Physics 42A, Mr. B. Panas - Course Outline

Introduction

Welcome to Physics 42A. If you are willing to put some work and effort into this subject, you are sure to find it rewarding, enjoyable and very interesting. On the other hand, if you choose not to put some work into this subject, you are sure to find it miserable, mysterious and difficult. As with physics 30A, keeping up with the material is very important, so please don't allow yourself to fall behind.

Topics

On the side counter is a book which details the content of the AP physics program. I would expect that you will look through this book, and learn everything you need to know about the course from it.

Homework

As with 30A, there is loads of homework! I will be perfectly honest with you right now: this course is not easy. When I asked last year's class what the hardest course they had taken in high school was, most of them said it was physics. (A couple said calculus, but I'm going to work you guys even harder this time to ensure this doesn't happens again). On the positive side, everyone who took it said that they would do it again. Seriously, you can expect about an hour of homework every night in physics, even accounting for the fact that I only see you every other day. If you cannot commit to this, then come and see me in private to discuss what options you have.

When you get "stuck" with a homework problem, I advise the following: First try the problem in a different way, draw a diagram if applicable. If that doesn't work try getting help from a friend. If that doesn't work try looking at the solutions book (I'll talk about this in class). If none of these help you, then (and only then, please) come to me for help. **I am always happy to help, but only after you have attempted to help yourself**. By the way, getting "unstuck" on problems is called *learning*.

Marks Breakdown

Before Exams		
1	Tests / Quizzes	80%
2	Computer Quizzes	20%

After Exams		
1	Tests	40%
2	Computer Quizzes	10%
3	Grade 11 Exam	20%
4	Practice AP Exam	30 %

1 **Tests**: When we have completed our study of a given topic, I will announce a test date. Tests are based on all the material we have covered up to that point (including "old" material that may have already been tested). A friendly reminder here that the textbook does not contain all of the material you will be responsible for - you will have to study from your notes and assigned problems. **If you are absent for a test you will receive a zero for that test**. You will be able to write the test late only if you talk to me, and convince me that you missed for a legitimate reason. The only legitimate reasons for missing a test are illness, and school sponsored events. Doctor / Dentist appointments etc. are not legitimate reasons, unless you speak to me in advance, and can convince me otherwise - in which case you should try to arrange to write the test.

Quizzes: Once in a while, class will begin with a short quiz, which will test you on recent topics covered (as recent as the last class!) This is my way of forcing you to keep up with the material. These quizzes are marked on a "right or wrong" basis with no possibility of part marks, which may be earned in tests. **If you are absent for a quiz (including quizzes missed due to tardiness) you will receive a zero for that quiz**. The zero will be changed to "omit" only if you specifically talk to me the next day you are back in class, and convince me that you missed for a legitimate reason, such as illness or a school sponsored event.

2 **Computer Quizzes**: On my computer is a program which contains multiple choice quizzes on the topics we are studying. I will be assigning these quizzes frequently, which are to be done **outside of class time**, by a certain date (usually the day before a test). The computer is available for you to do these quizzes at almost any time that I am in the room, including when I am teaching other classes - the exception is when I need to use the computer.

3 **Grade 11 Exam**: The exam is 3 hours, and will consist of multiple choice and long answer. It will be based on the material that was already covered in Physics 30A. This exam will take place at the same time that the 30A students write their exam, during the January exam block.

5 **Practice AP Exam:** This exam is also 3 hours and will be similar in format to the grade 11 exam. It will be based on the material that has been covered in Physics 42A along with what was in 30A. This exam will take place approximately 2 or 3 weeks before the real AP exam, and will be done outside of class time.

6 **Physics AP Exam:** This is not for school marks, but in my opinion is the most important of the three exams. Unlike the previous 2 exams, this one is an international exam which is prepared by the College Board (a US - based organization). This is the exam where you will prove your ability in what I have taught you. I will not see any of the exam until after you have written it, and even then I will only be allowed to see the long answer questions (not multiple choice). Your papers will be sent to the US to be graded (I am one of the people who will be marking the exams), and you (and the school) will be mailed the results. They are scored on a 1 - 5 mark scale, 5 being the best. With a 4 or a 5, many Universities will offer you either "placement" in a first year physics course (which means you can skip to 2nd year), or "credit" which means that for all intents and purposes, you have already done first year university physics. To be eligible to the write the AP exam, you must have an 80% or better in the course.

Physics 42A - Class #1

Today's class: Sections 16-1 to 16-4

Welcome to Physics 42A. I sincerely hope you are prepared to work harder than you have ever worked before. We will be moving fast, so keep up.

All electric (and magnetic) effects are the result of <u>electric charge</u>. Electric charge is a fundamental property of matter - as is mass. Unlike mass (which only comes in one kind), there are two different types of charge, named (by Ben Franklin) <u>positive</u> (+) and <u>negative</u> (-). A third charge state is <u>neutral</u> which can either mean no charge or equal amounts of positive and negative (described as having "no <u>net</u> charge").

The Law of Charges: Unlike charges attract; like charges repel.
The Law of Conservation of Charge: A <u>net charge</u> can never be created or destroyed.

Electric charge is ultimately found in only two places: electrons and protons (this is a lie but is close enough for now). Negative charge is a fundamental property of the electron, while positive charge is a fundamental property of the proton. Remarkably, they both have the exact same *amount* of charge. The amount of charge on a single proton is defined as one <u>elementary charge</u> (symbol: "e") The property of charge is designated as the letter "q" (as mass is designated with "m"). The following tables (which you will immediately memorize) should clear this up:

Fundamental Property	Designated Symbol	Unit (abbreviation)
mass	m	Kilogram (kg)
charge	q	elementary charge (e)

proton (p^+)	electron (e^-)	neutron
$m = 1.67 \times 10^{-27}$ kg	$m = 9.11 \times 10^{-31}$ kg	$m = 1.67 \times 10^{-27}$ kg
$q = e$	$q = -e$	$q = 0$

Atoms always have the same number of protons and electrons. For example: An atom of Uranium has 92 protons and 92 electrons. The net charge is therefore zero (92 protons:
$q = +92$; 92 electrons: $q = -92$. $q_{\text{uranium atom}} = q_{\text{protons}} + q_{\text{electrons}} = (+92) + (-92) = 0$

If an atom loses or gains electrons, it is not called an atom any more: it is called an <u>ion</u>.

Most objects are usually neutral. A ball, for example, typically has the same number of protons as electrons. We would say that it is <u>uncharged</u> or <u>neutral</u>. Note that there is lots of charge in the ball (very large numbers of protons, and very large numbers of electrons), but the fact that they are present in the same amount is what is implied by the words "neutral" and "uncharged".
A neutral object may become charged in a number of ways. The three most common ways are:

1. <u>Friction</u>: Charging by friction involves <u>rubbing</u> two different materials together. Since

electrons are attracted to the different surfaces to different extents, some of the electrons will tend to be transferred from one surface to the other. This leaves one surface positive (which lost electrons, leaving behind more protons) and the other negative (which gained the electrons).

2. Conduction: A neutral object may become charged by touching it to another object that is already charged. The effect is most effective if both are conductors (as opposed to insulators - see below). Electrons move from one to the other, leaving both objects charged with the same type of charge. Note that the original charged object will have less of a charge now. See figure 16-6, page 458

3. Induction: A conductor (but not an insulator) may be charged by bringing it near (not touching) a charged object. The electrons will move in the conductor towards a positively charged object, or away from a negatively charged object. At this point, the object is still strictly speaking neutral, but the charge has been separated. It is said to have an "induced charge" - positive on one side and negative on the other. If the object is grounded at this point, a net charge will remain, opposite in sign to the original charged object. Note that the original charged object has not lost any charge. See figures 16-8 and 16-10, page 459.

Note that it is the electrons that (almost always) do all of the moving around. Usually, when there is a movement of charge, you should recognize that it is the negative charges (electrons) and not protons that move. There are exceptions, such as in a fluid consisting of ions - where the ions themselves move around redistributing charge. An example of this would be when charge "leaks off" of objects. A charged object will discharge over time, as the air around it ionizes, carrying negative charges onto or away from the object, reducing its charge.

A material is described as being a conductor if electrons can move through it easily, and an insulator if electrons cannot easily move through it. Examples of good conductors are metals, while insulators are glass, plastic, rubber, etc.

Grounding means placing in electrical contact with the earth, which can be considered to be a very large conductor. Grounding an object provides it with a source of electrons (if near a positive charge, or is positive itself) or a place to get rid of excess electrons (if near a negative charge, or is negative itself).

An electroscope is a device which detects charge. See description and diagram, page 459.

For Homework:
- Read Sections 16-1 to 16-4
- Questions #1, 2, 4, 5, 6, 8
- Read ahead into 16-5 to 16-9

Physics 42A Class #2

Today's class: Sections 16-5 and 16-6

We have seen a unit for charge: the elementary charge (e) which is defined as the amount of charge on a proton (q = e) or electron (q = -e). This is far too small of a unit to be of much use under most circumstances (when large numbers of these particles are involved). The SI unit of charge is in fact the <u>Coulomb</u> abbreviated "C" (named after Charles Coulomb). "Coulomb" can be thought to be the same kind of word as "dozen" (meaning 12), and "mole" (more often used in chemistry - meaning 6.022×10^{23}). One coulomb is the amount of charge contained in 6.24×10^{18} protons or electrons. Notice then, that a single proton or electron has an amount of charge equal the reciprocal of this number.

$1C = 6.24 \times 10^{18}$ e $1e = 1.6 \times 10^{-19}$ C $= q_{proton} = q_{electron\ (but\ negative)}$

Recall Newton's Law of Universal Gravitation: $F = G\dfrac{m_1 m_2}{d^2}$. It describes the gravitational force between any two objects <u>with mass</u>. Similarly, Two objects <u>with charge</u> experience forces between themselves (ie the Law of Charges). The *amount* of force can be found using <u>Coulomb's Law</u>.

Coulomb's Law: $F = k\dfrac{q_1 q_2}{d^2}$ Where k is "Coulomb's Constant = 9.0×10^9 Nm^2/C^2 and q_1 and q_2 are in the SI unit of charge: Coulombs.

When using Coulomb's Law, ignore the sign (positive or negative) of the charge. Then determine the type of force (attractive or repulsive) or direction of the force from the law of charges. *Alternatively:* use the sign of the charges in the formula: a positive force implies repulsion while a negative force implies attraction.

You can safely ignore the stuff about "Permittivity of Free Space" found in the text. (ϵ_0) for now - use k, Coulomb's constant instead - if you're dying to know, k stems from a more fundamental constant named the "permittivity of free space" $\epsilon_0 = 8.85 \times 10^{-12}$ and $k = 1/(4\pi\ \epsilon_0)$

In situations where there are more than two charged objects, you can obtain the net force by considering them in pairs, obtaining the forces (vectors!) with Coulomb's law, and then adding the vectors to obtain a resultant (review this from last year! - Vectors can be found in the beginning of chapter 3 in your text).

Homework:
- Read 16-5 and 16-6
- Questions 9 & 10
- Problems 1-18
- Read ahead, the rest of the chapter (omit 16-10)

Physics 42A Class #3

Today's class: Sections 16-7 to 16-9

Think back to Grade 11 physics: gravitational "g". We described g as being the "acceleration due to gravity" where $g = 9.8\ m/s^2$ at earth's surface. Also, from Newton's 2nd Law: $F_G = mg$. There is another way of describing g, which is completely compatible with everything we know about it. Instead of thinking of g as an acceleration, we can think of it as being a property of space itself (pretty deep stuff - hey?)

We can describe a "gravitational field" as being a property of space itself. A gravitational field results in a force being exerted on any mass in that space. "g" is a measure of the strength of the gravitational field. Specifically: $g = F_G/m$ (the gravitational force per mass). Notice that g would have the units of Newtons per kilogram (which is exactly equivalent to m/s^2 !) How strong is the gravitational field at the surface of earth? Answer: 9.8 N/kg - each kilogram of mass at the earth's surface experiences 9.8 N of force on it. Where does the gravitational field come from? Answer: mass is the source of all gravitational fields (the mass of the earth produces the 9.8 N/kg at the surface). Specifically: $g=\frac{F_G}{m'}=\frac{\frac{Gmm'}{d^2}}{m'}=\frac{Gm}{d^2}$ where m is the mass of the earth, and m′ is the mas of an object at the surface (which has no effect on the value of g)

Now do the exact same stuff for charge: we know that charges exert forces on each other. We can describe an <u>electric field</u> as being a property of space: it exerts forces on electric charges, and is the result itself of electric charges. We will use the symbol **E** for electric field, and note that it is a vector (as is **g** though we are usually less concerned about it). Charges experience forces whenever they are in an electric field. The amount of force depends on the strength of the electric field, as seen in the definition of the electric field:

> $E=\frac{F_E}{q'}$ (Definition of Electric Field) $\quad E=\frac{kq}{d^2}$ (Electric field strength produced by charge q)
>
> Where q is a charge and q′ is a "test charge". The <u>direction</u> of the electric field (it is a vector) is defined as being the direction that a <u>positive</u> charge would experience in the electric field. The SI unit for electric field is (from F/q) Newtons per Coulomb (N/C)

Note that the right equation comes from substituting Coulomb's Law in place of F_E and simplifying, as seen above for gravitational field. Also note that the q in the right equation is the charge that is producing the electric field. If another charge (q′ - called a "<u>test charge</u>") is in that electric field, it will experience a force, as in the left equation. Test charges should be small so that they disturb the original charge as little as possible, and be positive by convention.

If there is more than one charge producing an electric field at some location, then the net electric field can be found by determining their individual contributions to the electric field (described above) and then adding the electric fields as vectors.

Field lines are a way in which the electric field can be visualized. A field line diagram tells the following information:

- the direction of the field lines shows the direction of the force on a positive charge placed in that field, as the force is tangential to the lines, in the direction shown
- the density (spacing) of the field lines indicates the strength of the field (the closer they are, the stronger the field is)

The following rules should help for drawing field lines:

4. Field lines begin on positive charges and end on negative charges (or infinity)
5. The lines are drawn symmetrically leaving or entering a charge at right angles
6. The number of lines leaving a positive charge, or entering a negative charge is proportional to the charge
7. Field lines can never cross each other, or overlap

Since electric fields exert forces on charges, it is reasonable (and correct) to assume that inside a conductor at equilibrium, the electric field must be zero. If there was an electric field inside a conductor, then the charges would move as a result of the electric force. Important consequences of this are:

For a conductor at equilibrium:

- any net charge must be found at the outer surface only, as they repel each other
- the electric field is always perpendicular to the surface of a conductor

Homework:

- Read 16-7 to 16-9
- Questions 13-16, 19, 20
- Problems 19-38, 54
- Read ahead into chapter 17

Physics 42A - Class #4

Today's class: Sections 17-1 to 17-5

I found these sections especially unclear - so concentrate on these notes more

Recall potential energy from grade eleven: it is energy "waiting to be released". For example, gravitational potential energy = mgh. An object (with mass "m") has energy, simply because of where it is located (height "h") in a gravitational field (g). If an object is dropped from a height, the gravitational field does work on the object, changing the potential energy into kinetic energy.

The same is true with charge. A charge placed in an electric field will have an electric force on it ($F = qE$). If released, it will accelerate, gaining kinetic energy. Where did that energy come from? Answer: from the electrical potential energy that the object had from being in an electric field E.

Remember that the electric field can exist in places that there is no charge: it is the force that would be exerted on a charge if there was a charge there ($E = F/q$). Electric Potential (or more simply "potential") is very similar: it is a description of points in space, describing the amount of electrical potential energy a charge would have, if there was a charge there (there need not be a charge there at all).

$V = U/q'$ where U = potential energy, q' = test charge, V= "potential" or "voltage"

Units of potential: Joules per Coulomb (J/C) which we call a Volt (V)

Note that Voltage is a scalar which makes it a lot easier to work with than electric fields

Note also that "electrical potential" and "electrical potential energy" are not the same thing. Potential energy is the amount of energy that an actual charge has by virtue of it being in a certain location. Potential is a description of that location. This can be seen in the above equation: U (potential energy) and V (potential) are clearly not the same.

When a charge q moves from a location where the potential is V_A to a location where the potential is V_B, (and $V_A \neq V_B$) then the potential energy of the charge changed. It started with potential energy $= qV_A$ and ended with potential energy $U_B = qV_B$. It's potential energy changed by an amount equal to $U_B - U_A = \Delta U = qV_B - qV_A = q(V_B - V_A) = qV_{BA}$ where V_{BA} is the "potential difference" from A to B.

Whereas we have ignored the sign (positive or negative) of charges in equations involving electric forces and fields, you must include them in equations involving energy and / or voltages. If ΔU is negative, then it lost potential energy (most likely to kinetic energy - don't forget that $KE = ½mv^2$). If ΔU is positive, then it gained potential energy (something did work on it).

Since W = Fd and F=qE, we have W = qEd. But W = ΔU = qV. This means:

V_{BA} = Ed (Provided E is constant)
where V_{BA} is the potential difference between points A and B, E is the electric field strength (constant) between A and B, and d is the distance between A and B.
This also means the E = V_{BA}/d which gives a new choice of units for electric field: Volts per metre

The above is most commonly seen in situations involving parallel plates. Parallel plates simply being two large sheets of conductor which may be subjected to a potential difference (V). Parallel plates are useful because they are a way to achieve a uniform electric field (E). Whenever you see a question involving them you should immediately recognize that the field is constant between the plates, and the above formula will work. It will not work unless you have a uniform electric field.

*** You should read section 17-3, but it is not very applicable anywhere other than in a lab.

An Electron Volt is a unit of energy (not of voltage as it's name might suggest)

$$1\text{eV} = 1.6 \times 10^{-19}\ \text{J}$$

The potential at a distance d away from a point charge q can be found by:

$$V = \frac{kq}{d}$$

By the way "ground" is usually taken to be V = 0, and other potentials are made with respect to this chosen zero point (just like in the gravitational case where "ground level" is usually taken as h=0).

For Homework:
- Read Sections 17-1 to 17-5
- Questions #1, 4-9
- Problems #1-21 (Omit 17, 19 b)
- Read ahead into 17-7 to 17-10

Physics 42A - Class #5

Today's class: Sections 17-7 and 17-9

These two sections finish off our study of Electrostatics. Note that these two sections are very straightforward, being summed up neatly in a couple of new formulas.

We have been looking at charge, which comes in two varieties (positive and negative, which attract each other). A capacitor is a device which stores charge. Usually, a charged object won't hold its charge for long. The Van de Graaff generator (that really is how you spell it) if charged, quickly loses its charge: why? Answer: the negative charges built up on it repel each other, and the charge "leaks off" via the air around it.

A capacitor can store charge for long periods of time. It works by putting equal amounts of charge "Q" separated onto two conducting sheets: positive on one (Q), negative on the other (-Q). It won't discharge as readily because the two sheets are kept close to (but insulated from) each other. The charges on one sheet "see" the nearby opposite charges, and are held in place by the electrostatic attraction.

If the sheets are flat, it is termed a "parallel plate capacitor" - as opposed to a cylindrical capacitor, where the sheets are rolled up. (Review what we said about parallel plates last class, as it applies here too). Capacitors are rated for their capacitance which is an indicator of how much charge they can store:

$$C = \frac{Q}{V}$$

C = "capacitance"
Q = charge stored on one of the sheets (amount only, so always positive)
V = Voltage (potential difference) applied to the capacitor

Unit of Capacitance is the "Coulomb per Volt" which is named a Farad (F)

For a parallel plate capacitor: $C = \epsilon_0 \frac{A}{d}$

Where ϵ_0 is the "permittivity of free space" (a constant) = 8.85×10^{-12}

A = area of one of the sheets (both same area)
d = distance separating the plates

Capacitor's stored charged has energy: $U = ½QV$ Note: Q or V in this formula can be removed by solving for one of them in the first capacitor equation: C=Q/V, and substituting to get $U = ½CV^2 = ½Q^2/C$

For Homework:
- Read Sections 17-7 and 17-9
- Questions #13, 14
- Problems #30-34, 37, 39, 43-47 omit 45(e)
- Review the sections of Chapters 16 & 17 that we have looked at.
- Test on Chapters 16 & 17

Physics 42A - Class #6

Today's class: Sections 18-1 to 18-6

Here we leave Electrostatics to study Electrodynamics - where the focus is charges in motion.

Ben Franklin was among the first to study charge movement. He named the charges positive and negative, and noticed that a current resulted in conductors when there was a potential difference between the two ends. He incorrectly guessed that it was the positive charges that were flowing. Due to this error, and the difficulty of changing tradition, even today we will talk of current as if it were the flow of positive charges <u>even though we know that the fact is that the negative (electrons) do all the moving around</u> in solid conductors.

Potential energy is the least stable form of energy. Anything with potential energy will naturally, if left to itself, lose the potential energy, as it transfers it to other forms, such as kinetic energy. This means that positive charges will naturally "flow" from places of high potential, to places of low potential, gaining kinetic energy, as the potential energy is lost. The flow of positive charges is known as "conventional current" and is to be distinguished from "electron flow". The word "current" by itself is to be taken as "conventional current". Current always flows from high potential to low potential (electrons actually do the exact opposite, but this fact is less important).
To maintain a current, one must maintain a potential difference. Any device that does this can generically be called a "voltage source" and examples would be batteries, generators, solar cells, etc. (note that section 18-1 is more chemistry than physics - read it but don't "study" it).

Current is the flow of charges, and is given the symbol "I". The equation $I = q/t$ defines current, where q is the amount of charge that moves past a fixed point in a time t. The units would be "coulombs per second" which is given the special name "Ampere" or "Amp" (A) for short.

The amount of current that results depends not only on the potential difference, but also on how good of a conductor the voltage is applied to - "resistance". "Ohm's Law" defines resistance, and is commonly written as $I = V/R$. Note that units of R would be "volts per amp" and is given the special name "ohm" (Ω).

The resistance of an object depends on the physical characteristics of it: what it is made of, how long it is, and how "wide" it is. The resistance can be calculated by $R = \rho L/A$ where ρ (Greek letter "rho") is the resistivity of the substance (good conductors have a small value, while insulators have a large value), L is the length of the conductor, and A is the cross sectional area.

Certain materials exhibit the rather strange phenomenon of "superconductivity" which usually only takes place when a sufficiently low temperature has been achieved. A superconductor has a resistance of 0.

For Homework:
- Read Sections 18-1 to 18-6
- Questions #1-8
- Problems # 1-14, 19, 20

Physics 42A - Class #7

Today's class: Sections 18-6 to 18-8
Note that Sections 18-7 and 18-8 are not in the AP Curriculum, but we'll peak at them anyway - ignore the math stuff in 18-7

Please understand that the <u>reason</u> electricity is useful, is because it is a convenient way to transport <u>energy</u> from place to place. (Recall from last year that energy is the "ability to do work" and is measured in joules (J)). "Electrical energy" is nothing more than the kinetic energy of the moving electrons. These moving electrons have the ability to do work. In particular, one thing they can do is smash into the atoms of a resistor. This speeds up the vibrational movement of those atoms, which is detectable as an increase in temperature. Any electric appliance that provides heat does so on this principle, including incandescent light bulbs, where the resistor is a filament which radiates most of the energy as heat, with a small portion of the energy being light.
(It is literally "white hot").

The amount of energy carried by the charges is found by $U=q\Delta V$ (remember that ΔV was defined as the change of energy per charge). A very useful concept here is that of <u>power</u> which is the rate at which energy is delivered or consumed.

$$P=\frac{W}{t}=\frac{\Delta U}{t}=\frac{qV}{t}=\frac{q}{t}V=IV$$

In other words, electric power is the product of current (I) and voltage (V)

Note that in $P = IV$, we can use Ohm's Law ($I = V/R$) to substitute the I out to get $P = V^2/R$. We can also use Ohm's law ($V = IR$) to substitute the V out to get $P = I^2R$

$$P=IV=\frac{V^2}{R}=I^2R$$

From $P = \Delta U/t$ we can see that $\Delta U = Pt$. The SI unit of energy is of course the Joule. With electricity, however, there is another energy unit that (though not SI) is very common. In particular, it is commonly found in electric bills, as well as appliance ratings. Energy is the product of power and time. If you have power in watts, and time in seconds (both SI) then the product will have the SI unit of Joules. However, if power is in <u>kilowatts</u> and time is in <u>hours</u>, then the unit of energy is called a "kilowatt hour".

$$1\text{kW}\cdot\text{h} = 1000\text{W}\cdot 3600\text{s} = 3.60 \times 10^6 \text{ J}$$

As stated above, electricity carries energy. Too much energy is a bad thing, as "overloaded" circuits can overheat and catch fire. In particular, a "short circuit" is one which has a resistance far lower than it should, resulting in a dangerously high power level. ($P=V^2/R$) Two devices that help prevent such an overload from overheating are <u>fuses</u> and <u>circuit breakers</u>.
A fuse is simply a device that is built to burn out when a certain current is reached. This breaks the pathway for current, and all current is therefore stopped. A circuit breaker works the same way,

however, instead of burning out, the heat causes a bimetallic strip to bend, which again halts all current until it is reset.

There are two fundamentally different ways in which electricity exists. They are "Alternating Current" (AC) and "Direct Current" (DC). As the names imply, the current in DC is direct, or non-reversing (it always flows in the same direction). With AC, the current changes direction repeatedly. In the USA and Canada, household electricity is AC, the voltage is 120 V, and the frequency for the "alternation" is 60 Hz, meaning that it completes 60 cycles of the alternating current per second.

For Homework:
- Read Sections 18-6 to 18-8
- Questions #9-14
- Problems # 22-31, 33, 39
- Test, chapter 18

Physics 42A Class #8

Today's class: Sections 19-1, 19-2, and 19-5
Note that the notes here are brief, as the emphasis will be working with the material (which will be done in class, together)

Three very important aspects of a circuit are voltage, resistance and current. Voltage is provided to circuits by "voltage sources" or "EMF sources" such as batteries. Note that the term "EMF" stands for "electro-motive force" and basically means the same thing as voltage, except that it specifically refers to a source of voltage (thus, a battery is said to be a source of EMF). Resistance is found in many devices, however, we shall consider them to be generic "resistors". The current can be found by ohm's law, where the current running through any resistive device is the voltage across that device divided by the resistance of the device.

I will use the word "device" to mean any type of circuit component, including voltage sources, resistors, capacitors, etc. Devices can be placed in a circuit in one of two ways (or in combinations of these two ways). They are series and parallel. These are easily identified by the pathways that the current will follow.

In series, devices are placed one after the other, so that there is only one path for the current to flow: from one device to the next. Devices in series therefore all have the same current flowing through them. An analogy would be two garden hoses, if hooked up in series would form one longer hose, which would carry the same current of water through each of the two sections. For resistors placed in series, the voltage drops across each one, so that the total voltage drop is equal to the voltage applied to the circuit. For EMF sources in series, the total EMF is the sum (or difference) of the individual EMF's, depending on how they are placed - read section 19-5.

In parallel, devices are placed side by side, so that the current will branch - some going into each branch. For resistors placed in parallel, the current will add up to give the total current, and the voltage drop across each branch will be the same. A parallel hose arrangement would have two hoses carrying water side by side, as they split from a common source.

When more than one resistor is in a circuit, they act as a single equivalent resistance. The equivalent resistance for the entire circuit is known as the total resistance. For resistors in series, the equivalent resistance is the sum of the individual resistors. For resistors in parallel, the resistance is less than the smallest resistor, and can be found as follows:

$$R_{series} = R_1 + R_2 + R_3 + \ldots + R_n \qquad R_{\parallel} = \left(\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \ldots + \frac{1}{R_n}\right)^{-1}$$

For Homework:
- Read Sections 19-1, 2 and 5
- Questions #1 - 4
- Problems #1-23 (odd), 14, 22

Physics 42A - Class #9

Today's class: Sections 19-3, 19-4, 19-6 and 19-10

Our previous analysis of circuits is limited to circuits with one voltage source, or more than one voltage source in series. We have no way of working with more complicated circuits, especially ones where two or more voltage sources exist, not in series, but rather in different parts of the circuit completely. These types of circuits are called "multi-loop" circuits. To work with these types of circuits we need Kirchhoff's Rules, of which there are two.

Kirchhoff's First Rule is also known as the "junction rule", which is an application of the Law of Conservation of Charge. A junction is a place in which 3 or more wire are electrically connected. The junction rule simply states that the amount of current entering the junction must be the same as the amount of current leaving the junction - ie the current doesn't pile up in the junction, or come from nowhere.

Kirchhoff's Second Rule is also known as the "loop rule" which is an application of the Law of conservation of energy. The loop rule says that the algebraic sum of the changes in potential around any closed path of a circuit must be zero - ie, a charge going around any loop may lose and gain energy (voltage), but if it returns to the same part of the circuit, it will have no net change in energy (voltage).

We will work with these rules in class, note two things: it is very important that you know which laws that Kirchhoff's rules are based on (underlined above), and also that page 539 has a good summary of how to use them.

Capacitors, like resistors, can be placed in circuits in either series, or parallel, we can obtain the "equivalent capacitance" by the application of the following, which is exactly the opposite of the resistor rules:

$$C_{\parallel} = C_1 + C_2 + C_3 + \ldots + C_n \qquad \frac{1}{C_{series}} = \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3} + \ldots + \frac{1}{C_n}$$

A galvanometer is a device which measures very small currents. The galvanometer is the heart of ammeters, which can measure larger currents, voltmeters, which measure voltages, and ohmmeters, which measure resistance. The details of this section are not extremely important, so be sure to read it, but the important facts are that:

Ammeters measure current, and are placed in series with the part of the circuit where the current is to be measured; they should ideally have a very low resistance. Voltmeters measure voltage (potential difference) between two points of a circuit. They are placed in parallel with the devices between these two points, and should ideally have a very high resistance.

Ω Fd: - Read Sections 19-3, 19-4, 19-6 and 19-10
- Questions #5 - 15 - Problems #37, 39, 40 - 45

Physics 42A - Class #10

Today's class: Sections 20-1 to 20-4, 20-12

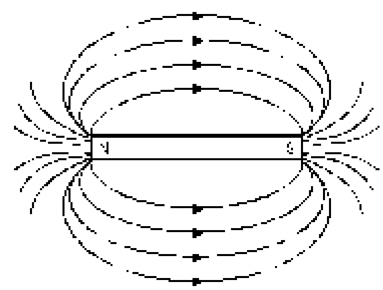
The history of magnetism begins with an ancient civilization in Asia Minor - in a region known as Magnesia, where "rocks" were found that would attract each other.

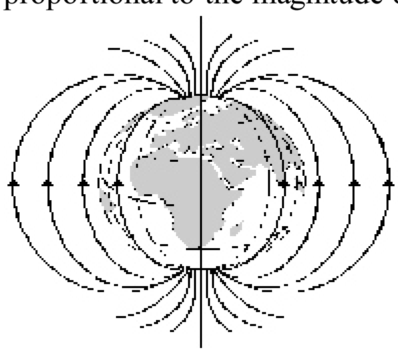
Magnets have two "poles" - *North* and *South*. The north pole will point north if the magnet is allowed to freely rotate about (as in a compass). Magnets are like electric charges in some (not all) ways . As with charge, like poles repel while unlike poles attract. However, isolated magnetic poles do not exist: Every magnet has a north pole and a south pole. Even if a magnet is cut in half, each half will have both poles.

All materials are affected by the presence of a nearby magnet. For most materials, however, these effects are far too small to be detected normally. *Ferromagnetic* materials are materials that readily show large attractive forces to nearby magnets. Examples of ferromagnetic materials are iron, cobalt, nickel and gadolinium. *Paramagnetic* materials are those that are very weakly attracted to either pole of a magnet (Aluminum, Magnesium, Tungsten, etc). While *Diamagnetic* materials are those that actually repel either end of a magnet (though extremely weakly). Examples of diamagnetic materials are bismuth, copper, gold and silver.

In the same manner that we speak of electric fields, we can describe the space surrounding a magnet as containing a *magnetic field*.

We draw the magnetic field lines so that: the direction of the magnetic field is tangential to the field lines and points in the direction of the force felt by a magnetic north pole. The number of lines per unit area is proportional to the magnitude of the magnetic field.





Compasses work because the earth itself is like a magnet. However, the Magnetic poles do not exactly coincide with the geographic poles - the geographic North Pole is about 1500 km away from the nearby magnetic pole. *"Magnetic declination"* is a measure of the angular difference between these two locations … which can vary from 0° to 25° in North America, depending on your location.

Since the magnetic field of the earth is not tangential to the earth's surface at most locations, there is also an "angle of dip" which refers to the angle the magnetic field makes with the horizontal.

The connection between electricity and magnetism is not obvious. For example, an electrically charged object does not show an attraction or repulsion to a magnet, regardless of the type of charge (+ or -) or type of magnetic pole (N or S). In 1820, Hans Christian Oersted discovered a connection.

Oersted found that a compass needle is influenced by an electric current in a wire. ***Electric Current*** (moving charge) produces a magnetic field. The magnetic field lines form concentric circles perpendicular to, and centered on the current. The direction of the field around these circles may be determined by the (first) right hand rule: If the thumb of the right hand is taken as the direction of current, the curled fingers point in the direction of the magnetic field about that current.
Magnetic Fields add in the same way that Electric Fields add: as vectors. This means that equal and opposite magnetic

fields cancel each other, and parallel fields reinforce each other. Fields that are at some angle to each other have a vectorial sum, which is the net magnetic field.

Of particular interest is the magnetic field produced by a current loop. Inside the loop, the magnetic field is perpendicular to the loop itself, in a direction given by the right hand rule. A *solenoid* is a long coil of wire consisting of many such loops of wire. Since Magnetic fields add as vectors, the magnetic field inside the solenoid is the vectorial sum of the individual loops which make up the solenoid. Solenoids can be used to produce large magnetic fields.

Recall that ferromagnetic materials are strongly attracted to a magnet. In addition to this, ferromagnetic materials may actually become magnetized (all permanent magnets are made of ferromagnetic materials). What is so special about ferromagnetic materials?

Ferromagnetic materials consist of large numbers of tiny regions known as *domains*. These domains are at most about 1 mm across. Each domain acts like a tiny magnet, with a north and a south pole. These magnetic effects within a domain are the result of the intrinsic motion of electrons.

Normally, the magnetic fields associated with these domains are arranged randomly, which results in a cancellation of the fields. However, in the presence of an external magnetic field, they tend to "line up" so as to produce an overall net magnetic field, in the same direction as the external field. These magnetic domains, once lined up, may stay lined up for long periods of time … a "permanent" magnet. This is best demonstrated by magnetizing an iron nail by stroking it with a magnet. The stroking tends to line up the domains in the iron nail, to such an extent that they remain lined up. The nail has become a permanent magnet itself.

The orderly arrangement of domains in a permanent magnet may be disrupted by vibrations, or by heat. In either case, (heating a magnet, or causing it to vibrate - such as by dropping it), this will cause the domains to lose their order, resulting in the magnet weakening.

Each domain can be thought of as a small current loop. Each small current loop creates a magnetic field, which we can describe with the right hand rule. Inside the material, these current loops cancel each other out, however, along the surface of the material, the currents will not cancel. A ferromagnetic material then, can behave rather like a solenoid, with current carried around the surface. This explains why the magnetic fields for the two are so similar:

Newton's 3rd Law states "for every action, there must be an equal and opposite reaction". Since a current exerts a force on a magnet, a magnet must exert a force back. The force on a current is not towards either of the poles, but rather at right angles to the magnetic field. Direction of the force is given by another right hand rule: point your right thumb in the direction of current, outstretched fingers in direction of magnetic field. Direction of the force is in the direction that your palm is "pushing".

Just as we symbolize the electric field as "E", we call the magnetic field "B". The force on a current carrying wire is in the direction given by the rule above, and has a magnitude proportional to the magnetic field "B," the current "I" in the wire, and the length "L" of the wire.

$$F = I\mathit{L}B \sin\theta = qvB\sin\theta$$

Where B is in the SI unit "Tesla" abbreviated "T" Note that charges moving may be represented as either IL or qv where IL is usually used for charges moving in a wire (current) and qv is used for individual charges moving at speed "v" In the latter case, it is important to realize that the charge will undergo <u>circular motion</u> as the magnetic force is always perpendicular to the velocity. (recall from last year that $F_C = mv^2/r$) and so $qvB\sin\theta = mv^2/r$

ΩFd: - Read Sections 20-1 to 20-4, 20-12
- Questions #1-7, 10-21

Physics 42A - Class #11

Today's class: Sections 20-5 to 20-9 (omit 20-8), 20-13

Ampere's Law is really quite complicated, and you need calculus to understand it fully. What it does, is relate current to the magnetic field it produces. It gives us a couple of handy formulas for determining the strength of the magnetic field (B) in different current arrangements. (note that they are not all on your "cheat shirt" - better memorize these). In these equations, μ_0 is a constant known as the "permeability of free space" and has the value of $4\pi \times 10^{-7}$ T·m/A

$B = \frac{\mu_0 I}{2\pi r}$ for a long straight wire, where r is the distance away from the centre of the wire.

$B = \frac{\mu_0 N I}{2r}$ for the magnetic field at the centre of a wire of N loops of radius r

$B = \frac{\mu_0 N I}{L}$ for the magnetic field inside a solenoid of length L, with N turns of wire.

Note that there are two things we have now looked at:

8. A current (I) in an external magnetic field (B) feels a force acting on it. ($F=ILB\sin\theta$)
9. Electric currents produce magnetic fields; for a straight wire $B = \mu_0 I/(2\pi r)$

Simply combining these two properties results in the fact that parallel wires carrying currents will exert forces on each other: the magnetic field <u>produced by one wire</u> being the <u>external field</u> in the other. It is important that you recognize that the field that one wire produces does not affect that same wire's current in any way; <u>only "external" fields (fields coming from somewhere else) will exert a force</u>.

The direction of the force can be seen by first determining the direction of the field produced by one current, and then seeing the direction of force the other current feels in that magnetic field. The result is that parallel wires carrying currents in the same direction (parallel) <u>attract</u> each other, while those carrying currents in opposite directions (antiparallel) <u>repel</u>.

The magnitude of the force per unit length can be found as: $\frac{F}{L} = \frac{\mu_0 I_1 I_2}{2\pi d}$ where d is the distance between the two wires carrying currents I_1 and I_2

Previously in physics, we have seen that the unit of current, the Ampere is equal to one Coulomb per second (1A=1C/s). Furthermore, the value of a Coulomb is equal to about 6.24×10^{18} elementary charges. The Ampère (and therefore the size of a Coulomb) is in fact <u>defined</u> as *that current flowing in each of two long parallel conductors 1m apart, which results in a force of exactly 2×10^{-7} N/m of length of each conductor.* The reason for this is to make the Ampère *operational* or "measurable" based on previously established units (the Newton and the metre).

The importance of the magnetic force on a current should not be underestimated. As a physics student, you should have a good understanding of how electricity "works". Most electric devices work in one of two ways: some produce heat when an electric current is provided. This includes electric heaters, light bulbs (which also emit light when hot), ovens, irons, toasters, etc. How these devices work is simple: the current loses power to thermal energy (heat) when it passes through the resistance ($P=I^2R$) of the device.

Other electric devices work by providing <u>motion</u> of some sort. The motion can be used in many ways. The motion could be a fan turning, a weed whacker whacking, a refrigerator compressor compressing and expanding coolants, a speaker cone vibrating (producing sound as a result), etc. Virtually all electric device which produces motion (even if it is internal) do so by virtue of the force a magnetic field exerts on a current. In general, an electric motor is the device which transforms electrical energy into motional (kinetic) energy. We will study them in class, and so you should have an understanding of how they work, even though the details are not in this handout.

ΩFd:
- Read Sections 20-5 to 20-9 (omit 20-8), 20-13
- Questions #22-24, 28, 29
- Problems 19-35 (odd, omit 29, 31, and do 35 (a) only), 28, 53, 55
- Test on Chapter 20

Physics 42A - Class #12

Today's class: Sections 21-1 to 21-4

Here it is, the grand finale to the electricity and magnetism section of the course: electromagnetic induction, which to be quite honest is fairly in-depth stuff. We (luckily, but also unfortunately) are only going to scratch the surface, and much of the chapter will be omitted - but be warned that there's still lots of heavy duty physics here.

In the previous chapter we saw that currents are the cause of magnetic fields: charges in motion produce magnetic fields. It turns out that the converse of this is also true: magnetic fields can produce currents, a fact first discovered by two great men of physics: Joseph Henry and Michael Faraday. Here's how it works (note my approach is quite different than that of the text):

Consider a conducting wire of length L, moving at right angles to a magnetic field at speed v. Will there be a force on the wire? No, because the wire is (assumed to be) neutral - only <u>charges</u> moving in a magnetic field have forces on them. But the wire itself is made of charges - in fact it has electrons (negative charges) that are able to move (since it is a conductor). If the wire is moving in the magnetic field then the electrons themselves are moving in the magnetic field - also at a speed v, with the wire.

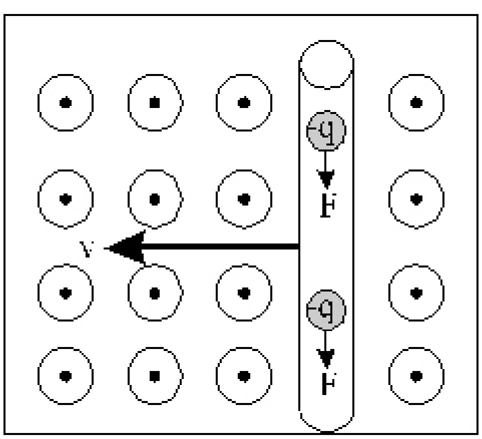
- The electrons in the conducting wire will have a force on them: $F=qvB$ (ignore $\sin\theta$, as $\theta = 90°$).

- Since the force acts over a distance (the length of the wire), <u>work</u> is being done to them: $W=Fd=(qvB)L$

- But work is a transfer of energy, so each electron receives energy $U = W = qvBL$

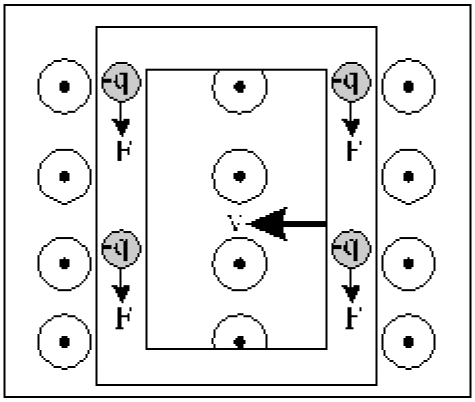
- This means that the energy per electron (U/q) is: $U/q = (qvBL)/q$

- But energy per charge (U/q) is <u>voltage</u>! (also known as EMF, ε) so the potential difference between the ends of a conductor moving at right angles to a magnetic field is: $\varepsilon = Blv$



This is extremely important: A voltage can be created across a conductor simply by moving it (at right angles) to a magnetic field. The direction of the voltage (which end will be positive / negative) can be found by thinking of which way charges will move with the right hand rule: as shown in the diagram here, the conducting wire segment is moving to the left, so charges (including the shown electrons "-q" are moving to the left, in a magnetic field coming out of the page. The electrons will have a force on them down the wire, and so there will be a potential difference (ε), with the top of the wire being positive, and the bottom being negative. Note that you can also think of the positive charges being pushed up the wire, which gives the same results, but realize that the protons are not free to move about within the conductor.

Another equation that pops out of this line of thinking is that within the conductor, in the situation shown, there must be an electric field, since the charges have forces on them ($E=F/q$) so $E=(qvB)/q = vB$ again, where v and B are perpendicular to each other, and perpendicular to the wire (as shown).



The previous discussion involved only a wire segment, which would develop an EMF across it. For there to be a current, the wire must be part of a closed circuit. The simplest way to close the circuit is to use a wire loop instead of a straight wire (the loop shown here is rectangular for simplicity). Unfortunately, dragging the loop across the field will not produce a current, as the two opposite sides of the loop will produce voltages that cancel, as shown in the diagram here (electrons are being pushed against each other). In order for there to be a current throughout the loop, one side needs to be outside of the region where the magnetic field is. But there is a better way of thinking about it ...

We define magnetic flux to be the amount of magnetic field "caught" in a wire loop. In the above diagram, you can see that the loop "caught" about four of the circles representing magnetic field. This number represents (very loosely) the magnetic flux. Clearly the magnetic flux could have been increased in one of two ways: make the loop bigger, or make the magnetic field stronger (which would place the circles representing it, closer together). The Greek letter Phi (Φ) represents flux, and so Φ_B represents magnetic flux (stick around for luminous flux when we look at light). Magnetic flux is proportional to the magnetic field strength (B) and also the size (area A) of the loop. In fact $\Phi_B = BA\cos\theta$ where theta is the angle between the magnetic field and the normal to the loop (which goes straight through the loop - the diagram above would have $\theta = 0°$ as the normal is parallel to the magnetic field.

Magnetic flux has the units of magnetic field (Tesla) times area (m^2) or Tm^2 which is given the name "Weber" (Wb)

The highlight of the chapter is "Faraday's Law of Induction" which states that an EMF will be induced (produced) whenever the magnetic flux through a loop changes. In fact the rate of change of magnetic flux will be the induced EMF: $\varepsilon = -N\frac{\Delta\Phi_B}{\Delta t}$ Where N is the number of turns of wire in the loop.

The negative sign in the above law indicates the direction of the EMF, and is interesting in its own right ... it even has its own name: "Lenz's Law" which in my opinion is extremely fascinating (we're talking neat-o demonstrations). Lenz's Law states that the induced EMF will be in the direction such that the induced current will set up its own magnetic field to minimize the actual change in the magnetic flux. This is best seen by demonstration: make sure you understand this when we look at in class; if not, come see me, and we'll play with it so you can see it yourself.

My advice is to ignore the negative sign in Faraday's Law when simply doing the calculation, then apply Lenz's Law separately to determine the direction of the induced voltage and/or current.

Homework:
- Read Sections 21-1 to 21-4
- Questions #1-4
- Problems 1-7 (all), 9-17 (odd)
- Test on Chapter 20

Physics 42A - Class #13

Today's class: Sections 21-5 to 21-8 (omit 21-9 to 21-15)

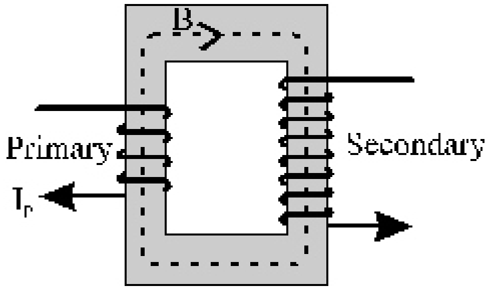
In the previous chapter we saw how a (DC) motor works. Here we see how a Generator works. A generator involves a loop (actually many loops) of wire inside a magnetic field. Something turns the loop around and around, which results in a magnetic flux which rises and falls (a changing flux) which gives rise to an induced voltage, and if there is a closed circuit, an induced current as well. The something that turns the loop has mechanical energy: a rushing river for example, in the case of a hydroelectric generator. Or it may be steam that turns the loop, if forced through a turbine. The steam may have been produced by burning something (coal) or in the nuclear fission of uranium, which produced tremendous amounts of heat. Or of course, it could have come from the food energy a person ate, when someone turns a small generator.

So a turning loop inside a magnetic field induces a voltage. But wait a minute - a motor is a loop (wrapped around an armature) turning in a magnetic field! That's right, a motor and a generator are in principle, the same thing! The main difference is their intended use: A motor will be supplied with electrical energy, which it transforms into mechanical energy. A generator is supplied with mechanical energy, and transforms it into electrical energy. This is a very important concept, and results in the following fact: A motor, while turning, produces its own voltage! Don't jump to the conclusion that a motor, once turning will be a perpetual motion device (it's own voltage keeps it going). That darn Lenz's Law spoils all the fun: the induced voltage results in an induced current which acts to oppose the turning (not keep it going). This is called "Back EMF" - a motor turning produces a voltage opposite to the applied voltage, effectively reducing the overall voltage.

This means that a voltage, when applied to a motor does the following: At first, the motor is not turning, and so an applied voltage (EMF) results in current through the motor, which should obey ohm's law, where $I=V/R$. This current in the motor results in a torque which gets it turning, but as it turns it produces a back EMF, effectively lowering the current through the motor ($I=(V-EMF_{back})/R$) eventually the motor will be going full speed, at which time the current going through the motor is comparatively small. This explains why the lights in your house often dim briefly when your fridge kicks in. When the motor is first turned on it draws a lot of current. This large current goes through the wires in the whole circuit. Even though the wires have small resistance, the voltage drop $V=IR$ is not negligible when the current is large, and so the wires themselves use up a portion of the 120 V, leaving less than that for other parts of the circuit, including the lights, which dim as a result. It doesn't last long because the motor turns, and the current drops as described above.

The equivalent situation for a generator is called "Counter Torque". The armature is turning because something is making it turn, resulting in an induced EMF. If there's no closed circuit, then that's all that happens, but when there is a circuit, an induced current is allowed to flow: a current in this loop, which is in a magnetic field has a torque on it (behaving like a motor) which again by Lenz's Law acts to oppose the motion. Meaning that it is hard to turn a generator when there is a current running in it.

A few chapters ago, we saw that it was best to transmit electrical energy at high voltage, so that the current could be low in the power lines ($P=I^2R$) reducing the power lost in them. However, a high voltage would be a dangerous thing to have in the house. A transformer is the device that saves the day: it can change any voltage into any other voltage, provided that it is AC and not DC.



Here's how it works: currents produce magnetic fields. Changing magnetic fields induce voltages (and therefore currents). So get yourself a current that changes (alternating current - AC). This should produce a magnetic field that also changes. Harness this changing magnetic field to induce a new voltage. In practice this is done by taking the wire carrying the input voltage and current, and wrapping it around an iron (hmm .. that's ferromagnetic ... coincidence?) core. Call this set of windings the "primary" or "input". Somewhere else on the iron core, put another set of windings for another wire. Call this the "secondary" or "output". The cool thing is that the primary and secondary are not electrically connected at all! (the wires themselves are insulated). However, the primary carries alternating current with a set amount of energy (voltage). This produces a magnetic field inside the core. Since it is iron, the magnetic field is transmitted throughout the core, which happens to go right through the secondary. The secondary core now has a magnetic flux through it ... which is changing! Faraday's Law steps in and says that there will be an induced voltage, depending on how many turns of wire there are there! ($\varepsilon = -N\frac{\Delta\Phi_B}{\Delta t}$) So choose how many turns of wire you want (N) and you can choose your secondary voltage. The number of turns in the primary also plays a role, as the more turns there are there, produces a stronger magnetic field (in fact, it kind of resembles a solenoid, except for the iron core, in fact it is behaving exactly like an electromagnet). It mathematically becomes a ratio:

$\frac{V_S}{V_P} = \frac{N_S}{N_P}$ where N and V are number of turns and voltage, for the primary and secondary.

One thing might bother you: Can we take 1 volt and turn it into a billion volts? Yes, we can. Doesn't this violate the law of conservation of energy? No it doesn't. Why not? Because it's THE LAW

Actually, energy is conserved: the voltage might increase, but this will result in a corresponding decrease in current. In fact, for a "perfect" transformer (which wouldn't heat up at all) the power in the primary and secondary would be the same. This means that $P_P = P_S$ so $I_P V_P = I_S V_S$.

Note that Induction has a whole lot of applications: make sure you read 21-8

One other thing I didn't include here is a description of "eddy currents" be sure to read up on that, and I will also be talking about them in class.

Homework: - Read Sections 21-5 to 21-8
- Questions #8, 10 - 15, 17
- Problems 19, 21, 23, 28-36, 38

Physics 42A- Class #14

Today's class: Sections 11-1, 11-2, 11-8, 11-9
(Note that these and 11-11 to 11-13 are the only sections of Ch 11 we will be looking at here)

This chapter starts off with some familiar concepts - as they were discussed in grade 11. Simple Harmonic Motion (SHM) is motion which is regular and repeating. An example of SHM would be the regular vibration of an object about a rest or "equilibrium" position. Fundamental measurements of this type of motion include period (T) - the time it takes for one full cycle of the repeating motion to occur; frequence (f) - the reciprocal of period, being the number of cycles completed per unit time; and amplitude (A) - the maximum displacement from the equilibrium position.

Springs are also discussed here. Important equations that you may or may not remember from grade 11 are $F = -kx$ and $U = \frac{1}{2}kx^2$. I won't go into the details here, and I am not stressing sections 11-1 and 11-2 at this point, as in my opinion they are more along the lines of grade 11 material, which we will review later. Basically, skim over these two sections to reacquaint yourself with what I have already said in the first paragraph above.

Our focus begins in section 11-8: Wave Motion. Up until now, energy has always moved from one location to another by being carried by matter. This may be done in a variety of ways: the object may be moving fast (kinetic energy) or may be hot (thermal energy), but in general, a physical object has always moved from one place to another, carrying energy with it. We now look at how energy may be transported from one location to another, but not by being carried by a physical object. A wave is "a travelling disturbance that carries energy" and while it may exist in matter, it is important to realize that energy is transported, while matter is not.

Waves fall into two general categories: Mechanical and electromagnetic (there is also a third category we will see much later: the "matter wave"). We will see the latter in a future chapter, and so focus on Mechanical waves here; from this point on in this section, when the word "wave" is used, I will be meaning "mechanical wave". Mechanical waves are those waves that need a medium (matter) to travel through. Every wave has a source, which disturbs the medium. Each disturbance is called a pulse, and a wave is a continuous series of pulses being produced by a source which is undergoing SHM. An important type of wave is one in which the source's SHM is sinusoidal, in which case the wave itself is sinusoidal. For a sinusoidal wave, we call the maximum displacement *above* the equilibrium position a crest and the maximum displacement *below* the equilibrium position a trough.

When matter moves, we focus on such properties as *momentum, kinetic energy, velocity,* etc. For waves, we focus on such properties as wavelength (λ - Greek letter lamda), which is the distance between any two successive identical points on a wave, period (T), and frequency (f) (already discussed above). We can also discuss the wave velocity (v) as being the velocity of the pulses through the medium. Notice that $v = d/t$, and a wave travels a distance equal to the wavelength (λ) in a time equal to the period (T). This makes the velocity of a wave $v = d/t = \lambda/T$ and since f and T are reciprocals: $v = \lambda/T = \lambda f$.

One important thing to realize about the above equation for wave velocity: it is not proportional to either wavelength (λ) or frequency (f). In fact, the wave velocity is constant FOR A GIVEN MEDIUM. One common medium which carries waves is a string. As stated above, the wave velocity is constant through a string, depending only on the string itself, namely the tension in the string, and the mass per length of the string. In fact, for a string, the wave velocity is $v = \sqrt{\frac{F_T}{m/L}}$ Notice that the source sets the medium into motion, as the pulse propagates through it, though the medium immediately returns to its original position after the pulse has passed (hence energy is moving from place to place though matter itself is not, as discussed above). Waves are categorized by how the medium moves in comparison to the direction of the wave motion itself. If the disturbance moves the medium at right angles to the disturbance, it is known as a transverse wave. A transverse wave can be created in a spring or string, by shaking one end from side to side. If on the other hand, the disturbance moves through the medium parallel to it, it is known as a longitudinal wave. A longitudinal wave can be created in a spring by quickly compressing and then expanding the spring. For a longitudinal wave, the words "crest" and "trough" are replaced by "compression" and "rarefaction" for areas in which the medium is bunched up, and pulled apart respectively. A third type of wave is a surface wave, in which the motion of the medium is a combination of the two basic types.

If matter is in motion, and it hits an obstacle, it may "bounce" back. The same is true with waves, however the "obstacle" for a wave is the point separating two different media. The bouncing of a wave is called reflection. There are two types of reflection, depending on what kind of medium the wave is attempting to enter. If the new medium supports the wave at a higher wave velocity, then part of the wave will reflect back, with the reflected pulses on the same side of the equilibrium position. If on the other hand the new medium supports the wave at a lower wave velocity, part of the wave will reflect back, but will become "inverted" meaning that the pulses will appear on the opposite side of the equilibrium position to where they were originally.

Notice in the above that it is "part of the wave" that reflects. The other part will continue into the second medium, a process called "transmission". The extent to which a wave is reflected depends on how different the new medium is from the first. Two special situations are when the second type of medium is different to the extreme: when it is "fixed" (unable to support the wave at all because it is unable to move - in which case essentially all of the wave is reflected, and inverted) and when it is "free" (unable to support the wave because the second medium simply doesn't exist - the first medium simply ends, in which case the wave is essentially all reflected, but remains "upright").

Waves aren't always constrained to a line. Two and three dimensional waves are actually more common. With these waves, we focus on the wave fronts which are the entire width of a wave crest. A ray is an arrow representing the direction of motion of a wave front in one particular line. When a wave front encounters a barrier, it undergoes reflection. The Law of Reflection states that the "incident wave" (before reflection) and "reflected wave" strike the barrier at equal angles: the angle of incidence (θ_i) is equal to the angle of reflection (θ_r). Where both of these are measured with respect to normal to the barrier.

Two pieces of matter can't be in the same place at the same time. If you try to bring two pieces of matter into the same place at the same time, they will hit each other, and refuse to be put into that situation. Not so with waves: waves never "collide" - two waves occupying the same place at the same time is commonplace. Waves in this situation obey what is known as the "Principle of Superposition" which is simply that the overlapping waves yield a resultant displacement that is the algebraic sum of the separate displacements (treating "above" the rest position as positive and "below" as negative). This phenomenon is known as interference, which may be constructive (when two crests and /or two troughs overlap to produce an even bigger crest / trough) or destructive (when crest meets trough, resulting in total or partial cancellation of the wave).

Homework: - Read Sections 11-1, 11-2, 11-8, 11-9
- Questions 13-16, 18-21 - Problems #34-36, 39

Physics 42A - Class #15

Today's class: Sections 11-11 to 11-13

We left off talking about interference, which is the phenomenon of two or more waves being in the same place at the same time. If the waves are "in phase" (crests are aligned with crests) then constructive interference occurs, leading to larger crests. On the other hand, if the waves are "out of phase" (crests aligned with troughs) then destructive interference will occur (the waves "cancel"). A third possibility is a combination of the above, when only partial destructive interference occurs.

We have also discussed wave reflection. Of particular interest now is the reflection off of a fixed end. This is when the wave hits an end that cannot move. We have said that the reflected wave will be inverted. This is of interest because the reflected wave and the original wave will interfere with one another! At times, the reflected crests will meet original crests - constructive interference, while at other times, destructive interference occurs. This leads to regions which do not move (nodes) and regions which will oscillate back and forth with a large amplitude (antinodes). Such an arrangement is called a "standing wave". Don't let the name (or how it looks) fool you - a standing wave is actually two waves moving in opposite directions. I will discuss this further in class, and show demos - you need to have a good understanding of this phenomenon.

The critical thing to remember about standing waves is that the length of the medium contains an integer number (n) of half wavelengths ie $L = n\,(\lambda/2)$ and since $v = \lambda f$, we also can say that a standing wave has a frequency $f = n\,v/(2L)$. A wave frequency which leads to this result is referred to as a "natural frequency" for that medium. The lowest such frequency (when $n = 1$) is called the "fundamental frequency". All others are called "overtones" with the first overtone being when $n = 2$. The importance of standing waves should be clear: whenever an object is set into vibration, it does so with a combination of natural frequencies. This will lead us into our study of sound next chapter.

Two and three dimensional waves exhibit two more characteristics: <u>refraction</u> and <u>diffraction</u>. Refraction occurs when a two dimensional wave enters a "new" medium which supports the wave at a different speed (note that "new" can include the same medium in a different physical condition, a water wave for example, may enter water of a different depth). When such a wave enters a new medium, the direction of the transmitted wave is different than the direction of motion of the incident wave. The Law of Refraction mathematically describes this as:

$$\frac{\sin\theta_1}{\sin\theta_2} = \frac{v_2}{v_1}$$

Where the angles are measured between the wave velocities and the normal to the boundary separating the two media. It is worth noting that when a wave enters a faster medium, it bends away from the normal, and when entering a slower medium, it bends toward the normal.

Diffraction is the bending of a wave around an obstacle. This may seem counterintuitive, but what we are essentially saying is that waves cannot be cleanly blocked - the "shadow" of a wave can seal itself up reforming the wave. Again, I will discuss the reason behind this further in class. It is worth noting that extent to which a wave diffracts depends on the wavelength (longer wavelengths diffract more) and on the width of the obstacle (smaller obstacles lead to more diffraction).

Homework: - Read Sections 11-11 to 11-13
- Questions 22 - 24
- Problems #50-61, 63-64, 67

Physics 42A Class #16

Today's class: Sections 12-1, 12-5, 12-7 - 12-8

As I'm sure you are already aware, sound itself is a wave. Specifically, it is a longitudinal wave that travels through a medium (usually air) that results from a vibrating source that compresses and rarefies the air (which correspond to crests and troughs). The speed of sound, as with all waves, depends only on the state of the medium. For sound in air, the speed can be found approximately by v = 331 + 0.60T where T is the temperature in Celsius. Notice then that at room temperature (20 °C, and at 1 atm), sound travels at 343 m/s. Notice also that sound travels much faster in liquids, and faster yet in solids.

What we sense as "pitch" (treble / bass) is actually the frequency of the sound wave. What we sense as "loudness" is actually an indication of the amount of energy being carried by the sound wave. Humans, on the average, are able to hear sounds varying from 20 Hz (low frequency = low pitch) to about 20 kHz (high frequency = high pitch). Sound intensity (which relates to loudness) is measured in decibels (dB) and is rather complicated, due to the complicated nature of how we hear. Nonetheless, humans can hear sounds ranging from 0 dB (by definition) up to 120 dB. Above 120 dB can be heard but is actually painful, and can cause hearing problems.

Sound is a pressure wave. Its source is anything that moves back and forth (vibrating). It is the back and forth movement which pressurizes / rarefies the air. Of particular interest are strings that vibrate back and forth, and therefore set the air around it into vibration, and air that is in a tube (or column), which can itself be made to vibrate by a variety of techniques. The reason for these two choices should be obvious: They describe the large majority of musical instruments. When an instrument's string (guitar, piano, etc) or column of air (trumpet, flute, etc) is set into vibration, only resonant frequencies (Harmonic frequencies) will persist.

In a stringed instrument, for example, each end of the string must be a node, as they are fixed in place. This leads to what we discussed last class: that $L = n(\frac{1}{2}\lambda)$. With air columns it is a little more complicated, as the ends of the tube may be open (to the atmosphere) or closed (sealed). Furthermore, we investigate the tubes in two different ways: in terms of displacement (which corresponds directly to the way we investigated strings) and in terms of pressure.

	at an open end	at a closed end
Air displacement	antinode	node
Pressure	node	antinode

Open at both ends: $L = n(\frac{1}{2}\lambda)$; $f_n = n(v/2L)$	Closed at one end: $L = m(\frac{1}{4}\lambda)$; $f_m = m(v/4L)$

n = integers (1, 2, 3 ...); m = odd integers (1, 3, 5 ...)

<u>Beats</u>: when two sources have a different frequency, the resulting wave will vary in intensity (loudness). The sound will "come and go" with a frequency that is the difference of the two source frequencies.
<u>Doppler Effect</u>: When either the source, listener, or both move with respect to the medium, the frequency will appear to be different than that of the source $f' = f\frac{v \pm v_o}{v \mp v_s}$ where you use the upper signs for the <u>o</u>bserver / <u>l</u>istener approaches the other

Homework: - Read chapter 12, - Questions 1-6, 15-16, Problems #1-8, 27 - 57 odd, omit 6, 41, 51

Physics 42A - Class #17

Today's class: Sections 22-5, 23-5, 24-1 and 24-2

Here we enter our study of light. A fundamental question to address is "What is light?" This simple question has no simple answer. There are two models which attempt to describe what light is. They are the "corpuscular model" and the "wave model". The corpuscular model basically says that light consists of small particles which go by the name of photons - we will look more at this later. The wave model says that light is a wave unlike any other. It is unlike other waves in that it does not need any medium to travel through: it can propagate through a vacuum, and does so at the incredible speed of $c = 3 \times 10^8$ m/s.

What kind of wave needs no medium? James Clerk Maxwell, in studying electric and magnetic fields, discovered that an oscillating electric charge produces both electric and magnetic fields that will propagate at "light speed" and therefore concluded that these "electromagnetic waves" (EM for short) in fact were light.

The above is almost true. "Light" is in fact EM waves within a certain, narrow range of frequencies. In other words, EM waves can come in any frequency. These different frequencies are given different names, but only "light" is visible to us. The <u>Electromagnetic Spectrum</u> is the name given to the entire series of different type of EM waves. You should know the order of the types of waves: from lowest frequency to highest frequency (and therefore largest wavelength to smallest) they are: Radio, Microwaves, Infrared, Visible Light, Ultraviolet, X-rays and Gamma rays. Visible light itself comes in a range of frequencies. Different frequencies of light are perceived by us as "colour" (just as different frequencies of sound are perceived as pitch). The light spectrum (again in order of increasing frequency / decreasing wavelength) are Red, Orange, Yellow, Green, Blue and Violet (note that physicists generally have no use for "Indigo" ... whoever heard of indigo anyway?) Note that higher frequency EM waves are also associated with higher energies (look what gamma rays did to Bruce Banner!)

One of the problems that people initially had with the wave model of light is that it predicts that light will do all of the "wave things" we studies in the last unit, such as diffraction, refraction, reflection, etc. Light can certainly reflect, but this could also be thought of as particles bouncing. Even refraction can be explained (not very well, but ...) if light is thought to be a particle (but there are problems with what it predicts), but diffraction is unique to waves: If light diffracts, then it for sure a wave! (Note that another problem with the wave model was that Newton didn't go for it ... and Newton's opinion carried a lot of weight).

<u>Huygens' Principle</u> helps us understand many wave phenomena: It simply says that a wave propagates as each point in the wave acts as a source of new waves: the overall wave that continues is the sum of all of the small "wavelets".

The Law of Refraction (also known as Snell's Law) describes, mathematically, the angle of incidence and the angle of refraction (both measure to the normal) in terms of the speed of the EM waves in the two media. When light (or other EM wave) passes from one medium to another, then $n_1 \sin\theta_1 = n_2 \sin\theta_2$ where "n" is the "index of refraction" defined as $n = c/v$ (v being the speed of EM wave in that medium).

Homework: - Read Sections 22-5, 23-5, 24-1 and 24-2
- Ch 22 Questions 1-8, Problems 10-15; Ch 23 Problems 31-39 odd;
- Read ahead to the rest of chapter 23

Physics 42A - Class #18

Today's class: Sections 24-3 to 24-11

The way we know that light is for sure (ha ha ha) a wave is that it exhibits interference. There is no way to explain interference when thinking of light being made of particles. The interference of light is typically demonstrated by shining it through one or two "slits".

The double slit experiment is attributed to Thomas Young. In it, light falls on two small slits - which, by Huygans' principle, then act as individual, new, in phase sources - just like the speakers we discussed in the sound chapter. These two sources will result in constructive interference when the light coming from them arrive "in phase", which will be when they travel the same distance, or, if one of them goes a full integer number of extra wavelengths in distance. Destructive interference will result when the two waves arrive "out of phase" which results from one of them going further than the other by an amount of a half wavelength, one and a half wavelengths and so on. A little geometry reveals the following:

$d \sin\theta = n\lambda$ (Constructive Interference ... Bright Spot) $d \sin\theta = (n + \frac{1}{2})\lambda$ (Destructive ... Dark)

d = distance between two slits; θ = angle away from central region; $n = 0, 1, 2, \ldots$

Notice the dependency on wavelength in the above. When it was observed that white light going through two slits forms bands of color, it gave strong evidence to believe that different colors are different in wavelength (by the way, "brightness" or "intensity" corresponds to the amplitude of the wave). The separation of white light into its component colors is known as <u>dispersion</u>, and the color pattern resulting is known as the <u>visible spectrum</u>. Dispersion can also result from refraction, and in fact this is essentially what creates a rainbow ... see page 687 for the details.

A single slit can also give an unusual pattern, but the reason is different than that for double slits. <u>diffraction</u> is the cause, as light (waves) bend around obstacles, or thing openings (slits). Light shining through a single slit will diffract, resulting in a maximum of intensity straight ahead (which would be expected), and a series of dark regions (minima) and bright regions (maxima) around that central maximum. For single slits:

$D\sin\theta = n\lambda$ (minima) ($n = 1, 2, 3 \ldots$) maxima occurring in between minima and in central region

(Be sure to read 24-6 and 24-7 about "diffraction gratings" I will discuss them further in class). Yet another way to get an interference from light is to shine it on a "thin film" such as a soap bubble, or a layer of oil on water. The idea is simple: there are two boundaries (air/soap and soap/air OR air/oil and oil/water). This will cause two separate reflections. Depending on how much further one path is, the two resulting waves may be in or out of phase. The "difference in path length" will be twice the thickness of the film.

Homework: - Read Sections 24-3 to 24-11
- Questions 1-10, 12, 14, 15, 21, 22, 25-29
- Problems 3-25 odd, 39, 41 - Read ahead (back?) into chapter 23

Physics 42A - Class #19

Today's class: Sections 23-1 to 23-10

Up until now we have been studying the physical properties of light itself, namely the wave properties of light. Here we begin a study of the behaviour of light, as it interacts with such things as lenses and mirrors. Much of this chapter will make a lot of sense if you understand two basic things: ① you will only "see" the light that enters your eye; ② your brain assumes that the light entering your eye came from its source along a straight line. The fact that light usually does travel in a straight line leads to what we call the "ray model of light" - in which the straight path of a narrow beam of light is represented by a ray. We "see" things when many such rays of light emanating from things around us enter our eyes.

As mentioned before, light travels at a fixed speed "c" in a vacuum, $c = 3 \times 10^8$ m/s, but slower than this in media such as air, glass, etc. (Be sure to read about the Michelson - Morley experiment to determine the speed of light). The ratio of the speed of light in a vacuum to the speed of light in a given medium is termed "the index of refraction", "n", where $n = c/v$. Notice that $n \geq 1$, as $c \geq v$.

When light encounters something in its path, one of (or a combination of) three things typically occur: The light may be absorbed (in which case the obstacle warms up slightly), it may enter the new medium (transmission) or may bounce off of the boundary between the two media (reflection). Absorption is not terribly interesting in the sense of geometric optics, but reflection and transmission are.

Reflection is described by the "Law of Reflection" in which the angle of incidence θ_i (incidence means "incoming") is equal to the angle of reflection θ_r, that is $\theta_i = \theta_r$. One thing about angles in geometric optics: they are always measured with respect to the normal, which is the perpendicular line from the boundary of the two media. There are two types of reflection, depending on the nature of the surface light is reflecting off of: if the surface is very smooth, then "specular" reflection occurs. This is when a bundle of parallel rays reflect off of the smooth surface, and remain a bundle of parallel rays after the reflection has taken place. If the surface is not very smooth, then "diffuse" reflection occurs. Here the light rays reflect, each one of them obeying the law of reflection, but because of irregularities in the surface, parallel incident rays are reflected all over the place.

A key area in geometric optics is the formation of <u>images</u>. An image is the appearance of an object (often in a location where it is not). A common example of an image is your own "reflection" in an ordinary mirror. You see an image of yourself (an object) in a location where you are not (behind the mirror). Images result when a bundle of rays appear to originate from one spot. Understand that this is how light actually originates from objects. For example, if a cat was on the floor, then light is reflecting off of the cat, and spreading out in all directions (diffuse reflection). The fact that light is <u>originating from the cat</u> allows us to "see" it regardless of where we are (provided light can get to our eyes). If the light coming off the cat were to strike a mirror, and reflect from it (specular reflection) then those rays of light will seem to be originating from a point behind the mirror, and this is where the cat's image will be. There are two types of images: <u>real images</u> and <u>virtual images</u>. A real image is one in which there really is light emanating from the location of the image ... if you look directly at the cat, you see a real image, as light really is coming from that spot. A virtual image is one in which the light is in fact not originating from that spot, but seems to be, since our brains automatically assume that light was going in a straight path the whole time.

Notice that for flat or "plane" mirrors, the image will appear upright, in a location the same distance behind the mirror, as the actual object is in front of the mirror. We would say the "image distance" (d_i) is equal to the "object distance" (d_o). It is a virtual image, as the light is in fact not coming from back there. Another important type of mirror is a spherical mirror ... one which is part of the surface of a sphere. We call the outer surface of such a mirror convex and the inner surface concave. Spherical mirrors are not the ideal type of mirror to use for what we would like, but they are easier to make than parabolic mirrors, so our study of spherical mirrors is actually only an approximation, which works well when the size of the mirror is small compared to the radius of the sphere from which it is a part. Here's what you need to know about the rays striking the concave side of a spherical mirror:

- Rays that come in parallel to the principal axis (the straight line perpendicular to the centre of the mirror) will reflect and all go through one point, which is called the focal point (F). The distance to the focal point is the focal length (f), and is half the radius ($f = r/2$).
- Rays that go through the focal point reflect off the mirror parallel to the principal axis.
- Rays that go through the centre strike the mirror at right angles, and so are reflected right back.

"Ray tracing" is the process of locating the image of an object by using the three rules above. See page 652 for the details. Some terms to keep track of: h = height, d = distance (to mirror), f = focal length, r = radius of spherical mirror, m = magnification. Relationships: $1/d_o + 1/d_i = 1/f$, $m = h_i / h_o = -d_i / d_o$
Notice that heights are positive if above the principal axis, and distances (including focal lengths) are positive if on the reflecting side of the mirror. If more than one mirror (or lens) is present, then the image of one becomes the object for the next.

In the case where light enters another medium, the transmitted ray will generally not be parallel to the incident ray, but rather refraction occurs. The Law of Refraction (Snell's Law) has been mentioned already, and is $n_1\sin\theta_1 = n_2\sin\theta_2$. It is nice to know that when light travels into an object which is more "optically dense" (n is larger) then it is bent towards the normal, whereas light entering a less optically dense medium (n is smaller) bends away from the normal. The extreme case of the latter situation gives an interesting result: mathematically it is possible to have a refracted angle giving impossible results (a sine value larger than one). Physically, the refracted angle is restricted to a maximum of 90° and this occurs when the incident angle is at the "critical angle" θ_c. Setting $\theta_r = 90°$ in Snell's Law reveals that $\sin\theta_c = n_2/n_1$. For incident angles larger than this, refraction does not occur, but rather the incident light is totally reflected back into the originating medium. Again, remember that this will only occur when light is trying to enter an optically less dense medium (such as from glass to air).

Lenses are similar to mirrors in that they can form images. Our approach to lenses is identical to that for mirrors, except for the obvious fact that light passes through them. Even the same equations apply as for mirrors. A noticeable difference is in how we categorize lenses: converging and diverging. A typical converging lens is a double convex, and a typical diverging lens is a double concave.

A general point here: This chapter should be read very carefully, and you should go over the examples in detail. I believe this to be an easy chapter, in that you can pick it up easily if you allow yourself the time to do so.

Homework: - Read Sections 23-1 to 23-10
- Questions 1-4, 12-20,
- Problems 1-10 all, 15 - 27 odd, 41-44, 49 (a only), 53, 57-61 odd

Physics 42A - Class #20

Today's class: Chapter 27 (all, but don't worry too much about 27-2, 7, and 9)

Yes ladies and gentlemen, we're studying quantum physics (don't panic, it's not so bad)

As we have said before, light is for sure a wave. It does things that can only be explained in terms of waves, such as interference. All is good, boy are we smart for figuring this out! Then along comes this guy named Albert who wrecks all our fun (and makes us wonder if we're really that smart). In one year (1905) Einstein produced three papers. One was on Brownian Motion (which gave strong evidence for the atomic theory), another was on Relativity (for which he is most well known), and another was on the photoelectric effect (for which he won the Nobel Prize in physics in 1921 ... NOT for relativity!)

Basically Einstein revealed that light is of course a particle, which is now know as a **photon** (as in "photon torpedo"). The concept was first conceived of by the great Max Planck, who came up with an expression which would describe the energy associated with EM waves, as coming in integer numbers of fixed amounts (or "quanta"). Each quantum of energy having energy $E = hf$ where E = energy, f = frequency of the EM waves, and h = "Planck's constant" = 6.63×10^{-34}. Whereas Planck considered this to be merely a convenient formula, it was Einstein who realized what it actually meant: that light comes in discrete little packages (photons). The energy of each photon depends only on the frequency. This particle view of light was confirmed by experiments concerning the photoelectric effect - see 27-3 for the details, and I will talk about it in class. The mathematical results are that the photons have energy $E = hf = pc$ (where p is the momentum ... notice that $p = hf/c$ ... yes, light has momentum!). Also, electrons don't go flying out of metal objects because they are attracted to all of those positive nuclei. If a photon has enough energy, it can kick an electron right out of a piece of metal (the energy needed to kick an electron out is called the "work function" (W). Any extra energy a photon has in excess of W ends up as kinetic energy in the electron. Hence $U_{k\,max} = hf - W$.

Physics of the Atom

Okay, so we figure that maybe matter is composed of these little things called atoms, but what are atoms made up of? The first piece discovered was the electron, which is easily removed by applying a large voltage across a sample of low pressure gas. Historically, the name "electron" came later ... they were first called "cathode rays" since they came from the cathode (negative end). The first significant direct measurement of electrons was made by J.J. Thomson, who was able to determine the "charge to mass ratio" (q/m) for them. See pages 772-773 for the details. Thomson is also credited with the "plum pudding" model of the atom - in which positive charge is uniformly spread out in an atom (like pudding) and electrons are little particles lodged in it (like plums). On another historical note: Millikan later measured the charge on an electron, which then allowed us to determine the mass of an electron (knowing the q/m ratio).

The plum pudding model of the atom was shattered by Ernest Rutherford. Rutherford directed a beam of high speed alpha particles (helium nuclei) at a very thin sheet of gold foil. Consider these alpha particles to be like bullets being shot at tissue paper. The expectation is that the alpha particles are going to go straight through the foil with little scattering (which would make sense if atoms have positive charge spread out in too low a concentration to significantly repel the alpha particle - as Thomson's model states. Rutherford in fact observed most of the alpha particles going straight through, AND some of the alpha particles being scattered all over the place (imagine a bullet bouncing right back off of tissue paper).

To explain this, he gave a new description of the atom in which all of the positive charge is located in one very small central “nucleus” ... on occasion, an alpha particle would be wildly deflected by a close encounter with it. He then added that the electrons must “orbit” the nucleus in a similar way to the planets orbiting the sun. Hence Rutherfords “planetary model” of the atom. This is probably as detailed as you have come across before, in terms of physics of the atom. Before I go any further, I need to tell you about matter waves.

You may still be asking “is light a wave or a particle?” - and I can assure you that the answer is “yes”. Of course, a better answer might be “light is really unlike anything you have direct experience with (or think you have direct experience with); it does behave like waves (interference, diffraction, etc)_and it also behaves a lot like particles (has momentum, can undergo collisions, etc). One important particle like behaviour of light is known as the “Compton Effect”. We do not need a mathematical understanding of this effect, but it basically says that photons can interact with other particles - when they do, they obey all of the usual laws ... momentum and energy are always conserved. In other words, we can treat the interaction between light and particles as being collisions! The mathematics becomes more complicated, as the theory of relativity needed to analyse these “relativistic collisions”. It will suffice for you to know that if a photon hits a particle, that if energy is transferred to the particle, then the photon will lose energy: it will have a lower frequency / larger wavelength after the collision.

Light is neither a wave nor a particle, so let’s invent a new phrase called “wave-particle duality” which basically means “both”. One thing you should know is that light distinctly exhibits one of these two aspects at a time, depending on what you’re doing to it. (You may not like this, but that’s just too bad).

Well, if we’re going to let photons be both wave and particle, we have to let all the other particles (including protons, electrons, bowling balls, etc) have this privilege too, or else we may be looking at a particle strike. Louis de Broglie determined an expression which will determine the wavelength for matter ... a “matter wave”. In it, $\lambda = h/(mv) = h/p$ (which you can show is exactly the same as for photons of light). Ordinarily, matter has wavelengths that are ridiculously small, since h is so small, and p is usually quite large. Careful experiments have confirmed that matter does exhibit wave properties. In other words, matter (usually in the form of electrons) undergo diffraction, and can even interfere with each other. If you feel okay about this, then you don’t really understand what we’re saying.

Back to the atom. Bohr modified the Rutherford model of the atom in a couple of ways. For one thing, not just any orbit is allowed. Only certain energy states are possible. Realize that Bohr did this quite arbitrarily, without proof or reason. We now understand this to be essentially true, and explain it in terms of matter waves. One interpretation is that the electron wave can only surround the nucleus at certain locations, where standing waves result, that is the electron orbit must satisfy $2\pi r = n\lambda$ (see page 797). In any case, the result is that the electrons exist only in “allowed” orbits, and each orbit has a precise amount of energy associated with it. Electrons become excited when they absorb energy (such as from a photon) to kick them into a higher state, and lose energy (such as by emitting a photon) when they drop to a lower state. If an electron absorbs enough energy, it will be removed from the atom, leaving it ionized. The lowest energy level is known as the “ground state”, where n = 1. It will be important for you to work with problems of this nature. Realize however, that the picture of an atom just described is truly not the most up to date picture of an atom. The truth is that atoms are incredibly complicated - you had better sign up for physics in university.

Note that the electron volt is the standard energy unit when discussing atoms (1 eV = 1.6×10^{-19} J)
One last thing: $E=mc^2$, I’ll talk about this in class too.
HW: Read chapter 27, Questions 1-10, 14-16, 22-26, Prob 1-3, 11-29 odd, 33-39 odd 43-51 odd

Physics 42A class #21

Today's class: Sections 30-1 to 30-9

Much of this chapter should be familiar from chemistry. Please note that you must not "brush off" this material thinking that it is not important. There are usually a fair number of questions on the AP exam on the material here. The concepts are easy, and so, if you read through this material, it's free marks.

We now believe the nucleus to consist of two particles: protons and neutrons, which we collectively call "nucleons". Note that a particle view is adequate, although we acknowledge that they too have wave properties, as per the wave/particle duality mentioned previously. SI units of are usually not the preferred units when dealing with mass and energy at the atomic level, as kilograms and joules are far too large to be convenient at this level. Instead, mass is measured in "unified atomic mass units (u), where $1\text{ u} = 1.66 \times 10^{-27}$ kg, and energy is measured in electron volts (eV) where $1\text{ eV} = 1.6 \times 10^{-19}$ J. Because of mass-energy equivalence ($E=mc^2$), 1 u of mass is equivalent to 931.5 MeV of energy.

One thing that always troubled me was why it was necessary to have the masses of the atoms given in the periodic table. It seemed unnecessary, as you should be able to calculate them, based on knowing how many protons electrons and neutrons are contained in an atom. For example, you can calculate the mass of a helium atom by adding the masses of 2 protons, 2 neutrons, and 2 electrons. If you do this, you will find that the calculated mass is 4.032988 u, when in fact (from the periodic table) it is 4.002602 u, which means that a helium atom is 0.030386 u less massive than the collection of 2 protons, 2 neutrons and 2 electrons. This puts a twist on the saying "the whole is greater than the sum of the parts".

Atoms are less massive than the sum of the masses of the individual particles because of "Binding Energy". When nucleons form a nucleus, energy is released. This "new energy" comes at the expense of some mass being lost. In fact, it is the binding energy that results in a nucleus being stable, as the nucleus cannot break apart into individual nucleons, unless a sufficient amount of energy is provided (again - the binding energy).

Another problem I had was in understanding why the protons (positive) don't repel each other within the nucleus, causing it to break apart. In fact, they do repel each other, as Coulomb's Law would suggest, however another force acts to hold them together ... the "Strong Nuclear Force", which we really don't know much about, other than that it only acts over very small distances (ie within a nucleus).

Of course, not all nuclei are stable. Some nuclei will spontaneously change, releasing energy in the process. This phenomenon is called Radioactivity, and the nuclear change itself is called decay, of which there are three main types: Alpha, Beta, and Gamma. Note that two conservation laws govern all types of decay: the law of conservation of charge, and conservation of mass number.

<u>Alpha Decay</u> is when an alpha particle (helium nucleus) is emitted from a nucleus. Obviously, the nucleus from which it came loses two protons and two neutrons, so the daughter nucleus will have Z-2 and A-4 (where Z is the atomic number of the parent, and A is the atomic mass number of the parent). Since the alpha particle is quite massive, alpha radiation is not very penetrating.

<u>Beta Decay</u> is when the nucleus emits a "beta particle" which is now known to be an electron (another particle is also emitted: a neutrino). The electron comes from within the nucleus, as a neutron changes into an electron and proton. This means that the daughter nucleus will have Z+1 and A that is unchanged. Beta radiation is considerably more penetrating than alpha.

Gamma Decay involves the nucleus emitting a photon of very high energy. This very high energy photon is very penetrating, and is the result of the nucleus itself dropping to a lower energy level.

The decay of a single unstable nuclei is unpredictable as to when it will decay. In fact, the best we can do is to describe it as being random, similar to the toss of a coin. Large numbers of unstable nuclei however, decay in a predictable, statistical manner. The decay of a sample of such nuclei is said to be exponential, as the rate of decay depends on the number of unstable nuclei present, which is a decreasing number as they decay. A very useful description of the rate of decay is the half life, which is a period of time, which will result in about half of the unstable nuclei decaying.

Homework:
- Read Sections 30-1 to 30-9
- Questions 1-18 omit 15

Physics 42A -Class #22

Today's class: Sections 13-1 to 13-9, omit 13-5

Here we enter our study of Thermal Physics. We will be looking at chapters 13 to 15. If you have taken chemistry in grade 11, you should find the beginning of chapter 13 very straightforward. Before you jump to the conclusion that we're doing chemistry, I'll remind you that chemistry is just a branch of physics. (You may even start to agree with me after this chapter, and considering how anything to do with electric charge is ours). This chapter is very readable, and so you should try to be less dependent on my notes, as I won't have all the details in them - focus on reading the chapter.

As familiar as the concepts of heat and temperature are, most people do not have a good understanding of what these terms actually mean. One thing to note is that you cannot trust your own personal experience with these phenomenon. A common misconception is that your body can "sense" temperature. It can't! Metal feels colder than wood, even if they are at the same temperature. A -20° C day may be bearable one day, if there is no wind, but seem much colder another day if there is a wind (ie windchill), even though the temperatures of the two days were the same. Comfortable room temperature is about 20° C; try jumping in a swimming pool also at 20° C and you'll feel extremely cold.

The problem is that your body actually senses energy transfers, and not temperature directly. Thermal energy is transferred between two objects when they are at a different temperature. The energy that is transferred is called "heat". If you touch wood at 0° C, heat will flow from your hand to the wood, and you perceive this as "cold". If you touch metal at 0° C, heat will again flow, but at a greater rate because metal is a better "thermal conductor" than the wood. You perceive this greater transfer as "colder" even though it is at the same temperature.

Two objects allowed to exchange thermal energy with each other are said to be in "thermal contact" (they need not be touching). Heat will flow as long as the temperatures are different. When heat stops flowing, they are said to be in "thermal equilibrium", and therefore at the same temperature. The zeroth Law of Thermodynamics got its unusual name because it seemed too obvious to even officially state as a law at first. (First came the 1st and 2nd, when it was realized that this law was technically needed before the 1st law could be proved). The zeroth law simply states that *if two systems are in thermal equilibrium with a third system, then they are in thermal equilibrium with each other.* This makes temperature a useful concept, as it means that if object A and B have reached thermal equilibrium, and B is also in thermal equilibrium with C, then A and C are necessarily also in thermal equilibrium. This makes temperature a meaningful concept, as we can now officially say that all things at the same temperature will be in thermal equilibrium with each other.

Even though our bodies are not good indicators of temperature, it is an extremely important concept, as many physical characteristics are temperature dependent. The first characteristic is "size" - the size (length, area, volume) of objects is to some extent temperature dependent. The concept is called "thermal expansion". Read 13-4 for the details and relevant formulas.

Sections 13-6 13-9 are very much covered in chemistry. If you have not taken this in chemistry please see me for the details - one thing that may (?) not be covered is that pressure has the SI unit of the "Pascal" (Pa) where P=F/A, one pascal is defined as one Newton per metre squared: $1Pa = 1N/m^2$

Homework: - Read Sections 13-1 to 13-9, omit 13-5
- Questions 1-14 - Problems #7, 9, 10, 12, 28, 34, 40, 41

Physics 42A - Class #23

Today's class: Sections 13-10 to 13-14

The Kinetic Theory is one of the building blocks of modern science (note science, and not just physics). Kinetic Theory is simply the idea that matter is made up of little pieces called atoms which move around. The word "theory" in the name of a scientific principle is often misused by nay-sayers along the lines of "it's only a theory". I don't much care for that line of thinking. A theory is not "just a theory" but rather an explanation that we have come up with for how something in the universe "works". Kinetic Theory is very well founded and backed by large amounts of evidence and experimental data.

A theory becomes a good theory if it can explain details about observed phenomenon. Kinetic Theory can explain much about matter and its properties, including Boyle's Law (pressure and volume are inversely related). The proof is in the text, and I'll go over it in class. The mathematical result reveals the relationship between the kinetic energy of gas molecules and temperature (that they are directly related to each other). It also gives insight into what temperature really is: Temperature is a measure of the average kinetic energy of the particles that make up matter.

$$\overline{KE} = \frac{1}{2} m\overline{v^2} = \frac{3}{2} kT \qquad (k = 1.38 \times 10^{-23} \text{ J/K, “Boltzmann's constant)}$$

One important thing to realize in the above equation is that we cannot find the actual speed of an individual molecule. We cannot even directly find the average speed. The equation contains $\overline{v^2}$ which is the average speed, squared. If you take the square root of this you do NOT obtain the average speed (try it with some numbers: the square root of the average of the squared values is not (usually) the average of the numbers themselves). The number obtained is the "square Root of average, or Mean of the Squared values - rms for short. So using the above equation, v_{rms} can be found.

The rest of the chapter (13-11 to end) I would like to look at qualitatively, in less detail - you should have an "awareness" of the material. Of particular interest to me is that you know:

- about the Maxwell distribution of speeds, and what the distribution curve looks like
- what "triple point" means
- what evaporation and condensation are, in terms of Kinetic Theory (why do you feel cold when wet?)
- the physics of a boiling liquid (what is so special about 100° C, when water boils?)
- what diffusion is, in terms of Kinetic Theory

Homework:
- Read Sections 13-10 to 13-14
- Questions 15-24, omit 20, 22
- Problems #42, 43-51 odd, 57, 70, 74, 78
- Read ahead into Chapter 14

Physics 42A - Class #24

Today's class: Sections 14-1 to 14-5

We are familiar with the concept of "thermal energy", as we came across it in grade 11 physics. Thermal energy is the energy of the random movement of large numbers of molecules. When two objects of different temperatures come in thermal contact, thermal energy will be transferred from the hotter object to the cooler object. Heat specifically refers to energy that has been transferred from one object to another, as a result of a temperature difference. Note that kinetic theory explains this nicely, as the collisions between fast (hot) and slow (cool) molecules.

Back when temperature and heat were not well understood, the calorie (cal) was introduced as being the official unit of heat, where one calorie was the amount of heat needed to raise 1 gram of water from 14.5° C to 15.5° C. Now that we know that heat is energy, the calorie is really not needed anymore, as the joule can be used. As you are well aware, however, the calorie is still very much in use. The "mechanical equivalent of heat" is simply the observation that heat is a type of energy, and that 1 cal is the same as 4.186 J. (Note that a Cal - capital "C" - is sometimes called a "food calorie" and is 1000 cal). It is also useful to know that 1 cal raises 1 gram of water by one degree Celsius.

The terms "thermal energy" and "internal energy" both mean the same thing: how much total energy is contained in an object, due to the random movement of the molecules (KE).

"Temperature" is a measure of the average kinetic energy of the molecules of an object.

"Heat" refers to the transfer of thermal energy, due to a temperature difference.

Note that the thermal energy depends on two things: How many molecules (or moles) an object contains, and how hot (temperature) it is. Specifically: $U_{Th} = 3/2\ NkT = 3/2\ nRT$ where N = number of molecules, and n = number of moles.

The temperature of an object will rise when you add heat to it. The thermal energy contained in it will be the sum of the thermal energy it started with, and the heat added to it. As a result, the temperature will rise because the KE of the molecules has obviously increased. It would be reasonable to think that adding heat (Q) will affect the temperature (T) of some mass (m) of any substance in the same way. This is not true: each substance has a unique value for what is called the "specific heat" (c), and $Q = mc\Delta T$

The equation above is useful for relating additions of heat with temperature changes, and is particularly useful for calorimetry - which is simply a setup in which substances of different temperatures are placed in an isolated container (a calorimeter). The two substances will certainly come to thermal equilibrium, which means that their temperatures will reach some value, T. The important thing to recognize here is that this resulted because the hotter object lost heat, while the cooler object acquired heat, and so, by knowing the specific heats of the materials involved, one can determine the end temperature.

"Bomb Calorimetry" isn't nearly as exciting as it sounds: it is the burning of a substance (usually food) while in a calorimeter, and determining the amount of heat gained as a result - the calorie content.

Homework: - Read Sections 14-1 to 14-5
- Questions 1-4
- Problems # 1-19 odd, 14, omit 11

Physics 42A - Class #25

Today's class: Sections 14-6 to 14-9

Recall that heat is a transfer of thermal energy. Heating a substance can do one of two things. The added heat can certainly cause the objects temperature to rise (remember that temperature is a measure of the average KE of the molecules). But surprisingly, adding heat does not necessarily lead to a temperature rise. At certain critical temperatures, the addition (or removal) of heat can lead to a <u>change of phase</u> in the substance. By "phase" we mean the three primary states of matter: solid liquid and gas (plasma being the fourth state, which we will not be discussing here).

The equation $Q=mc\Delta T$ is only valid for regions (ΔT) where a phase change does not occur. At the point of the phase change, additional energy will be needed (or released, depending on how the phase change is to occur - solid to liquid, or liquid to solid for example). <u>Latent Heat</u> (symbol l) is the term describing this extra quantity of heat involved in the phase change. l is defined as being the amount of heat per unit mass involved in the phase change. This means that

$Q = ml$	where l may be l_F (Latent heat of <u>F</u>usion) - Solid / Liquid phase change or l_V (Latent heat of <u>V</u>aporization) - Liquid / Gas phase change **see the table on page 408 for specific values**

Note that we only need to know sections 14-7 to 14-9 qualitatively (non mathematically). The main idea in these three sections is that heat can get from one place to another in primarily three different ways: <u>Conduction</u>, <u>Convection</u>, and <u>Radiation</u>.

Conduction, as the name implies, is a transfer of heat <u>by contact</u>. It is similar to electrical current, however, whereas current has electrons that move from one object to another, Thermal conduction has only the thermal energy flowing from one place to another.

Convection means that the thermal energy is being <u>carried by particles</u> (molecules / electrons / etc) while these particles are moving. For example, heat is transferred from the element of a stove to pasta in a pot by convection: Water is heated at the bottom of the pot (by conduction). The heated water then rises, carrying the thermal energy to the top of the pot, and to the pasta (convection).

Radiation is much more complicated, especially since we haven't yet studied light. Basically, all "hot" objects <u>emit electromagnetic radiation</u>. The filament of a light bulb, for example, is so hot that it glows, as does a hot element. However, even when you don't see a glow, there is still the emission of electromagnetic radiation - of a lower energy, non-visible type. In particular, Infrared radiation is associated with heat emission of this form. This is the principle involved in some "night vision" devices. All objects around you, including you, are "hot" (around 300 K!). They are so hot that they are glowing! Not with visible light, but with Infrared "light" which we cannot see, but can make machines that can see it (think of the movie Predator).

Homework: - Read Sections 14-6 to 14-9
- Questions 5-20
- Problems # 21 - 25
- Chapter 13 / 14 Test

Physics 42A - Class #26

Today's class: Sections 15-1 to 15-4

This is the heavy chapter of this unit. Note that you need to have a good understanding of chapters 13 and 14 before you can really understand the new material here.

Thermodynamics is the study of transfers of energy by work and heat. Work is a mechanical transfer of energy (a force acting through a distance), while heat is a transfer of energy due to a temperature difference. The first thing you have to have an understanding of is the word "system". We used this word before, when we talked about momentum. A system is simply the collection of objects. Anything not in the system is referred to as the "environment". Systems may be closed (only energy, and not matter are able to be exchanged with the system); open (matter and energy are able to be exchanged); or isolated (nothing is exchanged with the system).

The **First Law of Thermodynamics** is really a statement of the Law of Conservation of Energy. It says that the net change of internal energy for any system depends on two things: the amount of work done, and the amount of heat transferred. Note that this makes sense, as these are the only two ways in which energy can be transferred. There is, however, the possibility of confusing the signs (+ or -) inside the equation form of this law, which is:

$\Delta U = Q - W$ (Q = heat added ***TO*** system; W = work done ***BY*** system)

As far as systems go, we will be mostly interested systems which consist of ideal gases. Note that gases can be ① Heated or Cooled (changed in temperature), ② expanded or contracted (changed in volume), ③ increased or decreased in pressure. These changes can go about in several ways, referred to as "processes"

Process	Condition
Isothermal	Temperature remains constant (heat may be added or removed to ensure this) $W = Q$
Adiabatic	Heat is not exchanged between system and environment; ie $Q = 0$ and so $\Delta U = -W$
Isobaric	Pressure is kept constant
Isochoric	also known as isovolumetric, volume is held constant; $W = 0$

Note that a work is done on a sample of gas when it is compressed, and gas does work when it is allowed to expand. If the pressure is constant, then the work done by a gas is $W = Fd = (PA)d = P\Delta V$

The **Second Law of Thermodynamics** is a law describing that some things never happen, even though they would not violate any other laws of physics, if they were to occur. It is actually a complicated law that we will revisit later, but for now, one aspect of it is to say that "heat flows naturally from a hot object to a cold object; heat will not flow spontaneously from a cold object to a hot object".

Homework: - Read Sections 15-1 to 15-4
- Questions 1-8
- Problems # 1-15 odd
- Read ahead

Physics 42A - Class #27

Today's class: Sections 15-5 to 15-10

Heat is a form of energy, and as we saw in grade 11 physics, energy can be transformed from one type to another. Some energy transformations are easy to do: gravitational potential energy can be transformed into kinetic energy very easily, and kinetic energy can be transformed into thermal energy very easily as well (as this happens every time you drop something).

Thermal energy can be transformed into other types of energy. It is always being turned into "light energy" (energy of electromagnetic radiation, actually) as hot objects radiate - visible light if hot enough, infrared etc if not. Turning thermal energy into mechanical energy (getting heat to do work) is quite another story. It may be done by the use of a "Heat Engine".

A heat engine extracts thermal energy out of a hot object (cooling it somewhat in the process). The heat is transformed into mechanical energy, as it does work via the heat engine itself. It is important to understand that a heat engine operates at two temperatures: a "high" and "low" temperature. This is essential, as it is the lowering of the temperature of something that extracts the energy, which is used to do work. Since energy is conserved, the relationship $Q_H = W + Q_L$ holds. This equation simply says that the energy is the same before (when hot) and after (when cold, when the amount of work done is taken into account).

The efficiency of a heat engine is a measure of how good it is at "squeezing the energy out" of a hot object. It is the ratio of (how much work is done) to (how much heat energy there is). $e = W/Q_H$. Combining this with the equation $W = Q_H - Q_L$ results in $e = 1 - Q_L/Q_H$. In all cases, the efficiency "e" should be expressed as a percentage by multiplying it by 100.

Clearly, a perfect heat engine would extract all of the thermal energy out of a hot material, and use it to do work. But this would leave the material at absolute zero (no thermal energy at all) - which is a theoretical impossibility itself (you can get close, but never actually achieve "0 Kelvin"). This leads to another way of stating the 2nd Law of Thermodynamics: "no device is possible whose sole effect is to transform a given amount of heat completely into work".

Another important aspect of real heat engines is that not all of the extracted thermal energy can actually be used to do work - some of it will be lost in the engine, and reappear as heat! (due to friction, turbulence, etc.) An idealized engine would have none of these "real life problems" and is given the name "Carnot Engine". A Carnot engine would take material at a high temperature "T_H", cool it to a lower temperature "T_L". For this perfect Carnot engine, the efficiency can be found in terms of temperature, (T) as opposed to energy content (Q): $e_{Carnot} = 1 - T_L/T_H$. Note that even for the Carnot engine, the efficiency is not 100% as T_L can not be 0. The Carnot efficiency is the best possible efficiency for a given set of operating temperatures. A well designed, real engine performs at about 60 to 80 percent of the Carnot efficiency.

Note that you should know the rest of the chapter (15-6 to 15-12, omit 15-11) qualitatively. Reading through it should be sufficient. I will talk about entropy in class.

Homework: - Read Sections 15-5 to 15-12, omit 15-11
- Questions 9-23
- Problems # 16-25 omit 22 - Review chapter