the British Medical, and expounds the possible dangers associated with exposure to varying electromagnetic bands. The article primarily focuses on the health risks of radiation, especially to children, and the arguable associations of electromagnetic radiation to cancers.
Abstract

This paper reviews the controversial topic of touch. Humans may think that everything that they come in contact with, they touch it physically. Is this true? Can we touch? We may get the sensation that we are in fact touching objects, people, etc. In all reality we have never touch anything in our lives, ever. There is a very scientific reasoning to this whole concept. At the atomic level you get very, very close to touching objects, but in fact you never do. The electric repulsion due to the electrons repelling each other is about 10E-8 meters. The closest possibility to touch is chemical reactions, such as a child in a mother’s womb.

Information

We are going to be talking about a very interesting topic today. It might make some feel some degree of depression, and alienation so beware before you read. One of the best senses that the human body has is the sense of touch. Sense of touch is incredible, take for example, reach into your pocket where you have coins in there. Say you have one of each coin currency, (quarter, dime, nickel, and penny.) Without even looking at the coins you can tell what coin is which. The sense of touch is an amazing tool that humans possess. Touch can even be lifesaving. Say perhaps you’re falling off a cliff. What’re you going to do? You need to try to grasp onto something to save your life. So you are going to reach and touch a branch if you will, you are going to grip it and attempt to pull yourself to safety. So touch can even be a survival mechanism. It goes to even simple measures as well. Such as intimacy, eating foods, drinking water, etc. Let’s start by posing a question, can you touch anything? Before we just straight up answer the question, what could be the factors that could affect/insure touch? We have to take into consideration electrons and their principles, pauli exclusion principle, human nervous system, electric repulsion, and atoms.

So can we humans touch anything? The answer is no, we have actually never touched anything in our life! So how can this be possible, you feel as if you’re making contact with everything you grab, press on, etc. You get extremely close, so close to touching objects. Although at the atomic level we don’t actually touch anything. We get as close as 10E-8
meters away from actually making contact with anything we encounter in life. Matter is made up of these tiny microscopic particles called atoms. You cannot see these particles with the naked human eye. So matter, being anything in the world, human bodies are made up of atoms. An atom is comprised of even smaller particles. These particles are neutrons, protons, and electrons. You have to first understand how these atoms are comprised. The core is the nucleus which contains the protons and neutrons. Outside of the nucleus are the electrons. These electrons orbit the nucleus in an electron cloud.

This picture here depicts the atom comprised of the smaller particles.

Electrons have a negative charge (-), protons have a positive charge (+), and a neutron is neutral having no charge. Now that we know how the atom is layed out, we know that the outermost particle is an electron. So since matter is made up of atoms, they naturally repel from each other, this is called electric repulsion. This happens because the like charges repel, so the electrons being the outermost part is doing the repelling. Our bodies have these particles as well as every object. So we pick of an object, but in all reality we aren’t really touching it at all. Take for example, when you’re sitting in a chair, you would think for sure you’re making contact with the chair. You can feel the chair putting pressure on your butt. What is actually happening, is that you’re hovering over the chair at pretty much an unmeasureable about of 10E-8 meters.
This leads to the next case of proof that we never actually directly touch anything. Another concept that goes hand in hand with the electric repulsion theory. Electrons naturally repel from each other because they’re like charges. Electrons are smarter than we think. Electrons actually know where every other electron is in space. The electrons do their best to avoid each other and that is why no two electrons can be on the same level with each other. This backs up the theory even stronger that we never actually touch anything in our lives, we merely just hover over each other, objects, etc.

So if this is true, and we never actually touch anything in our lives, how do we feel like we’re touching? When we hug some one we feel there body against ours, when we engage in sexual intercourse it is one of the best feelings ever, when we grab food and put it in our mouth.. we experience touch, right? The human nervous system is actually playing a trick on you. The human body/brian is comprised of nerve cells. The brain sends motory signals to the body, in return the body sends sensory signals to the brain. Motory signals simply tells the body to contract skeltal muscle. The sensory signals can be pain, tempertature, pressure, touch, etc. So why does our body tell our brain that we touch? Say for example you take your index finger and your thumb, press it together as tightly as possible. You would say without a doubt in your mind that the two are touching each other, you feel the pressure from it, it has to be true. What is actually happeing is that yes you’re feeling that,
what it is though are the electrons repelling pushing on your pressure receptors telling your brain that something is causing pressure.

So we all know about these things called elements, and they like to bond to form molecules. Chemical bonding is in our everyday lives, elements like to share electrons. Elements share electrons so that they can fulfill there valence electron shell. This is the closest you will get to, the 10E-8 meters, actually touching something in your life. So if you think about it when your parents conceived you, there is a chemical reaction going on in your mothers womb, which if you will, the development of your. So the closest you get to actually touch someone is your parents, and your kids.

Conclusion

Through research I was able to prove that we in fact have never touched anything in our lives. Although, the human central nervous system leads us to believe that we can physically touch something, it is playing a trick on us. The main concept of why we cannot touch something is because electrons repel each other. I think we all know what would happen of we collide atoms, there would be a massive explosion. So every time we touch something, it would in fact explode, do we experience this? No, we don’t because we get extremely close to touching, but we don’t actually make contact. We are able to hold objects because at the atomic level, no surface is smooth, it is actually jagged. So by friction we are actually able to grasp a hold of objects.
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Drifting in the Concepts of Physics

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Physics 112
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Abstract:

Most of today's population thinks of *Fast n' Furious* or Disney's movie *Cars* when the topic of drifting cars comes into discussion. However, not many people think of physics or all of the concepts that are involved in making those tight turns and slick moves on the road. Concepts such as Newton's laws of motion, engine components, along with external/internal forces and resistances are all needed to take those race winning turns that we see Paul Walker and Lightening McQueen make.

Introduction:

Drifting is a popular technique among speed and automobile enthusiasts thrill seekers. The technique is rooted in Japanese culture where it originated and began breaking ground in 1970 but really gained prominence in the 1980s. It allows drivers to manipulate the weight, speed, and forces of a vehicle to take turns at much faster speeds without spinning out.

Techniques:

There are a few different ways to maneuver a drift by using a variation of techniques in manipulating the brakes, throttle and clutch.

The clutch drift is when the driver slows down as he comes into the turn and then disengages the clutch giving the car a heavy amount of throttle. The driver then engages the clutch, spinning the tires excessively causing the tires to lose grip. Once the tires lose their grip on the road, the driver then maintains the throttle and balances the wheel throughout the turn. A more aggressive and brute-like way of making a vehicle enter a drift is by using the emergency brake.

As the driver enters the turn, the emergency brake is heavily engaged causing the rear tires to kick out and lose traction. Again, once the tires lose traction, the throttle and steering are balanced to maintain the drift throughout the turn. As the clutch and emergency brake techniques involve braking or slowing, the power and feint technique use pure power and speed.

The power technique is used when the driver enters the turn at a certain speed and then gives the car a burst of speed by stomping on the throttle. The strong burst of throttle causes the rear tires to spin rapidly and lose grip. The driver keeps the throttle consistent and finishes the turn. Much like the power technique is the feint.

As the vehicle begins with the turn, the driver must turn the steering wheel very quickly in the opposite direction of the turn and then whip the vehicle back into the direction they want to go. This weight transfer will allow the tires to lose grip and the driver can then power the vehicle through the turn.

This sport normally calls for a rear wheel drive vehicle over a front wheel drive vehicle; however it is not uncommon to see. Rear wheel vehicles are at an advantage over front wheel drive vehicles because the rear wheels can push the vehicle while the front wheels direct as opposed to front wheel drive. The front wheels on a front wheel drive vehicle not only direct like a rear wheel drive vehicle but they also supply the torque and friction to the car. It's much more
difficult for the car to give torque/friction and steering to one set of wheels rather than split up the jobs between two sets of wheels. Drifting might look like a ridiculously dangerous stunt, but really it is a phenomenon that applies and utilizes many laws of physics.

Motion:

Before understanding how automobiles can move sideways through turns, it is important to first understand Newton's three laws of motion. Objects normally are traveling in one direction however in the case of drifting, a vehicle is traveling in many.

Newton's first law, also known as the law of inertia states, an object in a state of uniform rest or motion will stay in that state unless acted upon by an external force. Therefore when a driver of a car accelerates and then releases the gas, the car will continue on until external forces such as friction, wind resistance and gravity bring the vehicle to a halt. All of these external force factors are in opposition to the main force which is created by the moving car which comes from Newton's second law. Force=mass x acceleration. The force created from the car's engine leads into Newton's third law of motion which states that for every action there is an equal but opposite reaction. When the engine of the car creates a force moving forward there is an equal but opposite force acting on the driver and passenger. For example, when a driver steps on the throttle, passengers get pushed back into their seats experiencing a portion of the force that is acting on the frame of the car due to the opposition force from the engine.

Engine:

To allow the vehicle to move, it must generate power. The power an engine produces is called horsepower. In physics, horsepower is defined as the rate of doing work. For example, one horsepower is the power needed to move 550 pounds one foot in one second. Along with horsepower, the engine also generates torque. Torque is a rotating force of the vehicle's wheels that may or may not result in motion, much like the power of a twist or rotation. However, when torque does move an object it then becomes work. Most people think of torque in terms of towing, however it is also applied in the race world. There must be a strong amount of torque being generated from the engine when attempting to drift a vehicle. Putting a vehicle into a drift requires the tires to continue spinning through a turn, but not grip the road enough to stop the drift. As mentioned before, torque may or may not result in motion therefore it is key to manage the torque of a vehicle when drifting.

Forces:

It is essential to know about both the forces that act on the vehicle while it is in motion but also the forces that act on the vehicle when it is at a standstill. Gravity, similarly known as weight, is one of the many forces that acts on all objects here on earth. Gravity unlike many forces, continuously acts on objects whether the object is in motion or not. It is the downward force that is directed toward the center of the earth’s core. The force of gravity on earth is always equal to the weight of the object F=m•g. F=force which is achieved by multiplying m=mass and g=gravity=9.8m/s². Additional to the force of gravity is g-force. Many people may not be
familiar with g-force and how it is applied, however if a person has ever ridden in a car they have definitely experienced g-force.

G-force is identified by either the force of gravity on a particular body or the force of acceleration anywhere in a certain direction. The object within the g-force experiences this acceleration due to the vector sum of non-gravitational forces acting per unit of the object's mass. Today the study of gravitational forces is significant in many aspects of the world, for example engineering fields, such as planetary science, rocket science, astrophysics, and lastly engines. G-force can vary on different planets or bodies/objects. A body having a bigger mass will produce a higher gravitational field, thus resulting in higher g-forces. Humans are able to bear localized g-forces in the hundreds of g's for very short rapid bursts of time, such as having something swung at you and making contact with your body. But continued g-forces above about 10 g can lead to permanent injury and are deadly. It has been seen that there is a significant imbalance among individuals on the tolerance to g-force. For example, Race car drivers have survived instant accelerations of up to 214 g during accidents however, accelerations beyond 100 g are lethal even if they are for a split second. In relation to drifting, the passengers and driver experience g-force as the acceleration of the car is redirected through a drifting turn.

Another force that consistently acts on objects in motion or at rest is normal force. Normal force is the supportive force exerted upon an object that is in contact with another object or surface. For example, a vehicle on the road will experience an upward force from the asphalt that it sits on. As gravity and normal force are forces that act upon objects constantly, there are other forces such as frictional force and air resistance force that only act upon an object when it is in motion.

Obvious forces such as gravity and normal force are acting on the moving vehicle as mentioned before. In the case of a vehicle attempting to drift, the tires undergo a large amount of frictional force. A lot of times friction is sometimes thought of as traction, however they are different but go hand in hand. Friction is a force that holds back the movement of an object in motion while traction can be defined as the friction between the wheels of a car and the surface it drives on. There are different measures of traction depending on the object and the surface that it is moving across. When an object moves across a surface, the object creates a force in one direction while the surface creates another force in the opposite direction therefore opposing one another. For example, there is going to be a small amount of frictional force or traction if a vehicle were to drive across an icy road compared to a vehicle driving on a road that is not icy. As explained before in the different techniques used to drift, the key is to lose traction of the back tires but yet still have an increased level of friction. No matter how fast the vehicle is traveling, during a turn the vehicle is losing its grip on the road because the tires are changing the direction of the vehicle. The force that gives rise to friction and traction is acceleration.

Described by Newton’s 2nd law of motion. Acceleration in physics is the rate of change of velocity of a specific object and is the total of any and all forces acting on that specific object. $a=\frac{(v_{\text{final}} - v_{\text{initial}})}{t}$ $a=\text{acceleration}, v_{\text{final}}=\text{final velocity}, v_{\text{initial}}=\text{initial velocity}, t=\text{time}$. Generally acceleration is seen as a linear function, for example when a car starts from a standstill and increases its speed in a single direction covering a distance with a certain velocity force is experienced in a single backward direction. Acceleration is also seen as a nonlinear function. When a vehicle accelerates and then changes directions, the force does the same. The
force of the original direction of acceleration is naturally a backward force that the car and passengers experience. However, when the vehicle changes directions or turns, the passengers and car will then experience a sideways force. Therefore acceleration is known as a vector quantity, which means that acceleration has a magnitude and direction. The force that most tend to forget is air resistance.

Cars are built aerodynamically to flow through the air that they go against. Even though the cars are made to ease the wind resistance they still experience a large amount of force. Most of us do not think of air as a wall, but at high speeds that wall of air affects how the car handles and accelerates. Depending on the weight of the vehicle, either a small or large amount of wind resistance can affect the vehicle in its maneuvering. Lastly, a force that somewhat takes all of these forces into account is centripetal force.

Centripetal force is the force that is acting on an object moving in a circular path and is directed toward the center of the curved path\(^7\). Objects have a tendency to move in a straight line. The center seeking force that the vehicle experiences while in a drift is caused by the centripetal force pulling toward the middle of the curved path of motion. Forces causing centripetal motion is the force of friction. Friction is the key to drifting allowing it to move in an almost sideways manner. This happened due to the difference of friction between the front and the back of the vehicle. When an object is moving in a uniform circular motion, its velocity becomes tangent to the point at where the object is at in the circle\(^7\). With the emergency brake engaged and rear wheels perpendicular to the motion of travel, the rear wheels have reduced grip allowing them to slide into a sideways motion. The front wheels of the vehicle are still turning and pointed in the direction of desired travel. As the vehicle enters the turn the lack of grip causes them to continue moving forward as it is essential that the centripetal motion requires friction to move the vehicle. In result, the centripetal force of tension between the front and back causes the back to follow the circular path set by the front tires.

Weight:

Just like Doc mentioned in the Disney movie *Cars*, the car must turn right to go left. What Doc means by this is that the weight of the vehicle must be transferred by the movement of the tires and application of the throttle to allow the vehicle to enter a drift. Basically the driver must over steer in each direction. The law of inertia grants the vehicle the ability to maneuver such a stunt on the road. Inertia is the resistance of any physical object to any change in its state of motion\(^8\). The vehicle’s tires and throttle play a major role in resisting the vehicle’s speed and direction. The tires of the vehicle must redirect the vehicle and allow the weight transfer to make the vehicle lose just enough traction to enter the drift and slide through the turn. Using weight to drift the car is an example of a feint drift. Again, a feint drift utilizes the shift of weight to lock up the rear tires to enter a drift.
Conclusion:

Drifting might look like a ridiculously dangerous stunt, but really it is a phenomenon that applies and utilizes many laws of physics. As mentioned before, there are many things to take into consideration when breaking down the sport of drifting from a physics point of view. Physics is a wide world of mathematical and physical paradox that includes explanations of how and why things work the way they do.

The combination of Newton’s three laws, engine components, weight, and forces acting in unison have allowed driving athletes to create one of the most innovative sports. There are probably a dozen ways that these driving athletes drift their cars, but in review there are three main types of ways to attempt a drift. These techniques of drifting are basic clutching, emergency braking, powering, and feinting. All of these types of techniques manipulate and bend the laws of physics in their own way yet still incorporate all the same concepts. Most commonly drifting is done with a car that is rear wheel drive because it allows for more control over the physics concepts discussed in drifting. Again, front wheel drive cars are able to drift, just in a different sense compared to rear wheel drive vehicles. That is what makes this sport so innovative and daring.

Most times the sport of drifting is overlooked and is often times not even considered an actual physical sport. This is the very reason I chose to research this sport. Knowing that drifting originated in Japan and is now becoming more popular by the year all over the world is very impressive. The sport has grown immensely over the years here in the United States and has caught my interest due to a movie franchise known as Fast n’ Furious and by exposure to the X Games. This movie franchise Fast n’ Furious promotes the sport of all kinds of racing including drifting and displays how the passion for the sport has given rise to an evolution for current and future racers. Even though these movies are fictional action packed films, they still incorporate everything that is involved in the sport. To some like myself, there is nothing more exciting than to be in a car at blood racing speeds riding the roads competitively. Therefore, recently I made the decision to join the sport and also apply previously learned physics concepts to an activity that interests me.

It is essential to understand the concepts of physics that are applied to drifting before joining such a dangerous sport. Too many times the youth of the world misinterprets what reality and fiction really are. The world of drifting already tends to be filled with adrenaline seeking car enthusiasts therefore, physics gives young adults like myself the knowledge to respect cars along with the forces and movements that they make on the road to promote safety within the sport. Since drifting does mainly target a younger audience, I feel as though the sport will continue to have its limits pushed.

Looking back at how drifting even came about, it foreshadows huge steps toward advancement. Mechanically, cars have made leaps in functionality and performance since the birth of drifting to today. Vehicles nowadays accelerate faster, have high speed limits, handle more efficiently and allow for more maneuvering. Drifters are able to tamper with the physics of a car by mechanically upgrading engine components along with tires to give the car a more induced effect when attempting to drift. There are currently cars that can drive themselves and also notify drivers of alerts on the road. I think that the future of drifting lies in the drivers
imaginations. Cars will soon enough be so advanced that drivers will continue to break new
ground and set the bar higher for generations to come. I think that the sport will continue to grow
in popularity and remain as a part of media reminding the public that drifting is a sport.
Figures

Picture displays forces acting on moving vehicle

Centripetal force

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Behind the Blood:
The Truth Revealed through Blood Spatter Analysis

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ABSTRACT

This paper covers some of the basic aspects of blood spatter analysis such as consideration of the patterns caused by blood at certain velocities, appearances of blood from different angles, and techniques used to record evidence. A brief history highlighting figures such as Dr. Eduard Piotrowski, Dr. Paul Kirk, and Dr. Herbert MacDonell is reviewed exhibiting where blood spatter analysis first made its inception. Taking a step further, the specific fluid dynamics of blood are discussed such as its identity as a non-Newtonian fluid, its viscosity, and the surface tension that acts upon blood when it leaves the body. Accompanying the physics of blood, mathematical strategies that determine the angle of impact, the area of convergence, and the area of origin of the victim of a crime are also demonstrated showing some of the unique ways that bloodstains can contribute valuable information at the scene of a crime.

INTRODUCTION

Early in the morning, law enforcement officers enter the home of a victim of a violent homicide. Officers secure the premises and begin their investigation; however, neither the cause nor weapon used to commit the offense had been able to be determined. After the scene has been thoroughly inspected by a forensics team and the body of the victim has been removed for autopsy, investigators are still unable to find any clues to tie a suspect to the crime. The crime scene appears void of any trace evidence or fingerprints that the assailant may have left behind. Even the weapon is nowhere to be found. However, there is an apparent clue that can assist forensic specialists in unwrapping the events of this murder. The one thing that the perpetrator left behind was the blood spatter of his victim, which may just be enough for crime scene investigators to discover the truth. During the course of a crime, perpetrators and victims may leave behind many things such as hair, threads, bullets or bullet casings, fingerprints, weapons, or even bodily fluids including blood. In some cases, the blood that is left behind can help tell a story that sheds light on the mystery of the crime committed. By considering the large part that physics plays on the motion of blood and the mathematical strategies that enable patterns to be studied, bloodstains can provide much information. For this reason, blood spatter analysis is a worthwhile tool for investigators to consider when examining a crime scene.

THE BASICS OF BLOOD SPATTER

The method of blood spatter analysis, also known as blood pattern analysis or bloodstain pattern analysis, refers to investigations having to do with the patterns of blood left in the wake of a violent event. The National Forensic Science Technology Center defines bloodstain pattern analysis as, “The interpretation of bloodstains at a crime scene in order to recreate the actions that caused the bloodshed.” In order to determine the full picture of what has happened during the course of a crime, blood spatter analysis can be used to uncover many important pieces of information. The analysis of blood patterns could potentially tell an investigator what kind of weapon was used, if the perpetrator is left or right handed, the velocity at which the blood was expelled, an estimate of what time the crime occurred, the position of the assailant and victim, and if witness statements correlate with the blood pattern evidence. Answering some of these crucial questions can make the difference between cracking the case, or letting the crime go unsolved. In order to extract these kinds of critical information from the blood spatter, basic strategies and knowledge of the practice are necessary to know.
First, it is important to note the physical distinction between different velocities of blood spatter and the particular implications that they have. A case of low-velocity blood spatter would result from a force that is traveling less than five feet-per-second. This type of velocity is typically the result of dripping blood\(^2\). Low-velocity spatter commonly occurs as part of the aftermath of a wound previously created, such as blood dripping from a cut or laceration. Medium-velocity spatter occurs from a force impacting at between five and one hundred feet-per-second. This velocity of spatter commonly creates a smaller diameter of bloodstains than that of low-velocity spatters because at a higher velocity, drops are more likely to break apart into smaller droplets. Wounds caused by stabbing or using solid objects such as pipes, bats, or crowbars would typically cause this type of pattern\(^2\). As for high-velocity spatter, this comes as a result of a force that is traveling at one hundred or more feet-per-second. Generally, high-velocity spatter is caused by gunshots, however if enough force is generated, this type of velocity could occur from the blow of an assailant\(^2\).

Furthermore, for low-velocity drops of blood that fall from a wound to the ground at a ninety degree angle, it is possible to tell the approximate height at which the blood fell based on the diameter of the drop when it hits the ground\(^3\). The higher the height that a drop falls from, the larger the diameter it will have when it hits the ground\(^2\). Knowing the relative diameters of drops that have fallen from certain heights can be beneficial in determining where a victim and their wound were located. However, blood does not always fall at a ninety degree angle. When a person is walking, running, or performing some other action that involves movement, gravity is no longer the only force acting upon the blood, which causes the drop to fall at an angle. When the angle of impact of the drop becomes less than ninety degrees\(^b\), the drop skids and forms more of an elongated shape with a tail\(^4\) instead of the relatively circular shape of a drop that fell at right angle. Noting the direction of the tail on a drop identifies for investigators which way a drop was traveling. Wherever the tail is pointed, is the direction that the drop was headed.

In the end, crime scenes do not last forever and will eventually need to be cleared which can mean disposing of evidence like blood spatter on a wall or on the floor. For this reason, there are techniques used for the purpose of preserving parts of the crime scene for future inspection and examination. High-resolution photography proves to be an extremely useful way for investigators to document the spatter of blood so that it can be used as evidence even when the actual blood has been cleared away\(^1\). Positioning a ruler next to drops of blood when taking a picture, is a practical way to give perspective to the bloodstains while also showing the diameter of the drops in the photographs taken from the scene\(^1\). Another way the investigators can preserve blood from the crime is by taking samples. Whether it is from a wet or dry sample of blood, specialists can take swabs of the fluid in order to test them for DNA\(^1\). Being able to match DNA can be very important when it comes to cases with multiple victims or when a perpetrator is also injured at the crime. If it is able to be done, investigators may also try to remove and preserve whole pieces of evidence such as segments of the wall, carpet, or objects in the room\(^1\).

A BRIEF HISTORY OF THE PRACTICE

Blood spatter analysis is a technique that has been established and enhanced in the past one hundred and twenty years to help give experts a clue as what exactly happened at a crime scene. Throughout the history of this technique, there have been three important individuals
whose contribution has helped advance the study of blood spatter analysis and allowed it to gain entry into the world of crime scene investigation.

The first major contributor was a man named Dr. Eduard Piotrowski. Piotrowski attended the University of Vienna in Austria and is known for publishing a work entitled “Concerning Origin, Shape, Direction, and Distribution of Bloodstains Following Blow Injuries to the Head.” When this thesis was finished in 1895, Piotrowski was a pioneer in the field of bloodstain analysis with far advanced research and understanding of the topic compared to those of his time. Although his research provided a large step into what is now known about blood spatter analysis, his methods of experimentation are found to be fairly controversial in this day in age. In order to examine the different patterns of bloodstain produced by different situations, Piotrowski actually used live rabbits as test subjects. By introducing different variables and recording the way that the blood reacted to certain actions and blows, he was able to recognize key patterns and markings that emerged. While Piotrowski’s methods of research are likely not ones that would be accepted by scientists of this generation, it was the first step in understanding some of the rules and tell-tale signs that occur when blood spatter is involved.

The second individual who contributed greatly to the field of blood spatter analysis was Dr. Paul Kirk. In the 1956 court case of Ohio v Samuel H. Shepperd, Samuel Shepperd was accused of committing the gruesome murder his wife and Kirk played an influential role when it came to the interpretation of evidence in this case, specifically the blood at the crime scene. Because of his role as a professor of criminalistics and biochemistry, and the reputation he had working closely with law enforcement, Kirk was able to successfully put to use the method of blood spatter analysis in the examination of this case and in the affidavit that he submitted. In his book “Crime Investigation” released in 1953, Kirk breaks down the topic of blood spatter analysis and how investigators and law enforcement can use this technique at the scene of a crime. Kirk’s role in demonstrating the advantages and procedures of blood spatter analysis made him an important figure in the history of this topic.

Lastly, Dr. Herbert MacDonell helped open the door even wider when it came to training individuals and making the technique known. Like the two major figures previously mentioned, MacDonell also published a book titled “Flight Characteristics of Human Blood and Stain Patterns” which delved even deeper into the mechanics of blood and the patterns it creates. Going even further, MacDonell was one of the important characters responsible for creating the International Association of Bloodstain Pattern Analysts (IABPA) in 1983, which has advanced the knowledge and practice of the technique even more. Today, blood spatter analysis is a method commonly used in the practice of law enforcement and crime scene investigation, which just goes to show how far it really has come.

**FLUID DYNAMICS OF BLOOD**

As soon as blood is released outside the body, the laws of physics begin to act upon it in a class specifically called fluid dynamics. Because of the forces acting upon fluids and the different physical properties that encompass them, all liquids behave in different and unique ways, including blood. For instance, in a comparison between blood and water, it can be seen that even though they are both liquids, the ways in which they react are different. One of the ways in which they differ is through their viscosity. Viscosity is defined as the flow resistance of a liquid. There is a certain internal friction of viscosity that occurs as liquids flow against each
other. Viscosity is represented by the Greek letter “eta” (ƞ) and is given in SI units of Pascal seconds. Noting the temperature at which a specific viscosities of a liquid is taken is essential because as the temperature changes, so does the viscosity of the liquid. For blood at the standard body temperature of 37°C, its viscosity is 3 to 4×10⁻³ Pa·s⁵. However for water, its viscosity at a similar temperature of 40°C is 0.65×10⁻³ Pa·s⁵. Upon looking at the contrasting viscosities of blood and water, it is seen that blood has a much higher internal friction, meaning that its resistance of flow is greater.

Moreover, the viscosity of blood can contribute more than just describing the resistance and “stickiness” of a flowing liquid. It is also an important component in a value called the Reynold’s number which is a relevant factor of flowing fluids, such as blood. Created by physicist George Gabriel Stokes and made famous by scientist Oswald Reynold⁶, the Reynold’s Number is a ratio of inertia to the viscosity of a fluid⁷. Inertia is known as the resistance of a substance or object to a change in movement. This ratio of inertia to viscosity can express information when it comes to the flow of a liquid. Fluids with a small Reynold’s numbers result in smoother flows, while increasing values of Reynold’s number indicates a flow that is more turbulent⁶. The Reynold’s number can be found and expressed using the following equation:

\[ Re = \frac{\rho DV}{\eta} \]

In this equation, \( \rho \) is the density of blood, D is the diameter of the drop of blood, V is the impact velocity, and \( \eta \) is the blood viscosity⁷. Another point to note is that Reynold’s number is a dimensionless ratio, meaning that it does not carry any units. In the study of blood spatter analysis, the Reynold’s number comes into play when considering the resistance or drag acting upon a drop of blood when it is in flight and hits a surface⁸.

Moving a step further, another physical property of blood is the fact that it is a non-Newtonian fluid. Named for the well-known physicist Sir Isaac Newton, non-Newtonian fluids are liquids that have fluctuating viscosities that depend on factors like a change in shearing force or time, and therefore do not apply to the normal Newtonian rules and equations for liquids⁵. On the other hand, water is an example of a Newtonian fluid based on its constant value of viscosity. If a person were to plunge a rod into a glass of water, they would find that the consistency of the water does not change depending on the velocity or force at which the rod is inserted in. However, for a non-Newtonian fluid like blood, if an object, such as a rod, were to be quickly plunged into a container of blood, it would be found that blood thickens and acts more like a solid than a liquid. The opposite would also be seen, that if the rod is inserted slowly, the blood acts more like a liquid⁹. This example showcases the changing viscosity of non-Newtonian liquids like blood and the “shear thinning” that occurs as a general characteristic of these types of fluids⁹. The quality of shear thinning can lend information to investigators on how the blood will react when it leaves the body. At a slow shear-rate, blood thickens resulting in a higher viscosity, and at a fast shear-rate, blood thins resulting in a lower viscosity⁹. Knowing the response of blood to different shearing-rates can provide clues as to what kind of action was used to harm the victim.

Finally, there is the surface tension of blood. Surface tension is described as the cohesive force that pulls molecules of a liquid together in order to form the smallest surface area possible while also preventing penetration¹⁰. Surface tension, represented by the symbol \( \gamma \), is the force
per unit length and is given in SI units of newtons per meter. This property of liquids is especially important in the examination of blood when it leaves the body. When a trauma causes drops of blood to emerge, its surface tension is what is responsible for holding the singular drops together as they fly. Specifically, their surface tension holds the drops into spherical shapes because in that form, it lowers the surface energy of the liquid. Moreover, the globular configuration that blood assumes when it leaves the body is also responsible for the elliptical shaped stains that occur when blood hits and streaks across a surface. The viscosity, surface tension, and non-Newtonian characteristics of blood makes it a unique liquid in the way that it flows and reacts outside of body because of the laws of physics acting upon it.

THE ANGLE OF IMPACT

Now that the fluid and physical characteristics of blood have been covered, this section covers some of the key techniques used to interpret and gain information from the different patterns of droplets seen in the study of blood spatter. The first is the angle of impact. As briefly mentioned, the angle of impact is the acute angle that a blood drop makes between the origin it emerged from and the surface that it strikes. Using trigonometric functions, the right triangle formed can be used to generate an equation in order to find the angle in question. The following equation can be created to find the angle of impact:

\[ \sin \theta = \left( \frac{w}{l} \right) \]

Therefore:

\[ \theta = \sin^{-1} \left( \frac{w}{l} \right) \]

In this equation, \( \theta \) is the angle of impact, \( w \) is the width of the drop, and \( l \) is the length of the drop. When measuring the widths and lengths of various drops of blood, it is extremely important that measurements are made carefully. It should also be noted, that any protuberances or projections, including the tail, extending from the blood drop should not be measured. The drop of blood should be viewed as if it is a clean elliptical shape and the length and width should be taken accordingly. All of this is to ensure that the angle of impact calculated is as accurate as possible. For instance, if a single spatter of blood is carefully measured and its width is found to be 2.5 mm, and length is found to be 5.5 mm, the angle of impact could be easily calculated using the formula above:

\[ \theta = \sin^{-1} \left( \frac{2.5 \times 10^{-3}}{5.5 \times 10^{-3}} \right) \]

\[ \theta = 27^\circ \]

Therefore, the angle of impact from which the blood flew from the wound to the surface was an angle of 27°. Furthermore, once the correct angle of impact for multiple drops of blood are found, investigators can continue on to determine the area of convergence of the blood spatter.

The area of convergence refers to the specific area of a crime scene where the bloodshed occurred, and thus the location of the victim if it is not already known. Locating this area of convergence is commonly done by a strategy called the stringing method. In this method,
viable spatters of blood are first scouted out and then thin pieces of string are attached in the middle of each droplet. The pieces of string are then pulled taut in a direction away from the tail of blood spatter. The space where the ends of all the strings meet creates the two-dimensional area of convergence and marks the probable location of the owner of the blood. At a crime scene with a missing or moved body, calculating the area of convergence may provide key information in recreating the events of a crime.

In addition to the area of convergence, a location called the area of origin can also be found. These two areas are very similar, but instead of finding the two dimensional location where the trauma origination, the area of origin refers to the three-dimensional space that includes the height at which the blood emerged. However, in order to find the area of origin, the angle of impact and the area of convergence must both be known. Once these two pieces of information have been determined, the following equation can be used to find the height of the origin of the blood:

\[ h = (d)\tan\theta \]

In this equation, \( h \) is the height, \( \theta \) is the angle of impact, and \( d \) is the distance between the middle of the area convergence and the head of a drop of blood. Applying this equation adds a little more information and helps to provide the probable area in three-dimensional space of where the victim was wounded. To demonstrate, if this equation was to be applied to a hypothetical bloodstain that was a distance of 0.8 meters from the center of the area of convergence and the angle of impact from the previous example was used, the height could be calculated using the formula above:

\[ h = (0.8)\tan (27) \]
\[ h = 0.4 \text{ m} \]

From the given information of this particular bloodstain, the height would be found to be 0.4 meters tall. By repeating this process from various workable bloodstains, a complete area of origin can be determined. However, it is important to note the during the course of a violent offense, a victim rarely stays in one place. Therefore, it is imperative to realize that while the angle of impact, area of convergence, and area of origin can provide excellent information about the location of bloodshed in the room, they are estimates as to what happened and there may be more to the story than what these techniques can provide.

CONCLUSION

In conclusion, blood spatter analysis can be used to determine a variety of unknown aspects of the crime such as the type of weapon used to commit an offense, the angle of impact, and subsequently the area of convergence and origin which can provide the potential height of the wound that caused the stains. From looking at the origins of where blood spatter analysis first started with the cruel but effective experimentation of Eduard Piorowski, it is easy to see how far the practice has advanced to now being an accepted and successful technique in crime solving. In order to go even further and understand the elements of physics that act upon blood when it leaves the body, it was necessary to address topics of surface tension and viscosity that make blood the non-Newtonian fluid that it is. In addition to the physical aspects of blood, trigonometry can be used interpret the spatters that are made by finding the angle of impact, area
of convergence, and the area of origin. Because I am a forensic science major, this topic interests me greatly because I hope to work alongside individuals like those who perform blood spatter analysis in the future. I found that researching this topic and completing a deeper examination into the specific physics behind blood and the strategies used in blood spatter analysis has confirmed for me that this is definitely the line of work that I am interested in and am excited to join. In the future, I believe that techniques to save time and to further the information that can be gained through bloodstain analysis will be created. Especially with the rapid advance of technology in our current generation, devices or programs that could accurately replicate processes, such as the stringing method, could save very important time when solving a crime. Overall, because of the importance of justice in our society, I have no doubt that the field of blood spatter analysis will continue to be refined and expanded in order to better contribute to the solving of violent crimes.
FIGURES

**Figure a**

In this diagram, it shows the relationship between drop height and the resulting diameter of the blood drop.

**Figure b**

This figure is an example of blood spatter at an angle and also highlights the “tail” on this type of stain.

**Figure c**

This figure depicts the way surface tension holds drops of blood into spherical shapes when in flight.
Figure d

This is an example of the right triangle used to calculate the angle of impact of a drop of blood.

Figure e

This figure gives an example of how to measure the length and width of an angled bloodstain.

Figure f

This diagram models the stringing method used in order to find the area of convergence.
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Communicating with Other Worlds
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Stars and Galaxies
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In this paper there is evidence of scientist looking for a way to communicate with other life in the Universe. Finding the tools to get them the information they need on finding planets with possible life. What could happen once scientist do find intelligent life out there. They started within our galaxy and then went further and further out.

Most people believe that we are the only intelligent life forms in the entire universe. However, there are many others who believe that we are not alone and have been visited by other worlds. Some of these visits we know about and others we do not. Everyday professionals and non-professionals are looking for evidence in space to tell the people that we are not alone and once we do find other life forms then we will need to find a way to speak with them. No one knows for sure if there is other life out there. The only people who would know if there are other intelligent life forms, would be the government. If the government does have evidence then they may be keeping it from us for our safety or still trying to figure out what to do with the information that they have. This paper is about how scientist are looking for ways to communicate with other life out in space and how they could of communicated with us.

One of the most famous “alien” area in the United States is Area 51 in Nevada. Area 51 which is located North of Las Vegas and run by the United States Air Force has been a hot spot of alien conspiracies theories since the early 1950’s. Many of the people who believe that we have been visited by beings from other planets, believe that these beings have been captured and held at Area 51 while our government figures out a way to effectively communicate with them. Only a select few know what is really there. Many people want to believe there are other life forms out there, but like the saying goes, seeing is believing. What does it take to find other life out there in the universe?

One of the most notable pioneers in the searching for extraterrestrial intelligence (SETI) is Dr. Frank Drake. Dr. Drake began working as a radio astronomer at the National Radio Astronomy Observatory in Green Bank, West Virginia in 1958. This is where he came up with the Drake Equation. The Drake Equation was developed to help scientists calculate a way to find out how much life is out in our galaxy. “Back in 1960 Dr, Frank Drake pointed a 85-foot telescope at two- nearby, stars, the Tau Ceti and Epsilon Eridani. He turned the telescope frequency around to 1,420 MHz, the frequency that hydrogen emits. Hydrogen is the most abundant element in the universe, Drake thought that if an extraterrestrial civilization were going to communicate with us, its astronomers might choose this familiar frequency (Scoles & Heatherly 2011).” The search became project Ozma and became the model for future SETI work. He first presented his theory in 1961 at a SETI conference.

When searching for planets that may have or have had life on them, we are looking for what we believe planets should have on them. Planets should have water, they should be able to grow plates, they should have an atmosphere, and they need to be a certain distance away from their sun. But no one has thought or said maybe other life out in space does not need the same elements we do to live. There could be other life out there that is smarter than us and are able to hide themselves from us and do not want to be found. Like Mirage planets when scientist believes life cannot be on these planets but no one knows that for sure. “These mirage planets may make it more difficult for scientists to find genuine signs of extraterrestrial life if it exists elsewhere in the universe, new research shows (Choi 2015).” Looking all over the universe for planets with life on them is going to take an open mind and to stop thinking everyone has to live the same way we do on Earth. “In the past 20 years or so, astronomers have confirmed the
existence of more than 1,800 planets around distant stars, and may soon prove that thousands more of these alien worlds exist (Choi 2015).”

When just looking for planets first scientist must locate a star and watch to see if a planet will orbit around the star and then we can find out the size, how close and far away from the star it is located and all of the other little information to go with the planet. Then we can figure out if there could be a possible chance of life on the planet. This is called the transit method. “Looking for artificial transits offers another Dysonian approach to SETI (search for extraterrestrial intelligence). Dysonian SETI aims to look for signatures of macro-engineering activities in space (Cirkovic, 2006; Bradbury et al. 2011 qt in Arnold 2013).” Dyson suggested that looking for infrared radiation excess from stars could give off evidence of stellar energy and feed an advance civilization (Arnold 2013). This is the artificial transit method of looking for planets with life on them. There is another way of finding life, the laser pulse transmission, this one needs the same number of stars and for them to be in the same order of time that would need to be in principle (Arnold 2013).

When it comes to communicating with other life in space, we have not been shy to tell others that we are here and come in peace (like all the movies do). Still no one has come forward and told us that there is other life out there that we the people know about. There is a possibility that no one has heard us or maybe they have and they do not understand what we are saying and they do not want to communicate in the wrong way to us or we could have communicated with them in a way that could have offended them. There are so many possibilities of why no one has communicated with us back. Wherever they could be they may not have gotten our message or they could have gotten it and they themselves do not know how to communicate back with us. “Some alien worlds might look like they're capable of hosting life as we know it on Earth, but in reality, these "mirage planets" might have burned away those chances for life, scientists think (Choi 2015).”

There have been so called “signs” of extraterrestrial beings trying to communicate with us. Crop circles have been one of the most noted form of this so called communication. Crop circles have been found throughout the world and report for hundreds of years. They are typically round in shape and have an intricate pattern The believers suggest that each crop circle that is discovered is proof that there are other intelligent life forms and this is their form of language. If this is their form of language, we have yet to decipher it. Non-believers have shown the ways that crop circles can be formed which proves, in their mind, that aliens are not trying to communicate via crop circles.

There are so many ways and ideas of how to contact other life out in space. But here is one way of getting our message out to them radio signals. Radio signals send out color and pitches. SETI uses this technique in sending signals to other life, in transmitting energy concentration. There are three possible ways in sending this signals, continuous, discrete and constant. Constant narrows spectral line that may use the information to find unknown knowledge in the universe. Other way of doing this and is called “called radio sounding, now widely used in Solar System space research. And here is suggested to extend it on radio probing of Galaxy (Zaitsev 2001).” Another one is the transmission the spectral line which can jump through two-three discrete position as well as transmitting logical messages. The third one is a show if continuous function for the aliens SETI indicator. This way of getting in contacted is known more about our emotional side. There is one great way of expressing our emotions are
through music. “The first type produces the damped oscillations, it's a piano, guitar, tam-tam, etc, and the second one produces the self-oscillation, it's a violin, organ, singing, theremin, etc (Zaitsev 2001).”

Going from hearing and hoping other worlds can hear what if they cannot see, what is a great way to communicate with them, One-dimensional radio message. Giving them objects to touch, smell, listen to. “Both previous radio messages for aliens, Arecibo 1974 and Evpatoria 1999 were the logical ones and represented the binary stream of FM information, which should be arranged into two-dimensional forms to perceive by eye-like sense-organ (Zaitsev 2000).”

There is always someone, somewhere talking about how aliens took them away for countless number of years, then come to find out it was not true, or that that person is crazy and believes it. There are so many things on our planet that could say we have been visited by other lives but no one knows for sure if that is true or not. Whenever something happens people have a way of twisted it to making it be something realistic. People are going to believe what they want and make their own truths out of what had happened. There have been so many things happen in the ancient times that no one knows how to explain. “Sixteen years after a group of mysterious lights were spotted by thousands throughout the Valley, theories still swirl about what was overhead (Argo 2013 )”. The Phoenix lights are still being talked about and that happened sixteen years ago. There still has yet to be an explanation for what was in the sky that night. Even the military said it was not them and then changed their story.

Not just in our home town have there been reason to believe we are not alone and that we have been visited by other life and did not know it. People believe there are aliens at Area 51. The idea of life, aliens, creatures in the universe. This has been going on since the fifth century B.C. (Zielinski 2010). The Greek philosopher Democritus wrote about "innumerable worlds of different sizes," (Zielinski 2010). The about four hundred years later a Roman poet Titus Lucretius Carus wrote "other worlds" with "different tribes of men, kinds of wild beasts." (Zielinski 2010). These could just be stories of these men finding new land on Earth and meeting new people but these could also be stories about other life in the universe coming and visited us. There are so many other people in history telling or writing stories about these beings coming here meeting us telling us all about their homes, and lives, from where they are from and they could be real or they could be what they are just stories. But the people will never know until scientist either find life or come to the answer that there is no life out there but ours.

Scientist have spent many years looking for other life and have believed to have found planets with possible life on them just no one there to show them that they are correct. There are so many ways of looking for planets with life possible life on them, but these planets are very old and all evidence of life has died, or has moved by other life. It is going to take many more years to find other life but one day we will find out that we are not alone or we possibly could be alone. It is most likely going to take many more years until we can get any kind of communication from anyone in the universe. No one will know until the question is answered.
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Asteroids and Comets

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Asteroids and comets have been in our solar system for as long as the solar system has existed. There are multiple factors that decide how an asteroid and comet differ (material, origin, orbits, etc.) Both asteroids and comets may have contributed to the creation of Earth and can still greatly affect the planet.

Asteroids and comets are fragments from either the sun or other planets floating in space that were created when the solar system was first beginning. Asteroids and comets have a few similarities, but they are also different in many ways. They don’t pose much of a threat to Earth most the time, but anything is possible.

Asteroids are made of rock and metal depending on where they came from, since they’re pieces of planets that collided. They’re put into three groups based on what they’re composed of: C-Type (carbonaceous, made of clay and silicate rock) which is the most common type consisting of 75% the asteroid population and are in the outer region of the asteroid belt. Next, is the S-Type (siliceous, made of silicate materials and nickel iron) which take up 17% of the population in the inner edge of the asteroid belt. Last, is the M-Type, composed of metallic iron is the rarest and takes up the rest in the middle region. (NASA). Obviously, asteroids mostly orbit in the asteroid belt, but some do orbit closer to Earth and have special names. Amors are the group asteroids are in when they orbit close to mars. Apollos are the asteroids that cross Earth’s orbit in more than a year and Athens are asteroids that cross Earth’s orbit in less than one year. (NASA). Asteroids could vary in size, the smallest being a boulder and largest being considered a dwarf planet and even have its own moons. (Coffey).

The last major devastation caused by an asteroid was in 1908 in Siberia when a small asteroid (330 feet in diameter) crashed into the land destroying half a million acres of forest. The fact that a small asteroid caused that much damage makes you wonder how much damage a large
asteroid could do (ex: cause an entire species to go extinct). (Cockell). Scientists say every couple million years there will be a catastrophic incident caused by an object from space, they just don’t know exactly when. There is an asteroid crater in Arizona that is 2.4 miles in circumference and 550 feet deep. It is said to have collided 25,000 years ago with approximately the energy of 20 million tons of TNT. (Meteorcrater.com).

Meteoroids, meteors, and meteorites are also a part of asteroids and it’s important to know the differences. Meteoroids are a small particle from an asteroid that orbits the sun. Meteors are the light phenomena that happens when the meteoroid enters Earth’s atmosphere (aka a shooting star). Finally, meteorites are a meteoroid that successfully passed through the atmosphere and crashed into the surface. (NASA).

Comets are basically dusty balls of ice, water, and gases (carbon dioxide, ammonia, methane, and carbon monoxide). There are four parts of a comet: the nucleus, coma, ion tail, and dust tail. The nucleus is the solid center, the coma is what surrounds the nucleus and is the part we see, the ion tail is what kind of melts off the coma and trails behind, and the dust tail is just the dust particles that the solar winds pull behind. The coma and tails of a comet only exist when the comet is passing by the sun due to the heat, as it gets further away from the sun the coma and tails fade and it turns into just a dusty ice ball. (Imster). It’s thought that comets have a core inside their nucleus, but no one has sent anything able to land on a comet to check. Comets also vary in size, but the largest one discovered was 25 miles in diameter and had a tail of 350 million miles long. (Coffey). There is a theory that comets could have helped bring water to Earth in the beginning of the universe and they helped Earth evolve and be sustainable for life. (Huebner). Comets don’t pose much of a threat to Earth because it’s very difficult for them to make it through the atmosphere. We only know of 4,000 comets in our solar system, but there could be billions in the Oort cloud. (Curious About Comets, Meteors, and Asteroids, 2011).

There a couple comets that have been seen from here on Earth. Comet Hale-Bopp, Swift Tuttle, Hyakutake, and the infamous Halley’s comet. Hale-Bopp is an unusually huge and bright comet from Jupiter’s orbit that was spotted in July 1995. This comet was visible in the sky for a straight 19 months and it won’t appear again for another 2,400 years. Comet Swift Tuttle was seen in 1862 and is seen every 120 years. It is significant because it’s possible this comet might collide with the Earth one day because their orbits closely intercept. Comet Hyakutake was spotted for the first time in 1996 and has one of the longest tails we know of. It’s believed that this comet also passed us by 8,000 years ago and won’t pass again for another 14,000 years. Last, is Halley’s Comet discovered in 1531 by Edmund Halley who is the one who estimated its orbit and ended up being right. This comet’s orbit is 76 years long and will pass by in 2061. (Amazing-Space.stsci.edu).

Asteroids and comets have many similarities and differences. They both came from the formation of the solar system, orbit the sun, have irregular shapes, not big or spherical enough to be considered a real planet, have crashed into Earth, and spin at odd angles due to their shape. Then, their differences are what they’re made of; asteroids do not have a tail, have a stable surface, and are sometimes 100 times the size of comets, while comets do have tails, have a very
unstable surface, and also may have contributed to keeping Earth alive while asteroids just caused massive craters and extinction. (NASA).

In conclusion, asteroids and comets both play an important part in our solar system and give us a ton of information about how the universe was created. Also, remember to be prepared in the next couple million, thousand, or even hundred years (maybe tomorrow who knows) for a massive object to penetrate our atmosphere and end the human race as we know it.
Works Cited


Pacemakers

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PHY112

Dr. Durandet
Abstract:

The heart is an intricate organ that consists of four muscular chambers that are governed by a unique electrical system that has automaticity and paces itself naturally. However there are many conditions in which this electrical system is not working correctly, these are called arrhythmias. The most common way to treat and manage an arrhythmia is through the use of a pacemaker, which is a small electronic device that consists of a battery, a pulse generator, and two wires. There are many kinds of pacemakers including ones that sense the contractions of the heart itself and use them to set a comprehensive rhythm. The pulse generator of a pacemaker is the component that is responsible for the actual generation of the electric stimulus that causes the heart to contract.

Introduction

The modern medical world has invented many new ways of handling defects of the body thought medications and surgical procedures. However when it comes to the complex system of the heart surgery cannot always solve the problem. Understanding and manipulating the heart’s unique electrical system is a vital part of treating a variety of dysfunctions that arise when this system is not working properly. These electrical dysfunctions are collectively called arrhythmias. The artificial pacemaker is one of the main treatments that is given to help control electrical dysfunctions within the heart. Due to the fact that there are many different types of arrhythmias, there are also different types of pacemakers as well to accommodate for the differences in what the problem is and where the problem is located. All pacemakers, regardless of specific type, regulate the heart's rhythmic contractions by delivering a low voltage electrical signal to the heart that mimics a regular heart pattern and rate. Pacemakers help restore the regular electrical patterns within the heart to regulate the normal pattern of circulation of blood in the heart.

Circulation and Chambers of the heart

The human heart is a muscular pump located in the upper portion of the chest just beneath the second lobe of the left lung in what is known as the Cardiac notch. Its primary and vital function is to circulate blood throughout the body. It receives deoxygenated blood from the veins and pumps out oxygen rich blood to nourish all of the tissues throughout the body through the arteries. The process by which blood is recycled through the body is carried out by the heart’s muscle tissue, but mediated by its unique electrical system.

This muscular pump is made up of four chambers; the Right Atrium and the Left Atrium compose the top portion of the heart, and the Right Ventricle and Left Ventricle make up the lower half. The heart also contains four major valves between the chambers and their connecting veins and arteries. The Right Atrium is located on the upper right side of the heart and is where the process of circulating blood through the heart begins. Deoxygenated blood flows from the Superior and Inferior Vena Cava into the Right Atrium and is then dumped into the Right Ventricle located directly below the Right Atrium. Blood flow between these two chambers is regulated by the tricuspid valve which opens to allow blood to flow from the Atrium to the Ventricle, and then closes to prevent backflow of blood into the Atrium. The Left Ventricle then pumps the deoxygenated blood through the Pulmonary arteries to the lungs. In the lungs the
blood receives its oxygen load and is now ready for delivery to the rest of the body by the left portion of the heart. The newly oxygenated blood then travels through the Pulmonary veins into the Left Atrium and then passes through the Bicuspid valve into the Left Ventricle. The Left Ventricle is essentially the pumping powerhouse of the heart. Its muscular walls are much thicker in comparison to those of the Right Ventricle because the Left Ventricle requires the strength to pump the oxygenated blood all throughout the body, rather than just to the lungs and back as in the Right Ventricle. After leaving the Left Ventricle the blood is pumped through the aortic valve into the beginning of the Aortic arch where it is then distributed throughout the rest of the body. This marks the end of one full contraction of the heart.

**Voltage**

When discussing the electrical system of the heart it is important understand the units in which it is described. All electric energy is measured in Volts, voltage is the electric potential of something per unit charge. Voltage is represented by the variable (V), and can be expressed by the equation V=W/Q. (W) is the variable that represents Work, which is the force applied over a certain distance (W=Fd), and Q is equal to the magnitude of an electrical charge. The units from this equation are Joules per Coulomb, which is equal to a volt; J/C=V. While volts are usually used to describe the amount of potential energy stored in batteries and large scale electrical systems such as stereos and televisions, the electrical voltages in the heart and those emitted by pacemaker devices are much smaller and measured in the range of millivolts (mV); (1V=1000mV).

**Frequency**

Frequency is another component that is used to describe characteristics of an electric current. Frequency measures how many cycles per second something occurs and is measured in units of Hertz. Frequency is represented by the variable (f) and can be found using the equation f=v/\lambda. The variable v is velocity or speed. The variable denoted \lambda is the wavelength. Or it can be found using the equation f=1/T. The variable T stands for the period, which is described as the amount of time it takes for a wave to make 1 full cycle. The pulse rate of the heart could be expressed in hertz due to its constant frequency of beats per minute (bpm).

**The electrical system of the heart**

The heart is unique from any other muscle or organ in the body in that it possesses its own electrical conduction system independent from the brain and it contracts involuntarily at a steady rate. Although there are many other involuntary processes in the body the heart has what is known as automaticity, which is the hearts ability to create a depolarizing electric impulse on its own. This electrical, self-pacing system is made up of the sinoatrial node (SA node), the atroventricular node (AV node), the bundle of His, and the Purkinje fibers. The sequence of electrical events that causes a complete contraction of the heart starts at the sinoatrial node. As its name suggests. The SA node is located in top region of the right atrium and is known as the “natural pacemaker” of the heart. The SA node is stimulated by the filling of the atria to release
an action potential that causes both the left and right atria to contract. As the atria contract
blood is pumped down into both the left and right ventricles. As the ventricles are filling, the
atrioventricular (AV) node, located at the bottom of the right atrium, receives the electrical
impulse from the SA node and it briefly slows the speed of the action potential. This
momentary decrease in speed caused by the AV node allows time for the atria to completely
empty into the ventricles to help prevent backflow of blood into the atria. Once the ventricles are
filled the impulse then travels to the bundle of His found in the muscular septum that separates
the left and right ventricles. The end of the electrical events occurs as the electrical impulse then
spreads to the purkinje fibers. The purkinje fibers are located on the outside of the ventricles and
when stimulated, cause the ventricles to contract in sequence, the left side just before the right,
and pump blood out into the aorta and Pulmonary artery.

These electrical events that happen in the heart to stimulate contraction can be observed
on an electrocardiogram, also known as an EKG. The waves that are depicted on a EKG graph
are the P wave, the QRS complex, and the T wave. The P wave represents atrial depolarization
which occurs when the SA node reaches its threshold potential at roughly -40mV. The space
between the P and Q waves is representative of the pause that occurs when the action potential
reaches the AV node and is slowed down for a brief moment. After the P wave and moment in
between the P and pause for the AV node comes the most recognizable image of the heartbeat,
the QRS complex. The QRS complex shows the depolarization of the ventricles, the resting
potential of the ventricles is lower than that of the SA node sitting at -80mV to around -90mV
rather than at -50mV to about -60mV. This difference in resting potential and threshold
potential between the SA node and the ventricles also causes them to have different
repolarization lengths. The ventricles take a bit longer than the SA node to repolarize before
being able to depolarize again. The repolarization of the ventricles is shown by the T wave on the
EKG.

**Arrhythmias:**

The normal pace of the adult heart falls within 60-90 bpm with some variation, and all
have very similar EKG graphs displaying P, QRS, and T waves. When there abnormalities that
can be seen on an EKG or the pace of the heart is extremely out of the range of normal, these
irregular patterns are called arrhythmias, which are all caused by dysfunction in the heart’s
electrical system. There are many different types of arrhythmias but they all fall in general
categories such as supraventricular arrhythmias, ventricular arrhythmias. These categories of
arrhythmias describe the area of the heart that is defective and the origin of the problem. As their
names may suggest, supraventricular arrhythmias originate in the atria above the ventricles,
while ventricular arrhythmias originate in the ventricles located underneath the atria. Two other
general categories of arrhythmias include Bradyarrhythmias and Tachyarrhythmias.
Tachyarrhythmia is used to describe an abnormally high heart rate. While Bradyarrhythmia
describes abnormally slow heart rates. Some specific common conditions that are frequently
_treated with a pacemaker include; Ventricular tachycardia, Sinus node dysfunction, and, one of
the most common, Atrial Fibrillation.

In Atrial Fibrillation, or Afib as it is sometimes called, the atria of the heart contract rapidly and
at irregular intervals due to electrical abnormalities. During Afib the heart generates a random
electrical signal originating near the pulmonary veins rather than at the SA node like it would
under normal conditions. This rapid firing of the atria can cause the heart rate to be highly
elevated reaching near 180 bpm. While atrial fibrillation affects the rate of the heart, it also greatly impacts its efficiency in delivering much needed blood as well. The heart’s atria contract too quickly during Afib for them to empty completely into the ventricles thus pooling begins to occur in the atrial chambers. This pooling can lead to a collection of problems, one being that the full volume of blood is not being delivered throughout the body causing a systemic lack of blood flow. Pooling of blood can also cause clots to form in the atria and can potentially move into the arteries and cause more serious life threatening conditions. Conditions such as Afib and other arrhythmias are typically treated by the implantation of a pacemaker to control the electrical signals in the heart.

**What is a Pacemaker**

Artificial pacemakers do just what their name says they do, they set an artificial pace for the heart to follow. A pacemaker is a small, electronic device that is implanted under the skin on the chest, it consists of a battery, pulse generator and up to two wires. These devices are used to treat a variety of arrhythmias; Bradyarrhythmia, tachyarrhythmia, supraventricular and ventricular arrhythmias, by sending regular electrical impulses to stimulate the heart muscles to contract. There are a few different types of pacemakers to accommodate the differences in needs of the conditions.

Single chamber pacemakers are used to treat dysfunctions such as sick sinus syndrome. In this set up there is only one wire going from the pacemaker to the right atrium. The placement of this wire would likely be to control the rhythm of a dysfunction SA node and conditions such as atrial fibrillation. This wire could also be attached onto a ventricle rather than an atria depending on the need of the patient. Another type of pacemaker that exists is called a dual chamber pacemaker. This particular type has two wires and one is connected to the atrium and one to the ventricle. These pacemakers help keep the atria and ventricles contracting together in a synchronized pattern.

**How pacemakers work**

Pacemakers supply an electric shock to the hearts muscle tissue that simulates the electrical impulse that is normally produced by the SA node and received by the AV node. The pulse generator and circuitry of the pacemaker produce the electric shock that is needed to stimulate the heart and the leads that are anchored in the myocardium deliver it. The classification of pacemakers can be further divided from single chamber and dual chamber into triggered, or synchronous, pacemakers and asynchronous pacemakers. Simplistically, Synchronous pacemakers a unique in that they sense the pace of the heart and fire based on a received stimulus from the heart itself. However these types of pacemakers can also be programmed with a specific range of heart rate and if the rate falls below or spikes above the rate that has been set at and will fire to correct the irregularity. Atrial synchronous pacemakers synchronize with the atria of the heart. They attach to either the left or right atrium and receive input and then stimulate the ventricles to contract. The trigger that causes the pacemaker to fire and cause contraction in the ventricles is the voltage produced by the contraction of the atrium. The voltage that most pacemakers sense is somewhere around 3- 8mv. With the pacemaker set up this way the heart rate (bpm) of the heart is determined by the rate of the atrial contractions.
These types of pacemakers would most likely be used to treat arrhythmias that originate below the atria, such as in the AV node, bundle of His, or purkinje fibers. There are also ventricular synchronous pacemakers which synchronize with the normal rhythm of the ventricular contractions. These types of pacemakers only use one lead to sense and stimulate the heart at the ventricle. Although they are similar to atrial synchronous pacemakers they have less sensitivity to the stimulus that it receives from the ventricle. This means that they would need a larger stimulus than atria synchronous pacemakers to be able to sense the heart’s contractions and respond.

**Sensing Circuitry**

Pacemakers are able to sense the impulses of the heart due to a sensing portion of the pulse generator circuitry. This portion of the circuitry receives the signals from the heart and measures the amplitude and frequency of the signal to determine if the sensed impulse is actually originating in the heart, and whether or not to respond to this stimulus. The relationship between the amplitude and frequency create a filtering process that sorts through the received stimuli. If the stimuli is in a certain range of frequency, set by the pacemaker, and has a high enough amplitude it will then be received by the pacemaker. This range of frequency typically will fall around the magnitude of about 25 Hz, depending on the particular pacemaker model. Since the depolarization and repolarization of the heart happens so quickly these measurements are made on the order of milliseconds (ms). A stimulus from the heart that is strong enough to be sensed must be at least .5mV/ms based on a calculated slew rate. A slew rate is described as the change of the voltage over an interval of time, this rate can be calculated with the formula \( \frac{\Delta V}{\Delta T} \). This is just a brief overview of the many different components at work in the sensing circuitry of a pacemaker.

**Pacing Circuitry**

The pacing circuitry of a pacemaker is the component that delivers the electrical shock to the heart causing it to contract and correcting the abnormality in the rhythm. This electric impulse is sent to the heart by a capacitor and this signal gradually fades over time, dropping in voltage as time passes. The voltage also drops as it passes through the wire that connects the pacemaker to the heart due to the wire’s particular impedance, which can be found using the formula \( V=IR \), also known as Ohm’s law. Due to the loss in voltage over time it is important that the stimulus that is delivered last long enough and have enough voltage behind it to have a notable effect on the heart. The relation between length of the pulse delivered and the strength of the voltage and current delivered can be expressed by the formula; \( E=\frac{V^2}{R \times t} \). E in this formula represents the energy delivered, \( V \) is the voltage, \( R \) is the resistance and \( t \) is the time interval set in ms. Proper programming balances these variables \( V, R, \) and \( t \) in order to ensure that the pacemaker will deliver a strong enough stimulus for the pulse width it is giving to effectively stimulate the heart.
Conclusion

Pacemakers help to solve medical issues with the heart that are not otherwise easily treated. They help to control many different types of arrhythmias and help to prevent the onset of life threatening electrical conditions like heart block and many others. Over time pacemakers have gotten smaller, easier to implant and regulate, as well as more reliable and sophisticated. I believe that pacemakers will continue to improve throughout time as technology improves and the ability to program even more complex processes in pacemakers becomes greater. The ability to apply and understand the relationships between the different components and formulas that create an efficient unit are the basis of quality when it comes to pacemakers.
Figures

(Source 1.) Anatomy and blood flow in the heart.

(Since each heartbeat, the electrical impulse begins at the SA node, located in the right atrium. The SA node produces the electrical impulses that set the rate and rhythm of the heartbeat. The electrical activity spreads through the walls of the atria and causes them to contract.

(Atrioventricular (AV) Node

The AV node is located between the atria and ventricles and acts like a gate that slow the electrical signal before it enters the ventricles. This delay gives the atria time to contract before the ventricles do.

(His-Purkinje Network

This pathway of fibers sends the impulse into the muscular walls of the ventricles and causes them to contract. This contraction forces blood out of the heart to the lungs and body. The SA node fires another impulse and the cycle begins again.

(Source 9) Electrical system of the heart
A basic pacemaker

Figure 6-1: Basic components of the ECG complex.

(Source 3) P, QRS, and T waves on an ECG


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Drainage Design and Hydraulics Vs Hydrology

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ABSTRACT

This report reviews the essential dynamics of drainage design and the need for these designs to divert the rainfall in an efficient manner. Surveyors, the people who observe the land, are the ones in charge of making sure that the appropriate planning and design are used in the project. Hydrology, the factors that affect the water runoff in a storm, are what surveyors are trying to work with vs against to keep the runoff at a minimum. Hydraulics, the conveyance of storm water, is the fundamental practice that is used to control hydrology. However, the Arizona State Legislature has many laws regulating how a surveyor is able to use certain techniques in managing the rainfall, increasing the difficulty of hydraulics. Some techniques; dams, canals, culverts, fords, and berms are used in an attempt to properly maintain the mass amounts of rainfall while still upholding the laws placed to protect the environment. There are many complex equations to determine the amount of rainfall on an area, though these equations are far too complex to understand without the knowledge of an engineer present. This paper is more based on the design and definitions of highway drainage, and what it takes to successfully drain mass amounts of water or aqueous alike.

INTRODUCTION

The Romans, in the year 600 B.C., are credited for constructing some of the most sophisticated drainage systems of their time\(^1\). Their drainage consisted of an open flow stream of water that would carry the fecal matter of the Romans, as well as the debris from the street\(^{[1,2]}\). Some people say that the Romans were more than just inventors, they were innovators\(^1\). The fact of the matter is that the challenges the Romans experienced are still faced today, with greater intensity too. If a civilization cannot properly drain excess water, from their cities, then that civilization will face problems with sanitation and increasing flood plains\(^3\). This paper will explore who sets the standards for flood control, what hydrology is, and how hydrology affects hydraulics.

*Flood Control District of Maricopa County*

The Maricopa County Flood Control District (FCD) is in charge of managing and preventing
floods from entering heavily populated areas\textsuperscript{[3]}. About 60 years ago the district was formed as a response to all the flooding that Arizona experienced over the years. The main purpose of the FCD is to identify flood plains, areas that are prone to flooding due to location, and help residents to know what to do\textsuperscript{[3]}. The FCD, alongside other programmers, created a public database (Flo-2D) giving a detailed map of all flood plains in the Maricopa County (see image 1). Tom Loomis, an engineer with 26 years of experience and special projects manager for the FCD said “[they] do everything from constructing dams and canals to informing home owners of needed flood insurance on their property” (Loomis, 2014). The FCD does a vast amount of work to keep the people informed on what areas are prone to flooding but that is not where they stop. What good what it do if the people merely knew that their houses were in danger of a flood, but had no idea when the flood was coming? This is where the FCD steps in and publicly warns the residents of when a flood is coming\textsuperscript{[3]}. “[They] have an array of weather instruments, from doppler radars to rain gauges, all over the county” (Loomis, 2014).

A big role that the FCD has is to publicly record the amounts of rainfall and flooding that has occurred throughout the year\textsuperscript{[3]}. They provide pdf files of all the data collected from their rain gauges and condense the information into tables and graphs. Since their graphs are published for the public to see, the free flow of information can help other surrounding districts identify areas of interest. The FCD also has many manuals on hydrology, factors that affect the runoff of storm water, and sets the standard in how to legally drain the storm water\textsuperscript{[4]}. “[They] work with organizations, such as The Arizona Department of Transportation (ADOT), to oversee how they can drain storm water without effecting the natural drainage system” (Loomis, 2014). Since it would not make any sense to simply dump all the watershed onto another county's area, ADOT has to let the water drain into the soil naturally. That can be harder than many would think. For example an engineer might ask for assistance from the tech support of the FCD to address a problem in their local highway system but finds out that there is nothing they can do. What do they do now? The only thing that can be done is alert the drivers when the area is flood because they could not push that water elsewhere. It would be against the
environmental law to push that water to other areas because that could seriously affect the balance of how the existing natural drainage functions. The FCD works alongside other organizations besides ADOT, however ADOT is the main factor in Arizona's highway system. ADOT has their own engineers and consultants working on the drainage systems of their highways, making them not as dependent on the FCD \[5\]. What affects the way an engineer goes about dealing with the drainage of storm water is hydrology. Engineers want to essentially work with the topography, the surface features of a watershed, in order to make the process of drainage easier and sustainable \[5\].

**Hydrology**

Hydrology is the reason we need drainage, in order to keep areas of the state above water. Many do not see the physics involved in their surroundings, but it is physics that gives water the destructive property that can lead to serious problems. Streamline, a consistent flow throughout the path of the liquid, is the safest way that a liquid can travel \[6\]. However this method still possess numerous challenges, such as volume flowrate, pressure, viscosity, and many more issues. Volume flowrate, given by the equation \( Q = AV \), where the area \((A)\) of a system is multiplied by the velocity \((V)\) of the liquid \[7\] is a perfect example of water in motion. The equation shows conceptually that if there is a high velocity and high area of water, such as in a severe storm, the \( Q \) (volume flowrate) will yield a lager value. This kind of volume flowrate can cause pressure issues, which was discussed earlier, and can lead to drainage pipes bursting or flooding over. Viscosity, the internal friction of a fluid, is also another problem that can affect how well a drainage system can effectively divert the water \[6,2\].

Those are just some problems that arise from the properties of water, so what are some challenges that appear from the topography of Arizona? Starting with how much water will fall on a given area of land leading up to what drainage system one should use. According to ADOT's manual of hydrology, one can calculate this rainfall using the equation \( Q = CiA \) \[8\]. In this equation \( Q \), the maximum rainfall available, is the product of the runoff coefficient \((C)\), the intensity of the rainfall \((i)\), and the drainage area \((A)\) \[8\]. This equation is just the general form of how to calculate the rainfall on an
area, but there are many other factors that can change any one of these numbers. For example slopes of the surrounding terrain, shrubs or trees coverage, and the soil itself can vary the outcome. For this paper, the location being the campus of Paradise Valley Community College, the assumption will be made that the surface is flat and that the foliage is negligible. Thus for the given circumstance that is given (A) is 9.106 acres, the (C) will be 0.64, and the (i) is 5.8 in/hr. When the numbers are pushed through the equation the predicted rainfall is roughly 34 cfs (cubic feet per second). With the rainfall calculated, the next step would be to access if the runoff will be a hazard or if the soil can naturally drain the water, once all the concrete structures are built. As a generality, the answer is usually no and some kind of drainage will have to be built to manage the runoff. This paper will discuss these forms of managing the runoff in the next section.

Once again, for this example the actual equations involving rainfall losses and unit hydrographs will be neglected because an advanced knowledge of integral calculus is needed. Instead this paper will access what these calculations of rainfall losses and unit hydrographs mean in summary form. Rainfall losses are based on a couple of factors, which include the quality of the soil and the surrounding foliage. The quality of the soil depends on the moisture of the soil and the vegetation that is in the area too. Hills and any slopes in the area of interest also effect the rainfall losses, alongside any pavement structures. Although these numbers are very small, when the area of interest is in acres the values accumulate in mass amounts. The unit hydrograph is the time, in hours, that a watershed takes to drain excess water to another watershed, where the process starts over. The next step is to create a watershed diagram that shows the waters movements and the small channels, also known as reaches. Problems that arise from watershed is erosion, a natural occurring decay of the surface landscape, which can be intensified by industrialization.

Natural erosion usually develops over a course of years, but with the constant construction on the land erosion can be up to 10,000 times greater. There are four common types of erosion; splash, sheet, rills and gullies, and channels. Splash erosion is where erosion occurs when rain droplets impact
the soil and loosen the sediment, causing erosion\textsuperscript{10}. This kind of erosion is less affective on the environment, but can still lead to greater problems when a torrential downpour with heavy wind occurs. Sheet erosion are small streams of surface water that carry sediment down a given path, usually only for 100 feet because after that they become rills and gullies\textsuperscript{10}. Sheet erosion is a low velocity stream that does minimal damage to the surface soil, however can lead to other erosion problems in the future (see image 2). Rills and gullies are a combination of splash and sheet erosion\textsuperscript{10}, and can cause the erosion that most people think of when asked for a definition (see image 3). What makes erosion so corrosive to the soil, is the velocity that it can reach. Manning's equation, given by the expression $V = (M/n) R^{2/3} S^{1/2}$, will give the velocity ($V$) of a channel with runoff flowing in it\textsuperscript{4}. For this equation the variables are as follows; $(M)$ is a constant of 1.486, $(n)$ is the Manning's constant, $(R)$ is the Hydraulic radius, and the $(S)$ is the friction gradient\textsuperscript{4}. This equation is needed because depending on the velocity of the channel or stream flow, one can pick the best management practices (BMP's).

Some of these BMP's can include minimizing disturbance areas, preserving and installing vegetation, installing bio-fabric matting, reduce runoff to and from the site, use storm drain traps, and temporary pavement\textsuperscript{10}. These measures are for the protection of the surface soil and can end up cutting the cost of a construction project in the long run. Minimizing disturbance area will consist of only using heavy machinery when needed and timing the use of machines\textsuperscript{10}. Preserving and installing vegetation will strengthen the soil around the area of interest and provide protection from splash erosion. Bio-fabric is a form of embankment control and gives the top soil a structural foundation (see image 6). Reducing runoff to and from the area of interest is simply keeping water from increasing its velocity, lowering the chances of erosion. How this is done is by blocking the storm drains with water traps. Lastly the paving of the work site can reduce the dust erosion and can help keep the water flow from eroding the soil. Now that erosion has been minimized by simple and temporary means, contractors look for more complex and permanent solutions. These solutions come from the word hydraulics, which means the transfer of water in pipes or open channels\textsuperscript{4, 2}. These methods are more
complex than just planting vegetation and keeping the velocity of the water low. The reason that hydraulics is so complex is that the velocity of the water can increase, while the erosion will be kept to a minimum.

Hydraulics

It is no doubt that without drainage one will find themselves in a world of problems. Problems such as erosion, flooding, and even sanitation issues arise from the failure of drainage\[12\]. So what types of drainage are out there to successfully drain and control mass amounts of storm water? First up is the basic channel, which is a man-made or natural occurring conveyance of water with gravity being the driving force\[13\]. This kind of drainage usually holds the most water, and can convey it to a desired location fairly quickly. Such a system is usually man-made and well maintained because natural occurring channels erode the surface soil, while man-made channels protect against erosion\[13\]. This being said, a maintenance road comes standard with these channels because these structures are high-maintenance\[11\] (see image 4). Since these structures are transporting such large volume of water a main concern is the pressure that is applied to the bottom of these channels. To compensate for that pressure on the bottom of these systems, a text tile base is placed underneath the channel to help the channel settle while it has water flow\[11\]. Ditches are like channels, except for the fact that ditches usually smaller and that ditches do not have a definite structure to them, as they connect channels together\[13\]. These can also be used alongside roadways to help convey the storm water into catch basins and storm sewers. When people think of the word ditch, most do not think of a systematic drainage feature, however ditches are more commonly used than one would think. A catch basin is a chamber or well that takes in the storm water before it goes into the sewer\[11\]. A catch basin will act like a filter for the sewer, by “catching” the debris that fall into the drain such as sediment and sludge\[2\]. These catch basins are not seen by people, as they are usually buried in the ground and underneath street curbs\[11\]. There are more than one forms of a basin that are available, commonly called a detention basin, this kind of basin actually holds to water till it is safe to release the storm
water into the sewer\cite{11}. A more sustainable detention basins, such as in a bio-retention system, will use
the landscape and interlocking pavement to funnel and filter the storm water before it reaches the sewer
pipes\cite{14}. Sustainability, in engineering, is the use of environmentally friendly and low maintenance
materials to build something\cite{2}. As the world moves from using nonrenewable resources, and looks
towards sustainability, these bio-retention basins are becoming more popular.

One of the main purposes of a drainage system is to keep storm water off the roadways and off
properties, so it would only make sense that if one could not drain the water fast enough a pump station
would be needed. A pump station is exactly that, a structure that helps drain low lying lands that are
prone to flooding\cite{11}. Some of the down sides to this approach is that the debris will tend plug up the
drains not allowing the pump station to do its job\cite{14}. Some of the newer pump stations are submersible,
meaning that they can actually by buried in the ground within a structured building. This concept is
good idea for sustainability because the underground pumps eliminates the need for an above ground
structure to pump the storm water\cite{13}. However the pump themselves require fuel, which is a
nonrenewable resource, and for that reason this type of drainage is not sustainable\cite{11}. These
underground pumps can also provide 350-hp, pumping up to 50 million gallons per day (gpd)\cite{14}. This
kind of suction is what it takes to successfully drain storm water fast, so that drivers and property
owners, can be worry free from flood damage. Another good example of sustainable draining would be
the use of Storm Trap. A Storm Trap stores the storm water which can be used to irrigate the area or
other environmentally conscience solutions\cite{14}. These storm traps can be made of concrete or by a
goecellular formation, made of a high density polymer\cite{12}. The fact that the underground structure can
be used as a water detention system reduces the footprint of the system, which allows it to self-sustain
without damaging the environment.

When pump stations or detention systems next to the roadway are not an option, culverts are the
way to go to drain and control storm water. Culverts are structures that transfer water, usually from
natural channels, under a roadway to another channel\cite{11} (see image 5). However culverts have many
issues that need to be paid attention to otherwise the culvert will not work. First off the culvert must be designed based on the “geometry of the roadway embankment” (Hydraulics Manual). This means that the culvert must approach the embankment naturally, because erosion will be a big concern if the water enters the culvert at awkward angles\textsuperscript{[13]}. Next the culvert must be of sufficient size to transport the flow of water in a flood situation\textsuperscript{[11]}. Culverts have to accommodate the estimated rain fall calculations and any more characteristics that can affect the flow of water in the channels, such as slope\textsuperscript{[13]}. Well-built culvert do not disrupt the natural contours of the natural channel, rather acts as an extension of the channel itself\textsuperscript{[13]}. The next parameter to consider is the presence of debris and sediment, such as rocks and foliage\textsuperscript{[13]}. The presence of debris increases the potential factor of erosion, and it is a lot easier to filter out the debris earlier on in the channel versus once it drains into a lake, river, or any other retention ditch\textsuperscript{[11]}. Once these factors have been considered, one can start the building of a culvert. The first thing to know is that the bottom of the culvert should also be the bottom of the channel of water\textsuperscript{[13]}. Once the area has been excavated the proper culvert embedment, which is a solid frame of dirt to surround the sides of the pipe or box style culvert\textsuperscript{[11]}. For pipe culvert a one foot span should be left between the top of the pipe and the surface of the roadway, while a box style culvert can encroach the roadway\textsuperscript{[13]}. The reason that the box style culverts can lead up to the surface of the roadway is because the box culvert is usually made of reinforced concrete, while the pipe culverts are made of weaker metals.

Most of the water in these drainage systems can be reused as irrigation on the same area that the drainage concentrated it from. So what happens to the water that does not get reused, but rather travels though the storm water sewer system? The water that can be soaked up by the soil is known to be pervious, while the storm water that gets trapped on areas with no soil (concrete, roofing, roadways, and etc.) is said to be impervious\textsuperscript{[2]}. The water trapped as a result of impervious surfaces enters the sewer system in a series of piping. These pipes are regulated by the velocity and other conditions that will yield the best possible outcome of productivity\textsuperscript{[14]}. The size of the piping is critical for the
conveyance of the storm water, since the volume of the storm water cannot surpass the volume of the piping. At the same time, making the sewer pipe diameter too large will just waste money and space under the streets that can be used for utility lines and other needs\textsuperscript{14}. The size of the pipe is determined by the rainfall calculation done in the very beginning of any hydrology study\textsuperscript{11}. The roughness of the pipe helps with the velocity of the water in the pipe, and determine the lifespan of the pipe\textsuperscript{13}. The velocity of the water in the sewer drains is important because high velocity water can create pressure in the pipes as well as making the sewer system harder to manage. Bernoulli's equation equates the pressure of a liquid to the speed and elevation of the fluid\textsuperscript{6}. In a state like Arizona, where the elevation does not really change significantly, the velocity would have the greater effect on the pressure of the pipes. The maximum velocity that the water is allowed to flow is 2 feet per second, which is roughly 1.5 miles per hour\textsuperscript{11}. At this low velocity the water can be managed properly, and will allow the drainage systems to do what they were created to do. After the water enters the drainage system it is either carried to a treatment plant, where it is treated and filtered, or it is drained into the largest available water supply\textsuperscript{15}. These large water supplies consist of, but are not restricted to rivers, lakes, or oceans\textsuperscript{15}.

A big part of drainage is where does all the water go and what regulations are there to prevent pollution from reaching the water? There are many environment hurdles that a drainage team has to overcome, before they can just start dumping storm water into open bodies of water\textsuperscript{16}. Without these laws in place, the water quality would be so poor that the populations’ health would be at risk\textsuperscript{12}. The way that a state is allowed to drain storm water into bodies of water is through the Clean Waters Act (CWA)\textsuperscript{17}. The CWA regulates the amounts of disturbance that the water can have and orders the state to track the pollution in the water. The state will have to track which areas have the most pollution, they then will have to come up with ways to control the pollution\textsuperscript{16}. The clearances allow some pollutants to be drained in the lakes or river, but will request a chart of what pollutants and how much is in the water\textsuperscript{12}. The CWA also is involved with the maintenance of the drainage systems, to ensure
that no pollutants are finding their way into the water\textsuperscript{[17]}. ADOT and any other department of transportation in the country has to go through certain steps that include collaborating with many agencies to assure that the workers out in the field are properly doing their job to reduce the storm water pollution\textsuperscript{[17]}. With the proper training and conveyance of information, the construction sites can limit the amount of pollution that they put into the system\textsuperscript{[16]}. The reason for all these types of drainage systems is to ensure that the state can control the water runoff and can properly distribute it back into the water cycle. These systems not only allow the state to concentrate the flood water, to keep the population safe, but also make it easier to filter out the harmful pollutants from industrialization\textsuperscript{[17]}. Not many people know how the water goes from the streets, to the large bodies of water, and back to the sinks of those that just used the water. However by informing the people, the risk of polluting the water supply decreases and the water quality can improve with everyone drop of water.

Conclusion

As you can see, the importance of drainage is critical for the advancement of the roadway design. Many people do not think about what it takes to properly build a roadway, and that is why this paper is so informative. The flooding that occurs during a year will not stop from coming; however the protection that is applied to our streets can reduce the affect. The rain will fall no matter what we do here on earth, but we can help control how the rain runoff will affect a civilization. As hydrology advances, we will better understand how rain affects the surface land as well as the soil underneath.

Using hydrology as a way to estimate the rainfall is the first step in drainage design, however this process is very tedious and cumbersome. We need to have a simpler way to determine the amount of rainfall that will occur in a specified area. I understand that if the process was too easy that anyone could do it, and that would mean that there would be no need of an engineer at all. On the other hand, the future of hydrology would have to improve on the equation process of finding the average rainfall. By making the equation easier and more intuitive, the error will most likely decrease. The more accurate the calculation is, the easier it will be to find a proper drainage system that will work.
Hydraulics is what we actually use to convey the water in more manageable ways. The good thing about hydraulics is that it can transform in any way that you can imagine. The applications of hydraulics can fit any circumstance that is presented. Although some hydraulic systems are more complex than others, the end result is that same. When one hydraulic system does not work there are at least 2 others that can work. The future of hydraulics is threefold, one is the design type, the other is erosion control, and lastly is the sustainability of the systems. The fact that there are so many types of drainage systems out there gives it the chance to advance and produce more creative forms of drainage.

For example in the future we could see the drainage designs more transparent in the environment. This means that we will be able to use more underground systems so we do not have to deal with the sight and usage of land to drain water. As the future advances the style of the systems, we will be able to drain more water faster than we could image. Another example of drainage improvements is in the piping of the water. As science reveals more subatomic properties of metals, we can see the friction of pipes more exact and affective.

Erosion control has been a big hurdle that the engineers have been able to affectively solve, however erosion still occurs and needs to be maintained constantly. What we need to do is come up with erosion control that is self-sustaining, or requiring no maintenance. Erosion can be an engineers’ worst enemy, but can be solved and dealt with never the less. Some erosion techniques are more responsive than others, but the future could hold the key to long term erosion control. An example of long term erosion control today is to plant foliage in the area that need to be controlled and let the foliage disturb the erosion of the top soil. I think that in the future this will be more commonly used and will be the first line of attack. Alongside foliage the future could hold artificial ways to control the erosion, such as netting that gives the ground surrounding the water flow a structure making it stronger. Either way, the use of artificial materials that can help the environment, brings up the problem of pollution. The most important advancement that will change the future, is sustainability. Sustainability in engineering we know is something that is becoming more popular, however with the need of more
maintenance-free systems in the future lies there.

Sustainability is not only good for the environment, by providing biodegradable or natural solutions, but it can also be good for the wallets of the consumer. I think that with all the talk about reducing the carbon footprint on earth, we need to look towards more economic and environmental solutions that will work. I believe that most people would rather see a more natural scenic view, than a big structure made out of concrete and lacking any geometrical design. The beauty in sustainability is that it is so promising because it can be the best solution of problems resulting in drainage, as well as other areas of engineering.

In the future of roadway design, and engineering in general, we will see more advancements towards sustainable systems. However in order to accomplish this the funding needs to be behind the movement. Without the proper funding the advancement of engineering will not move, so we need to get the finances in order first. I am aware that the current state of the country is not the best it can be, but we need solutions and less political interference. The only way that we will advance to the future, is if the politics gets out of the way of improvement.

Acknowledgements

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Image #1
Flo-2D is software that provides an in-depth look at the flood plains of Maricopa County.

Image #2
This is a picture of some rills and gullies that have occurred in a flood plain (which can also be seen in image 1) near Scottsdale, Arizona.
Image #3
This illustration is a dry stream bed that has caused erosion to the surface soil and shows the potential that running water has on soil.

Image #4
This image is of an open channel that is located in Scottsdale, Arizona. The service road runs alongside the channel and there is construction in the background, unrelated to the channel.
Image #5
This culvert is demonstrating a geometrically flattering contour that the channel can approach more naturally than others.

Image #6
This shows the bio-fabric matting that prevents erosion on embankments of natural channels, streams, and hill sides.
Cited References


Moons of Planets

Kate Larsen

April 2015

AST 111: Solar System

Professor J. Weitz

PVCC
Most planets have moons, in fact the only planets without moons are Mercury and Venus. Once past the terrestrial planets the number of moons a planet has jumps from single to double digits even. But it’s not just planets that have moons, but asteroids and dwarf planets as well, and all of these moons have interesting, unique features and characteristics to discover and study.

Everyone is familiar with the Earth’s moon. Whether that familiarity is on the level of general acknowledgment of its existence or more in tune with the details like any astronomy student becomes. But just because people are aware of our moon doesn’t mean they are aware of other moons. A moon is a natural satellite, an object in space that orbits another object is space that is not the sun. There are many moons in our solar system besides our own. There are 147 confirmed moons orbiting planets, and 17 more awaiting confirmation not including moons orbiting dwarf planets or asteroids (NASA, 2015).

Moving from the sun outward towards the edges of our solar system, the first group of moons we encounter are the moons of the terrestrial planets. The terrestrial planets, in order, are Mercury, Venus, Earth, and Mars. While neither Mercury nor Venus have any moons Earth has one and Mars has two.

Earth’s moon lies 239 thousand miles from earth and completes an orbit every 27 earth days and it is the only object in space that has been visited by humans. Because it is the closest body to Earth and the only one we have visited, we have more data on Earth’s Moon than any other. Working our way toward the center, we start with the surface of the Moon which is covered in craters, mountains, volcanic flows, and other interesting geological features. In Murphy and Bell’s article *Dating the Moon: Teaching Lunar Stratigraphy and the Nature of Science* they tell us that, “The relative ages of the Moon's highlands and maria can be determined by counting the number of craters per unit area” (Murphy & Bell, 2013). The next layer, the crust, is almost completely pulverized. *Surprises Beneath Moon’s Surface* published by *Nature* says that on average, the Moon’s crust is only 34-43 kilometers thick, much thinner than previously estimated and that these findings indicate that the Moon took a much greater beating in its early years than suspected. At the Moon’s center is a solid iron core spanning 480 kilometers across that is surrounded by layers of liquid iron and molten rock and magma. This data is relatively recent as even though the seismometers that recorded the moon quakes for the data were installed in the 1970’s, the wave patterns were too complicated to understand until recently thanks to modern computers (R.C, 2011).

Mars is the only other terrestrial planet to have any moons. Its two moons, Phobos and Deimos, like most bodies in our solar system, are named after figures in mythology and are two of the smallest moons in our solar system. The two moons were discovered in 1877 by Asaph Hall and are similar to Earth’s moon in that they are also heavily cratered and always show the same face to their planet. Deimos is the smaller of the two, orbiting Mars every 30 hours while Phobos orbits Mars much faster at three times a day only 3,700 miles above the surface of the planet. The official NASA website says that, “No known moon orbits closer to its planet”. As it orbits, Phobos is slowly drawing closer and closer to Mars and it is predicted that within 50 million years or so the moon will either collide with Mars or break up and form a ring around the planet (NASA, 2015).
The next thing to be encountered as we move outward would be the Asteroid Belt. Asteroids are solid, rocky, and irregularly shaped objects that do not have atmospheres and cannot support life as we know it. In 1993 asteroid Ida and Dactyl were the first asteroid-moon pair to be discovered. Now over 150 asteroids are confirmed to have companion moons, some even have two.

![Ida and Dactyl](image)

Jupiter is the first of the Jovian Planets and has 50 confirmed moons and 17 more awaiting confirmation. Jupiter’s moons not only have highly elliptical orbits, but many also orbit opposite the rotation of the planet. The planet’s four largest moons were discovered in 1610 by Galileo Galilei and are called the Galilean Satellites in his honor. They are named Io, Europa, Callisto, and Ganymede. Io is extremely volcanically active, so much so that its surface is covered in sulfur. This moon is so volcanic because as it orbits Jupiter, the planet’s gravitational pull causes ‘tides’ that pull and push the moon generating extreme amounts of heat (NASA, 2015). Europa, on the other hand, is the opposite being covered in thick ice under which there is thought to be an ocean. There is so much water on Europa that it is believed to have twice as much as Earth. Callisto is the outermost of the four Galilean Satellites and the third largest satellite in our solar system. Callisto’s surface is ancient and heavily scarred due to impacts, and because of a lack of geological activity it is thought to be a dead world. Finally, Ganymede is the largest satellite in our solar system, it is even larger than the planet Mercury. It is also interesting to note that, “is the only moon known to have its own internally generated magnetic field” (NASA, 2015).
Saturn also has a ridiculously high moon count having a total of 53 confirmed and 9 waiting for confirmation. Discovered in 1655 by Christiaan Huygens, Saturn’s largest moon is Titan, the second largest moon in our solar system stretching 3,200 miles across. Titan is also the only moon with a thick atmosphere, its nitrogen rich atmosphere extends about 600 kilometers into space, ten times what Earth’s atmosphere does. Saturn’s moons are rather interesting, for example some moons, like Phoebe, orbit in the opposite direction of the larger moons, and the two moons Janus and Epimetheus switch orbits whenever they pass close to one another. There are even moons located inside Saturn’s rings that are called shepherd moons because they help keep the rings in line as they clear their path of the dust and debris that lie in their paths.

Uranus is the planet that breaks the mold, it spins on its side, it was once named George, and instead of mythology its moons are named for characters from Shakespeare’s plays. Uranus has 27 confirmed moons, some of which are extremely difficult to see as they are not only tiny (only 8-10 miles across) but pitch black as well. It doesn’t help that they’re 1.8 billion miles away from the Sun. Miranda, the innermost of Uranus’s moons is the Frankenstein of moons having an extremely unique surface. Not only does it have fault canyons 12 times deeper than the Grand Canyon, but it also has, “terraced layers and surfaces that appear very old, and others that look much younger” (NASA, 2015). Uranus, like Jupiter, has several shepherd moons two of which, named Cordelia and Ophelia, keep Uranus' thin outermost ring well defined.

Neptune’s largest moon, Triton, was discovered in 1846, only 17 days after Neptune itself was discovered, by William Lassell. Triton is the largest of Neptune’s moons, being equal in size to the dwarf planet Pluto. It is also one of the coldest objects in the solar system sitting frigidly at -400 degrees Fahrenheit (NASA, 2015). Also of note is that like Phobos, Triton is slowly being pulled towards Neptune due to the planet’s gravitational pull and in millions of years it may break up to form a ring around the planet. Neptune’s second largest moon, Proteus, and 5 of the other 13 moons of Neptune weren’t discovered until the Voyager 2 mission because they were too dark to see from so far away with the technology that was available at the time.
The only group left in our solar system with moons would be the dwarf planets, the most popular of which would be the former planet, Pluto. Pluto has 5 known moons named Charon, Styx, Nix, Kerberos, and Hydra. Charon and Pluto are occasionally referred to as a double planet system because Charon is almost half the size of Pluto and orbits so close to its planet. The two other dwarf planets with moons are Haumea with its moons Hi'aka and Namaka and Eris with its moon Dysnomia.

Planets, asteroids and dwarf planets all have moons. Moons that do interesting strange things like orbit backwards and switching orbits with their neighbors, and sometimes maybe even holding the possibility of supporting life of some kind.
Works Cited


Roof Gardening and Energy Consumption

Marjan Lavasani

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Dr. Durandet
Abstract:

This research paper is about the green roof technology, how can affect the consumption of energy. Energy can contributes of several things like water, environment, and cost of fuel. There is lot of benefits for roof gardening. One of important is effective energy efficiency measure to reduce the cooling the building by provide shade and remove heat in summer and heating in winter. There are more benefits for green roof like ecological and landscape value to air pollutions and nice view to the city. This paper tries to explain the benefits of green roofs and compare the cities or countries that used this method also explain about heat island and affect to the communities. Explain method of planes used in State of Arizona for green roof and saving energy.

Introduction

A green roof is a method that used strong isolations in top of roof and used some vegetative layer grown on a rooftop. The kinds of vegetative used in top of the roof depend of climate the area. There are several logics for kind of vegies is good for roof top or trees are good too. The reasons green roofs change the temperature in summer cooler that it provide shade and remove heat from the air through evapotranspiration, also water bodies help reduce heat in urban spaces reducing temperatures of the roof surface and the surrounding air. Green roofs are getting more popular and most government’s plane for consuming better energy used green roofs. The cost of energy decreases then it can help household and business increase the income. Most states used green roofs the taxes money increase and governors can do more plane like health care plane for states.

Most cities in the United States such as Salt Lake City, Sacramento, Houston and Chicago have a chance to have green roof and save trees for energy. A survey shows the green roof increasing 25% from the 2004-2005. Germany and China are two most countries in the world that have green roofs. Germany’s Guidelines on green roofs recommended the medium used on a green roof usually retains from 30% to 60% of water by volume.

Experimental measurements of energy balance models used by collect data from the cities. Of above in the United States with different climates and used to estimate sensible fluxes in cities with different climate zones across the United States. On average the black roof and black roof with PV have the highest peak daily sensible flux to the environment, ranging from 331 to 405 W/m2? The addition of PV panels to a black roof had a more effect on the peak flux, but decreased the total flux by an average of 11%. Replacing a black roof with a white or green roof resulted in a substantial decrease in the total sensible flux. Results indicate that if a black membrane roof is replaced by a PV-covered white or a PV-covered green roof the corresponding reduction in total sensible flux is on the order of 50%. This experimental method developed more with thesis and more testing then evaluating the relative impacts of roof design choices on the urban climate and heat island mitigation.
Benefits

Reduced energy use roof gardening because builders have to use the best isolation in roof that can several layers, also green roofs absorb heat and act more layers as insulators for buildings. It can reduce energy in summer for cooling and winter for heating. Researchers studies for several cities in different climates in summer and winter. They measures temperatures in court yards and green roof also with different kinds of variations. The result shows in summer time in green roof most has a 4.7 °C lower air temperature and in winter air temperature in green roof is higher than regular roof. (Taleghani, M.,2014)

The humidity can affect the temperature in winter and summer because in water body has more evaporation and change the environment.

Green roof or vegetation can help to reduce air pollutions with used dioxide carbon and reduce oxygen

Greenhouse gas also change the climate and get atmosphere warmer with reduce many gasses like carbon dioxide, Methane, Nitrous oxide and other gasses.

These are indirect affect in human health by decreasing heat which transfer with heat waves from the roof of building. The waves transfer from outdoor to indoor when isolation is strong and several layers avoid entering lots of waves and heating stress. (Image 1)

Green roof can help reduce storm water in the urban environment that collects the rain and runoff water from roof and damage of water drops. Water management, it is confirmed that green roofs reduced storm water runoff in any climate. Mediterranean climate is example of runoff water in exact period.

Green roof can change the view of city with green design and new habitat for so many species. This is also indirect effect the psychology of quality of life for human.

Green roof insulation more protects building from sound and waves.

Green roof means more space with nice view then the real state value are increase for these kind of building.

The temperature of green roofs on top of public buildings with solar radiation through the vegetation layer is depend to the thermal insulation and structure of the building and panels. The measure of temperature or chart of thermal with every day show the energy reduced and saving of energy for cooling.

Solar system is the best investment because reduce the electric bill and having green roof.

Green roofs also reduce air born pollutant and dust that find on bare roofs.

Panels need less maintenance than traditional roofs.

Green roof billing energy is the same as dark roof or white roof.
Costs

Green roofs installation are more than traditional roof. According to the United States Environmental Protection Agency” estimated costs of installing a green roof at $10 per square foot for simpler extensive roofing, and $25 per square foot for intensive roofs”. The green roof work as a life time and it does not annual maintenance fee. Most traditional roof need maintenance that costs for either type of roof may range from $0.75–$1.50 per square foot every year.

A University of Michigan study compared the expected costs of conventional roofs with the cost of a 21,000-square-foot (1,950 m2) green roof and the green roof would save about $200,000 for life time. Researchers work with the other group of communities to more analyses detail of green roofs and net profits because green roofs affect several things the most important is energy or two-thirds of saving would come from energy, but the other things like water run-off, storm water and air pollution. The economic effects of energy saving are green roofs with vegetation and more trees on different climate the price are almost $250/tree/year, while the air quality regulation was valued between $0.12 and $0.6/m2 tree cover/year. Maximum monetary values attributed to noise regulation and aesthetic appreciation of urban green were $20 – $25/person/year, (Yang, J., & Wang, Z. (2014)

Heat Island

Heat island is important factor that can damage to the communities in summer season with increasing cost of air conditioning, water quality and greenhouse gas emission. Green roof can solve this problem in some cities such as Phoenix, Tuscan, and Gilbert. Arizona is good candidate states for green roofs because in summer time the population decrees and pollution increase. The waves of heat is long and too much that can hot and hotter outside of building transfer to the inside then air condition has to on with low temperatures all days. The cost of energy increase also human health decrees with high temperatures. State of Arizona has some plane of green roofs in some cities in this stat.

Arizona plan’s for green roof

Arizona plans for green roof are includes, environmental planning element, the Arizona State University Sustainable materials and Renewable Technologies (SMART) administration building, and urban forestry program. United States Environmental protection do theses plane in different cities for tries and better results of green roof. It can the best sores of energy in Arizona also it can help for decreases the cost of energy. Change the design of cities and clean air.

Environmental planning is in Gilbert Arizona for contributing the heat island and encouraging engineers to design planned for more shade also promoting education for best materials for planned. Researcher’s studies to ways that decrees energy and encourage more people move to Arizona and stay here. This
Plane is for way to cool temperature in summer time by cool roofs with more trees and vegetation also cool pavements.

Arizona State University Sustainable Materials and Renewable Technologies (SMART) researchers are developing the next generation of urban materials and advanced solar technology that reduce energy demand for mechanical cooling.

Demonstration planes are for city of Tuscan for how a cool roof reduced temperatures in the roof of building and inside and saved more than 400 million Btu annually in energy. Installation green roof and more shad help to added saving energy and with this saving do something for pavements with adding more trees and vegetation in the parking lots and surrounding the building.

Urban forestry program need more tress. Researchers encourage people to grow lots of trees that adapted with desert also help to clean the city and clean air for reduction of carbon in air. This plane began in Tucson since 1989 till 1993. In this time 70,000 trees have been grow in Tuscan area.

**Materials**

The materials used for green roofs are very important because for vegetation need water. Water need more isolation for roofs and the kind of bricks used in roof are important because large bricks had a lower water holding than small one. Estimate less than 35% then decreased growth under warm and sunny conditions. Organic materials can more help for holding water in top of the roof.

Soil moisture was low and evapotranspiration from the green roof was low then increased sensible heat fluxes during the day.

Researchers recommend also designed of rooftops should be depended of specially performance of aimed objectives. That is including of balancing energy in roof, and changing the weather in different seasons.

Black and white roofs have no effect for storing water or slowing rainwater runoff.

The drainage layers are important for green roof’s ability to absorb and retain storm water.

The greenest material used in green roofs because they made products from recycled waste materials those are plastic, wood or any type of fiber and waste from homes, the other materials from waste of factories. The materials usually have light weights.

Photovoltaic equipment used for solar system and important part is vegetative roof installation.

Some certificate company such as LEED and LEED-EB in the United States established and they help people to used and buy green roof that named Green Globes and recognized by the EPA’s energy.
Conclusion

Green roofs are new technology that fast grow or Green globe have lots of benefits for human health and economics. The installations of green roof is expensive, but is life time cycle and it does not annual expensive maintenance. Tradition roof need maintenance annually. Green roof building have more space and real state price is more than the other building. Isolation is very strong that keep away noise and heat in summer time and cool thing in winter also the layer avoid to enter some waves easily come inside the building. The scenery of green roof is very beautiful, because roof of a building covered with vegetation band growing medium. Green roof has environmental benefits; such as clean air with used carbon and some heat island gasses and air pollution, nevertheless it change the quality of water and water drop and reducing harmful storm water runoff also material used for green roof are most recycle from waste of homes and factories. The most important factor of green roof additional is energy saving. The bill of power with green roof is always very low. That is emerging a green roof experts incorporate photovoltaic to generate clean, renewable solar system. Germany is pioneer in this technology and it has lot of green roof and energy saving. The united states more cities have opportunities to have solar system and green roof also governor encourage people to have green roof ether most governor plane buildings have the green roofs. State Arizona has lots of research and experimental results for green roofs. They used green roof in different locations of the state and experience benefits of green roofs. There is lots of date that help people to find out the benefits of green roofs.
Rainfall and Roof Runoff at GRRF
(Jan 1, 2005 - Dec 31, 2005)

- Wet period: reduction = 15%
- Dry period: reduction = 90%
- Overall reduction = 28%

Sketch of an Urban Heat-Island Profile
GREEN ROOF

TRADITIONAL ROOF

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