Physics
R.J. Marks II Class Notes
(1968-1970)

Physics I. Assignments_-Winter Quarter 1968-69

| Date | Assignment E Problems |
| :---: | :---: |
| 1/7 | Chap I. |
| 1/8 | Chap II, 2, 4, 5, 7 |
| 1/9 | Chap II, 11, 12,14.18 |
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| 1/14 | Chap III. 6, 8, 9, 10 |
| 1/15 | Chap III. 12,13,15,16 |
| 1/16 | Chap III, 23,24,29,31 |
| 1/17 | Test II \& III |
| 1/21 | Chap IV, 2, 4,5,7 |
| 1/22 | Ghap IV, 8, 10, 12,13 |
| 1/23 | Chap IV, $22,24,29,34$ |
| 1/24 | Chap IV, 20, V. 1, 4 |
| 1/28 | Chap V. 5, 6, 7, 8 |
| 1/29 | Chap V, 10, 12, 15, 17 |
| 1/30 | Chap V. 19, 20, 22, 29 |
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| 2/4 | Chap VI, 9, 10,11,12 |
| 2/5 | Chap VI. 16,18,20,25 |
| $2 / 6$ | Test IV. V, VI |
| $2 / 7$ | Chap VII, 2, 3, 5,6 |
| 2/11 | Chap VII, 9,12,13,15 |
| 2/12 | Chap VII, 19, 20, 21,22 |


| $\begin{aligned} & \text { pate } \\ & 2 / 13 \end{aligned}$ | Assignment \& Problems Chap VIII, 1, 2.5 |
| :---: | :---: |
| $2 / 14$ | Chap VIII, 3, 6, 7,9 |
| 2/18 | Chap VIII, 11, 12,13,14 |
| 2/19 | Chap VIII. 15,24,25,26 |
| 2/20 | Test VII and VIII |
| $2 / 21$ | Chap IX, 1, 5, 6 |
| 2/25 | Chap IX, 7,9,10,11 |
| 2/26 | Chap IX. 14, 16.17 |
| $\underline{2 / 27}$ | Chap IX. 19.22.27 |
| 2/28 | Chap X . 2, 3, 4,6 |
| $3 / 4$ | Chap X, 9.10.13.15 |
| $3 / 5$ | Chap X. 19,22, 27 |
| $3 / 6$ | Chap X, 29,31,33 |
| $3 / 7$ | Test Chap. IX and $X$ |
| $3 / 11$ | Chap XI, 3, 4, 6, 8 |
| 3/12 | Chap XI, 9, 10,13.16 |
| 3/13 | Chap XII, 1, 4, 5 |
| 3/14 | Chap XII, 13. |
| ! | FINAL EXAM. |
| 1 | $45 \quad 10 W K$ |




## 4 <br> 4 4 <br> $\sqrt{n}$

$$
\begin{gathered}
\text { 14) a) } 2.5 \mathrm{~m} 5 \\
\text { b) } 1.25 \mathrm{mN} \\
\text { e) } 6.25 \mathrm{~m} \mathrm{~S} \\
\text { d) } 10 \mathrm{~m} \mathrm{~N}
\end{gathered}
$$





$$
\begin{array}{ccc}
+1 & 0 & 0 \\
0 & 0-0 & 0 \\
0 & 0 & 0 \\
0 & 0 & 0 \\
0 & 0 & 0 \\
0 & 0
\end{array}
$$



$$
P_{6} 5 \leqslant
$$

$$
\begin{array}{r}
P_{8} s g \\
\text { 10) } \quad 50 \frac{M}{1}
\end{array}
$$

10
40
10
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11
in
$n$
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## 

$$
\frac{1}{5}
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$\mathrm{B}^{2}$

Pu 2
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& \text { 4 }
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\begin{aligned}
& \text { \% }
\end{aligned}
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$$
\begin{gathered}
F_{1}=F_{m} \\
F_{1}=M\left(\frac{p+m}{m+m}\right.
\end{gathered}
$$

S

$$
\begin{aligned}
& a=x=5 \sqrt{3} \quad 1=5
\end{aligned}
$$





P10



$$
\begin{aligned}
& 1+0 \\
& n \\
& \text { Po }
\end{aligned}
$$




$$
\begin{array}{r}
P_{8} 53 \\
10) \quad 30 \frac{9}{4}
\end{array}
$$

006
1101
361

$$
\begin{aligned}
& S \approx H \& k^{2}
\end{aligned}
$$

5
5
0
$b$

18
$\stackrel{\circ}{6}$
11
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an
4 18
MA U 裙


$$
\begin{gathered}
a=b \\
v=5 t \\
y=30 \\
(t=6414) \\
\frac{1+t^{2}+C_{2}}{L^{2}} \quad C_{2} \\
(6)^{2}=5(18)=
\end{gathered}
$$

r


PS $54-2$



8







$N$
$N$
+4y
20)



$\frac{44^{6}}{46}$


4
8
8
$140+\infty$ (Barab whatar



$$
\begin{aligned}
& a=x=5 \sqrt{3} \quad 1=5
\end{aligned}
$$





P10



$$
\begin{aligned}
& 1+0 \\
& n \\
& \text { Po }
\end{aligned}
$$




$$
\begin{array}{r}
P_{8} 53 \\
10) \quad 30 \frac{9}{4}
\end{array}
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\begin{aligned}
& S \approx H \& k^{2}
\end{aligned}
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$$
\begin{gathered}
a=b \\
v=5 t \\
y=30 \\
(t=6414) \\
\frac{1+t^{2}+C_{2}}{L^{2}} \quad C_{2} \\
(6)^{2}=5(18)=
\end{gathered}
$$

r


PS $54-2$



8







$N$
$N$
+4y
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$\frac{44^{6}}{46}$


4
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$140+\infty$ (Barab whatar



$$
\begin{aligned}
& a=x=5 \sqrt{3} \quad 1=5
\end{aligned}
$$





P10



$$
\begin{aligned}
& 1+0 \\
& n \\
& \text { Po }
\end{aligned}
$$




$$
\begin{array}{r}
P_{8} 53 \\
10) \quad 30 \frac{9}{4}
\end{array}
$$

006
1101
361

$$
\begin{aligned}
& S \approx H \& k^{2}
\end{aligned}
$$

5
5
0
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$\stackrel{\circ}{6}$
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MA U 裙


$$
\begin{gathered}
a=b \\
v=5 t \\
y=30 \\
(t=6414) \\
\frac{1+t^{2}+C_{2}}{L^{2}} \quad C_{2} \\
(6)^{2}=5(18)=
\end{gathered}
$$

r


PS $54-2$



8







$N$
$N$
+4y
20)



$\frac{44^{6}}{46}$


4
8
8
$140+\infty$ (Barab whatar





$$
\begin{aligned}
& \text { Pg } 12>46 \\
& 10) \\
& 1^{n t} \sqrt{4 k}
\end{aligned}
$$







ytu $s 418 d$


$$
\begin{aligned}
& \int_{4}^{5} \\
& \begin{array}{c|c|c}
40 & \quad 4 \\
4
\end{array} \\
& \begin{array}{c}
8 \\
0 \\
0 \\
0 \\
4
\end{array} \\
& \begin{aligned}
\Sigma F_{r} & =T-m g \\
T & =m a v^{2} \\
& =m \frac{I g}{G} \\
& =m \frac{1}{g} \\
& =4 m g
\end{aligned} \\
& \operatorname{coc}_{4}^{4} \\
& \begin{array}{c}
c \\
0 \\
0 \\
4 \\
4
\end{array}
\end{aligned}
$$





Z 506 c


$$
\frac{(9)}{0} \operatorname{mar}^{4} \times 5 \cdot \frac{(6)}{2}
$$














$$
\begin{aligned}
& x_{\infty}=\frac{\sum m x}{\sum m} \quad ; \quad y_{0}=\frac{\sum m b y}{\sum m}
\end{aligned}
$$



31
A' io ho block th pushall acuns a smod Fonce of 3Hs. H the blake stath a from $\operatorname{sos} 7$ at A, how morel wenk As dome by the 3/f fobedey as bract mouts A Pan A to s. What in the anoroge


R $x^{4} 2$
A Mock usenghing 2 Ha as fromed tagit

## Physics 1 Test I J

Name BLC
Score $\qquad$
I (i) Two vectors $\vec{A}$ and $\vec{B}$ lie in the $x y$ plane as indicated. Each has a magnitude of 10 units. A third vector $\vec{C}^{-}=-3 \hat{j}+4 k$. Circle the correct statements of those that follow. The vectors i, 角, and k are unit vectors pointing respectively along the $x$, $y$ and $z$ axes.
(a) $A_{X}=10$
(b) $B_{y}=-6$
c) $|\stackrel{\rightharpoonup}{\mathbf{c}}|=5$
(d) $\vec{A}+\vec{B}=2 \hat{i}-6 \hat{j}$
(e) $3 \hat{C}+\vec{A}=10 \hat{A}-9 j+12 \hat{k}$
f) $\vec{A} \cdot \vec{C}=50$

12
g) $\vec{C} \times \vec{A}=0$
(h) $\overrightarrow{\mathrm{A}} \times \overrightarrow{\mathrm{B}}=-60 \hat{\mathrm{k}}$
(i) $\overrightarrow{\mathrm{A}} \cdot(\overrightarrow{\mathrm{B}}+\overrightarrow{\mathrm{C}})=150 \times$
(j) $\overrightarrow{\mathrm{B}} \cdot \overrightarrow{\mathrm{C}}=18$
(ii) Sketch on the figure the vector $\vec{B}-$ -
(iii) Calculate the angle between $\vec{B}$ and $\vec{C}$.



Physics I Test I January 17． 1969


I（i）Two vectors $\vec{A}$ and $\vec{B}$ lie in the my plane as indicated． Each has a magnitude of 10 units．A third vector $\overrightarrow{\mathrm{C}}=-3 \hat{\mathrm{j}}+4 \mathrm{k}$ ．Circle the correct statements of those that follow．The vectors 今，今，and 令 are unit vectors pointing respectively along the $x$ ． $y$ and $z$ axes．
$\sin 30^{\circ}=.50$
（a） $\mathrm{A}_{\mathrm{X}}=10$
$\cos 30^{\circ}=.87$
（b）$B_{y}=-6$
$\sin 53.2^{\circ}=.80$
c）$|\overrightarrow{\mathrm{C}}|=5 \times$
$\cos 53.2^{\circ}=.60$
（d）$\vec{A}+\vec{B}=2 \hat{i}-6 \hat{j}$
（e） $3 \vec{C}+\vec{A}=10 \hat{a}-9 j+12 \hat{k}$
（f）$\vec{A} \cdot \vec{C}=50 \times$
12 （g）$\vec{C} \times \vec{A}=0 x$
（h）$\vec{A} \times \vec{B}=-60 \hat{k}$
（i）$\vec{A} \cdot(\vec{B}+\vec{C})=150 X$
（5）$\vec{B} \cdot \vec{C}=18$
（ii）Sketch on the figure the vector $\vec{B}-\vec{A}$ ．
（iii）Calculate the angle between $\vec{B}$ and $\vec{C}$ ．
2) An object moving in a straight line along the $x$ axis starts from the origin at time to. The figure below shows how its instantaneous velocity depends on time, motion to the right being represented by positive values of $v$ and motion to the left by negative values.


Determine:
(a) the magnitude and direction of the displacement of the object over the interval from $t=0$ to $\mathrm{t}=15$ seconds.

$x=\frac{a}{2} t^{2}+V_{0 t}$

$$
=3(10)^{2}-5(10)
$$

$=300-50$
=25 Oft
(6)
(b) the magnitude and direction of the average velocity over the interval from $t=0$ to $t=15$ seconds.
(6)
(c) The magnitude and direction of the average acceleration over the interval from $t=15$ to $t=25 \mathrm{sec}$ ends.
(6)
(d) the magnitude and direction of the instantaneous acceleration immediately after time $t=15$ seconds.

3.) From a high cliff a man shoots a body A straight up and a body B straight down with the same initial speeds $V_{0}=29.4 \mathrm{~m} / \mathrm{sec}$.
(6)
(a) At what time is the speed of body B twice that of body A?

$$
\begin{aligned}
& v_{B}=a_{x, t} t+v_{0 x_{0}} \quad v_{B}=a_{x} t+v_{0 x} \\
& V_{A}=9.8 t-29.4 \mathrm{~V} \quad V_{B}=9.8 t+29.4
\end{aligned}
$$

$2 V_{A}=19.6 t-58.6 \quad 2 V_{B}=19.6 t+58.6$
4

$$
t=90 \mathrm{pec}
$$

$19.6 t-58.6=9.8 t$
$9.8 t=88.0$
(6)
(b) What is the distance of separation between the bodies at that time?

$$
\begin{array}{r}
x=\frac{a}{2} t+V_{0_{A} t} \\
V_{\text {ox A }}=29.4
\end{array}
$$

$$
X_{B}=\frac{a}{2} t^{2}+V_{0 x_{4}}
$$

$$
\text { Vex: } 29.4
$$

$$
\begin{aligned}
& 4=9.8 x_{n}^{2} 29.4 \\
& t=20
\end{aligned} x_{B}=4.6(90)^{2}+29.4(90)
$$

$$
=3.72 \times 10^{2}+.365 \times 10^{2}
$$



$$
x_{a}+x_{b}=7.44 \times 10^{2} \mathrm{~m}
$$



A shell is fired straight upward and travels a distance of 543.9 m during the third second. Neglect air friction. Assume all motion is upward during the third second.
(6)

(c) What is the total flight time of the shell?
$T_{k}=={ }^{2} t^{2}+V_{0 x} t \quad Q=9.8$

$196=9.8 t$
$t=200 \mathrm{mes}$

(6)
(d) What was the initial speed of the shell as it O left the ground?


Name BOB MARKS

To receive credit on any test question it is necessary to indicate clearly how you arrived at your answer.
I. (A) A projectile is fired with an initial velocity of $160 \mathrm{ft} / \mathrm{sec}$ reaches a maximum height of 200 ft . What angle did its initial velocity vector make with the ground?


$$
\begin{aligned}
& Y=\frac{a}{2} t^{2} \quad V_{C_{6}}=160 \sin \theta \\
& -Y=16 t^{2} \\
& \begin{array}{l}
200=16 t^{2} \\
t=\frac{1026}{16}+\frac{141}{4}=3.53 \mathrm{sec} 0
\end{array} \\
& V_{4}==(3.53)(32)=160 \sin \theta \\
& \min =705
\end{aligned}
$$

(B) A 2000 lb. car is moving around a circular race track at a constant speed of $90 \mathrm{ft} / \mathrm{sec}$. The radius of the track is 500 ft.

What is the acceleration of the car? (Magnitude and direction)


What is the frictional force exerted by the ground on the car. (Magnitude and direction)


$$
F=\sqrt{5 \times 10^{6}}
$$

$=223$ ley

$$
\begin{aligned}
\text { ont } m, \quad t \quad \tan \theta & =2 \\
\theta & =63.5 \mathrm{NofW}
\end{aligned}
$$

$\mathbb{L I}$. The figure shows a box of mass $m=3$ slugs sitting on the floor of an elevator which is accelerating downward, speeding up 2 ft./second
 each second.
(a) Draw a free body diagram below showing the box and the real forces exerted on it by other objects.

(b) Determine the magnitude of each of the forces identified in part (a).
 $\sum F=m \vec{S}+m \vec{a}+\vec{N}=0$

$$
\begin{aligned}
& \mathrm{mb}=3 \text { lug }\left(32 \frac{t^{2}}{\text { Lt }^{2}}\right)=96 \text { lens down } \\
& \text { ma } \\
& N=\left|m_{\text {g }}^{\text {g }}\right|-|m a|=96-6=
\end{aligned}
$$

(c) According to Newton's third law, for every force there is an associated reaction force. What is the magnitude and direction of the reaction force associated with each force of part (a), and upon what object does the reaction force act?

III. An 8 lb. block and a 21 ib. block are tied together by a string running over a massless frictionless pulley as indicated in the figure. Assume that any additional weight added. to the $2 l$ lb. block would make the system move.
(a) Draw a free body diagram for each block, showing

(b) If the table top and the 8 lb. block are both made of oak calculate the coefficient of static friction for oak on oak.

$$
\begin{equation*}
\left.F, F_{2}=M 2\right)_{2}=(21)_{4} \tag{15}
\end{equation*}
$$

$$
\operatorname{Fen}^{2}=\operatorname{cis}^{2}
$$

$$
21 \mathrm{H}_{2}=\mathrm{H}_{2}=5 \mathrm{~S}=2
$$

(c) An additional 4 lb. block made of oak is placed on the 8 lb. block and a new block B replaces the 21 lb. block. The 41 b . block is tied to a vertical post as shown. What is the maximum weight of block B if the system is to remain at rest?



|  $p b d+0+b=d .$ pohenand wornumput <br>  <br>  $A_{p} d d==t$ <br>  <br>  <br>  <br>  $\frac{4}{8 \operatorname{tax}}=-2$ max maty <br>  <br>  <br>  <br>  <br>  $\operatorname{le}^{4}-\infty$ <br>  <br>  way may |
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4
$+1$

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| :--- | :--- | :--- |

$$
s \pm d t w 80 g
$$




$$
\therefore x^{2}+4
$$

$$
\begin{aligned}
& 8 \\
& =2 \\
& z
\end{aligned}
$$

解斯

$$
5
$$


















I. CHAPTER 30-CAPACITORS AND DIELECTRICS
$0$


BOB MARKS





$$
s>y \|^{\prime} y
$$



$$
\begin{array}{cc}
\frac{4}{4} \\
\frac{4}{4} & \frac{4}{4} \\
\frac{1}{4} & \frac{1}{4} \\
\frac{1}{4} & 4 \\
\frac{1}{4} & 4
\end{array}
$$












Physics III Test 1 October 24, 1969

Name $\qquad$ Robert Marks

BE SURE TO SHOW HOW YOU ARRIVE AT YOUR ANSWER
I.
a) In the figure below what is the magnitude and direction of $E$ at the point $x=y=a$ ?



Take $G=3 \times 10^{-10}$ caul.. $a=\sqrt{2 \mathrm{~m}} \quad \frac{1}{4 \pi \epsilon_{0}}=9 \times 10^{9} \frac{\mathrm{nt}-\mathrm{m}^{2}}{\operatorname{coul}{ }^{2}}$


$$
E=123 \frac{V}{n}=1.23 \frac{n t}{c o t}
$$

(b) What value of the negative charge will give a field of $E=0$ at $x=y=$ for positive charges as given above?

II. Two large metal plates face each other as shown in the figure. The surface charge densities are t $4.0 \times 10^{-6} \mathrm{coul} / \mathrm{m}^{2}$ and - $4.0 \times 10^{-6}$ coul/m2 on plates (1) and (2) respectively. (Circle the answer which is correct. All answers for E have the dimensions nt/coul.)
a) The value of $E$ at point $A$ is
b) The value of $E$ at point $B$ is


$$
\begin{array}{ll}
\text { 1) } & 2.26 \times 10^{-7} \\
\text { 2) } & 9.04 \times 10^{2} \\
\text { 3) } & 4.52 \times 10^{-7} \\
\text { 4) } & 1.44 \times 10^{12} \\
\text { 5) } & 0
\end{array}
$$

$$
\begin{aligned}
& \text { 1) } 2.26 \times 10^{-7} \\
& \text { (2) } 9.04 \times 10^{-7} \\
& \text { (3) } 4.52 \times 10-7 / 0^{12} \frac{1}{4 i t E_{0}}=9 \times 10^{9} \frac{n t-m 2}{\operatorname{cou} 1^{2}} \\
& \text { 4) } 1.03 \times 10^{-5 / C} \\
& \text { 5) } 0 \\
& E_{0}=8.85 \times 10^{-12} \\
& \text { could }{ }^{2} / \mathrm{nt}-\mathrm{m}^{2} \\
& \text { c) The value of } E \text { at point } C \text { is }
\end{aligned}
$$

1) $2.26 \times 10^{-7 / 0}$
2) $9.04 \times 10^{-7} \mathrm{O}^{2}$ USE BACK OF PAGE 1
3) $4.52 \times 10^{-7} 10^{2}$ FOR WORK AREA
4) $6.54 \times 10^{-5}$
(5) 0
d) The value of $E$ at point $D$ is
8. $2.26 \times 10^{-7} 10^{12} 50$
2) $9.04 \times 10^{-7} 10$
3) $4.52 \times 10^{-7} / 0$
4) $1.80 \times 10^{-6}$
III. Two charges one + and one - are located with respect to a point $P$ as indicated in fig. 1 .

a) Calculate the potential at point $P$ due to these two charges (12)


$$
V=45.8 \times 10^{3}
$$

b)

A third charge of $V=W / 4 / q$ $1 \times 10^{-6}$ coulombs is placed at $P_{\text {. This }}$ charge is now moved by some external agent from $P$ to the point $P^{1}$. At each point of its path (dotted line) it is acted on by a resultant force due to the other two charges. How much work is done by this resultant force as the $1 \times 10^{-6}$ coulomb charge is moved from $P$ to $P l$

$$
\begin{aligned}
W_{1} & =V q 0 \\
V_{1} & =\frac{\left(26 \times 10^{-6}\right)\left(9 \times 10^{3}\right)}{13} \\
& =-180 \times 10^{3} \mathrm{~V} \\
V_{P_{2}} & =\left(1 \times 10^{-6}\right)\left(9 \times 10^{3}\right) \\
& =9 \times 10^{3} V \\
V & =27 \times 10^{3} V \\
\text { in e } & =\left(27 \times 10^{3}\right)\left(10^{-3}\right)
\end{aligned}
$$

$$
\begin{align*}
& V=29 \times 10^{3} V  \tag{12}\\
& \left.W_{A B}=\frac{\left.27 \times 10^{-3}\right)\left(10^{-6}\right)}{W} \begin{array}{l}
=27 \times 10^{-3}
\end{array}\right) \text { lout }
\end{align*}
$$

$$
=1 \hat{1}^{2}
$$


c) What is the potential energy of the system consisting of the two charges shown in Fig. 1 .

$$
U=\frac{9.92}{4 \pi \varepsilon_{0} r}
$$





$\qquad$

Physics III Test 2

## In problem I circle correct answers.

I. A. The equivalent capacitance of the arrangement
of capacitors shown in the figure opposite is in microfarads.

1) 10 .
2) 11 .
3) 6.2
4) 3.875
(5) 2.506$) 4.43 \quad 71 \quad 30$


A parallel plate capacitor has a dielectric completely filing the space between the plates. The area of one surface of each plate is $0.01 \mathrm{~m}^{2}$. The dielectric constant of the dielectric is 2.0 . When a 50 volt battery is connected to the two plates, the total
free charge on each plate is found to be $30 \times 10^{-6}$ coulombs. (Note $\epsilon_{0}=8.85 \times 10^{-12} \operatorname{coul}^{2} / \mathrm{n}, \mathrm{m}^{2}$ )
B. $\therefore$ The capacitance of this parallel plate capacitor is

v) 0.6 /
(2) $1.2 \mu \mathrm{f}$
3) $150 \mu \mathrm{f}$
4) $75 \mu \mathrm{f}$
5) $300 \mu \mathrm{f}$ $3 \times 10^{-6}$
6) 16.6 ff

C. The energy stored in this capacitor is (in joules)

1) $1.5 \times 10^{-3}$
2) $7.5 \times 10^{-4}$
(3. $2.65 \times 10^{-2}$
3) $1.47 \times 10^{-15}$
4) $4.42 \times 10^{-10}$
5) $8.43 \times 10^{-2}$
6) $0.6 \times 10^{-5}$
$U=\frac{55}{2}\left(\frac{v}{9}\right)^{2}$


D Which of the following is not true.

1) The polerization $\overrightarrow{\mathrm{p}}$ is defined as the electric dipole moment per unit volume.
2) For most materials $\vec{P}$ is proportional to the electric field $\vec{E}$
3) The polarization $\overrightarrow{\mathbb{P}}$ is zero at every point in a vacuum.
4) For a given electric field, the polarization will be greater, the greater the dielectric constant of the material.
(5) The polarization vector is sometimes referred to as the electric displacement vector.

E The capacitance of the parallel plate capacitor shown opposite is

1) $\frac{K \epsilon_{0} S}{d}$
2) $\frac{\epsilon_{0} s}{K d+a+b+c}$
3) $\frac{\epsilon_{0} S}{\frac{d}{K}+a+b+c} V$
(c) $\epsilon_{0} S$
$\frac{d}{k}+a+c$
4) $\frac{\epsilon_{0} S b}{a+c+\frac{K}{d}}$
5) $\frac{\epsilon_{0} S(a+c)}{\frac{a+c}{K}+d}$
6) $\frac{K \epsilon_{0} S}{a+c+d}$

7) The power dissipated by a resistor is given by $P=i^{2} R$ or $P=V^{2} / R$. How does $P$ change if $R$ is increased or decreased? Clearly justify your answer. Explain all apparent inconsistencies or ambiguities.
8) | ARBTRAY <br> NETWORK |
| :---: |
| $+i$ | $\begin{array}{ll}t & p=i^{2} R \\ \frac{d P}{d R}=2 k R\end{array}$



$$
P \propto R \Rightarrow P=K R \Rightarrow k=i^{2}
$$

ERGO UNDER A CONSTANT CURRENT : THE POWER IS PROPORTIONAL TO THERESISTINEE IDEALLY
2)


THE POWER IN + RESISTOR IS IWVERSLY PROPORTIONAL TO THE RESISTANCE UNDER CONSeNT VOTHOE PA $\frac{1}{R}$ OR $P R=K \Rightarrow K=V^{2}$
III. For the circuit shown in the figure find the following

(8 points)
$6\left(i_{1} i_{2}+3 i_{1}+6 i_{i}+3\left(i-i_{2}\right)=V_{6}\right.$
$9\left(\lambda_{1}-i\right)+2 i=0$
$18 i=i_{1}=V_{b}$
$\left.i_{2} i_{1}\right)+3\left(i-i_{1}\right)+36=V_{b}$
$9 i_{2}-9 i_{1}=-\left(36+V_{b}\right)$$V_{a b}=$


3 equation with 3 mow indus.
solve fou $V_{b}$, then $V_{a} b$
B. If $S_{1}$ is $1 e f t$ open, and $\mathbb{R}_{1}$ is replaced by a battery with its positive terminal connected to and with an emf of 9 V , what is the power dissapation in $\mathbb{R}_{4}$ after $\mathrm{S}_{2}$ is closed?
(12 points)


$$
\begin{aligned}
& g+g i_{1}+g\left(i_{1}-i_{2}\right)=0 \\
& g\left(i i_{2}\right)=3 \\
& \left.V_{0}=-\frac{i_{1}}{1}+36\right) \text { ? } \\
& \text { sober foil }(1,-i) \text { and } i, \text { and }
\end{aligned}
$$ mutifoly by $V_{o}$

C. If switch $S_{1}$ is replaced by a capacitor $C=4 \mu f_{\text {, and the }}$ resistors $\mathbb{R}_{1}^{1}, \mathbb{R}_{2}, \mathbb{R}_{3}$, and $\mathbb{R}_{4}$ are in their original positions, what is the value of the charge $\underline{q}$ which appears on $C$ sometime after switch $\mathrm{S}_{2}$ has been closed?
(5 points)


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

$$
2 a=V_{0}(4 \mu f)
$$

$$
3 i_{1}+v_{c}+36\left(i_{1}-i_{3}\right)=0
$$

$$
6 x_{2}-V_{c}+3\left(x_{2}-x_{3}\right)=0
$$

$$
-36+86\left(i_{3}-i_{1}\right)+3\left(\hat{i}_{3} \hat{e}_{2}\right)=0
$$

$$
\text { notocepor } V_{C}
$$

$$
q_{\mu} \mu f=\frac{q}{V_{c}}
$$

IV. The internal wiring of a multirange voltmeter is shown in the figure. The galvanometer has a resistance of $1000=$ and will give a full scale deflection when a current of 1 x $10^{-3}$ amp passes through it. If the connections at A, B, and C are for full-scale voltages of lilt, lo volts, and 100 volts, respectively, what are the values for $\mathbb{R}_{1}$. Q $_{\text {, }}$ $\mathrm{R}_{2}$, and $\mathrm{R}_{3}$ ? (A full-scale voltage will give a full-scale deflection on the galvanometer.)
( 10 points)

(-)


$$
+1 v+10 v+100 v
$$

$$
R_{3}=e_{3} x^{2}=\frac{v_{3}}{i_{3}}=\frac{100}{10^{-3}}=10^{5} \Omega
$$


$\mathrm{R}_{2}=>$






## Experiment 3

Forced Oscillations and Resonance

## Purpose

To study forced oscillation and resonance of mechanical system.

## Reference

Study in $H$. ER. the linear harmonic oscillator (Sec. 15-2 and 15-3) and its rotational analog, the torsional pendulum (15-5).

Also study sec. 15-9 and 15-10. See H. \&R. Chap. 35, section 3 and questions 2, 4, for discussion of damping mechanism used in this experiment.

The oscillating system to be observed consists of a round flat disc which rotates about its axis of (cylindrical) symmetry. The "axle" of the disc moves in low-friction bearings. A spring provides an elastic restoring torque $t$ (proportional to angular displacement © from equilibrium). An electromagnet provides a damping torque proportional to the angular velocity $d \theta / d t$, providing in effect a variable amount of "friction". (Too much damping would "brake" the motion sufficiently to prevent oscillation.) You have seen references (above, text) to linear and rotational analogs. The equation of motion (15-6) of the linear oscillator, $\mathrm{d}^{2} \mathrm{x} / \mathrm{dt} 2+(\mathrm{k} / \mathrm{m}) \mathrm{x}=0$. its periodic solution ( $15-8$ ) and the period (15-10) , have their analog, equations (15-22), (15-23), and (15-10). There is a one-to-one correspondence between the sets of physical quantities:

## Iinear anquiar

| x | $\cdots$ | $\theta$ | displacement |
| :--- | :--- | :--- | :--- |
| t | $\cdots$ | t | time |
| K | $\cdots$ | K | K |
| m | $\cdots$ | B | "inestic constant" |

In our experimental study we need to carry the analogy further (Sec. 15-9 and 15-10). Thus $b$, damping const. $\longrightarrow$ B in the (electromagnetic) damping torque, ( $-\mathrm{B} d \theta / \mathrm{dt})$, and $\mathrm{F}_{\mathrm{m}} \mathrm{m}$ max. driving force $\longrightarrow \tau_{m}$. maximum driving torque, and we driving freq. $\rightarrow \omega_{d} d$, freq. of driving torque. The equation of motion (15-40) of the damped. driven linear oscillator becomes, for the damped. driven angular oscillator (our disc)

$$
\begin{equation*}
-K \theta-B d \theta / d t+T_{m} \cos \omega_{d} t=I d^{2} \theta / d t^{2} \tag{1}
\end{equation*}
$$

Eq'n. (1) asserts that the sum of the spring's restoring torque, - Ke, the damping torque, $-B$ d $\theta / d t$ and the periodic driving torque, $F_{\mathrm{m}} \cos u \mathrm{~d}$, produces an angular acceleration,. $\quad \mathrm{d}^{2} \theta / \mathrm{dt}^{2}=\ldots$, proportional to it:

$$
\tau(\text { restor, })+\tau(\text { damp. })+\tau(d r i v i n g)=I . \propto
$$

$$
I=\operatorname{constant}
$$

difference between the two periods (if any) in the direction predicted by theory: Is this difference too great to be explained by experimental error?.
II. A. With no current in the electromagnet start the disc from rest with a large initial amplitude $x_{0}$ and record peak positive displacements for the first six successive cycles. Repeat and average the corresponding peak displacements (amplitudes) A.
B. Repeat part $A$ with the terminals of the damping electromagnet attached to the 32 volt supply.
C. On a single set of axes: plot graphs of $\ln$ A versus from the data of the two preceding parts and determine the slopes at $t=0$. Use these slopes to find out the approximate relaxation times $t_{r l}$ and $t_{r 2}$

Hint: $\quad A=x_{0} e-t / t_{r}$

$$
\ln A=\left(-1 / t_{r}\right) t+\ln x_{0}
$$

and $y=\ln A$ versus $x=t$ should yield a straight line of slope ( $-1 / t_{r}$ ) if the damping torque is always proportional to the angular velocity
III. A. Plug in the motor controller and allow it to warm up. With the damping terminals connected to the 24 volt supply and the driving rod connected to the driving lever arm record the driving frequency and amplitude of oscillation for various motor speeds. Note: The apparatus must be positioned so that the pointer swings just as far in one direction as in the other. The driving rod should be connected to the lever arm in such a way as to give a large (on scale) amplitude at resonance. Observe phase relations (ques. 4 below). The frequency may be determined by timing several cycles. Be sure to cover the range of amplitudes near resonance carefully; the amplitude should be measured at several frequencies close to the resonant (maximum amplitude) frequency on each side. Before leaving this part make sure that you have enough points to plot a reasonable curve without having to guess at what happens to the curve between points. How does the resonant frequency compare with the natural frequency determined in part I?
B. Repeat Part $A$ with 48 volts on the damping electromagnet terminals. On one set of axes plot graphs of the amplitude versus the driving frequency for this and the preceding part. (See fig. $15-20$, H. + R.)

## Experiment 1 - Pendulum

## Purpose

Determine quantitative relations between parameters of oscillating physical pendulums: e.g.. size, masspmplitude and frequency.

## References

Text; and others on mechanics
Mechanics, Berkeley Physics Course, Vol. 1 , by Kittel etal
(p. 197; p.225, topic 1) for anharmonic pendulum.

With a set of pendulums formed from metal rod in the shape of isoceles triangles, use simple devices (clock, meter stick) to study their behavior.

How can you improve the precision of your measurement of period of oscillation? Partners should share in measuring techniques.

Record observations in a systematic way. Organize your work with the above purposes in mind.

Compare empirical relations - those found from your measurementswith theoretical relations - those deduced from physical laws. A table or a graph might be used. Show whether any discrepancies between theoryand experiment might be due to measurement uncertainties.

For example, suppose a calculated quantity were $I=m x^{2} / 3$, the moment of inertia of a rod about a certain axis perpendicular to it. If $m$ is measured with an experimental uncertainty $\pm \Delta m$, and $x$ with uncertainty $\pm \Delta x$, then the resulting uncertainty in $I$ is found as follows:

$$
\begin{aligned}
d \quad I & =d\left(m x^{2} / 3\right)=(m / 3) d\left(x^{2}\right)+\left(x^{2} / 3\right) d m \\
& =(2 / 3) m x d x+x^{2} d m / 3
\end{aligned}
$$

or

$$
d I / I=d I /\left(M x^{2 / 3}\right)=2 d x / x+d m / m
$$

This relates the fractional uncertainties $\Delta I / I$ etc. approximately. if they are small, or $\Delta x / x \Uparrow 1$ 。

In terms of percent uncertainty ( $\%$ error ${ }^{\text {p }}$ ) dividing through by 100 gives (\% error in $I$ ) $=2$ (\% error in $x$ ) $+(\%$ error in $m$ ).

Questions to consider in report
Could g (acceleration of "free fall") be calculated from your data. If so, indicate now. (DO it if you have time.)

Is frequency dependent on amplitude? This question is discussed in you: text for the analogous case of the simple pendulum; see also Berkeley text cited above. (The ambitious investigator might try to answer this question quantitatively.)

## I. Purpose of Laboratory

Laboratory work in physics has two important obiectives first, to give the student direct experience with some of the natural phenomena upon which physical principles are based, and second, to develop in the student some understanding of the experimental procedures. It is felt that some experience in the laboratory is necessary to give the student an insight into the methods of physics (or for that matter any experimental science). Without it he would be merely accepting principles as they were handed to him without an understanding of the experimental procedures on which they are based.

In the laboratory the student will work with real, rather than ideal, apparatus. This equipment (and the experimenter as well) will be subject to limitations which cause errors that must be taken into account before any conclusions can be drawn from the experimental results. Therefore error analysis is an essential part of all good laboratory work.

Although you will be assigned a certain group of experiments to do this quarter, and in many cases the procedure to be followed in performing the experiment is described in an instruction sheet, it is hoped that the student will use some of his own ingenuity in performing the experiments; it is intended that the instructions be used as an aid to understanding rather than something to be followed mechanically without thought. We also want to encourage students to think about possible experiments that they might do in place of one of the prescribed set. Within the limitations of equipment and time, substitution of an experiment which is more interesting to the individual student is permitted, provided it is a physics experiment and it is cleared with the instructor.

## II. Preparation for an Experiment

In order to perform an experiment thoroughly and accurately in the time allotted, it is necessary to put in some time beforehand thinking about the experiment. If an instruction sheet has been provided it is to be studied carefully before the laboratory period. You should come to the laboratory with as thorough an understanding as possible of what you are going to do during the period and why. This may require that you spend some time in the library, looking up references etcetera.

## III. Performance of the Experiment

An essential part of the method of solving an experimental problem is the preparation of a clear concise record of the data taken during the performance of the experiment. This record should contain, in a clear and legible form, all the "raw" data and information with which to make corrections (don't try to make corrections "in your head" while taking data) and also enough explanation of what you are doing and why so that your pages of
data can be analyzed later without confusion or ambiguity. Your instructor may require that this record be kept in a permanent motebook or he may ask you to keep this record on data sheets which are later included in a report on the experiment. In either case, all observations should be recorded directly into the notebook or on the data sheets (nothing on scratch paper and later copied) and an estimate of the accuracy of each set of measurements should be made and recorded also. Corrections can be made by crossing out errors with a single line (no.erasures). Before leaving the laboratory, the student should do enough calculation and graphical work to ensure that the data collected "makes sense" and there are no gaps in it which need to be filled in before he can continue the analysis without having to make any "wild guesses or assumptions. Your data record must be approved by the irfstruct or before you leave the laboratory.
IV. Laboratory Notebook (Data Record)

The following are specific suggestions concerning the form of the laboratory record of the experiments.
A. If the instructor has you keep a permanent laboratory notebook it should be one having cross-ruled pages (useful for graphs) and it must be labeled with the following information.

1. On the front cover in ink:

Physics Laboratory

Your Name
2. Inside the front couer at the top:

Fall (or whatever) quarter
Lab, day and hours
Group Number
B. For each experiment the student should record the title of the experiment and the date performed at the top of the data record. A very brief (not detailed) description of the procedure followed shouls precede the data record. which is preferably in tabular form. Label the datacarefully with the proper column headings and units. Whenever possible. the type and identifying number of instrments being calibrated or used in measurement should be recorded for later reference.
C. As suggested above the next step is to do the calculations required by the analysis of the experiment and draw the graphs. Bepeat any measurements which appear doubtful and make new measurements where needed to fill in gaps in the data.
D. If you are using a laboratory notebook rather than data sheets and if the instructor informs you that no report is required on a particular experiment, then the experiment should be completed in the notebook by writing a summary and conclusions. Final calculations should be summarized in tabular form and whatever additional graphs are required should be completed. State a conclusion in your own words and discuss the experiment briefly (for example a discussion of accuracy is always desirable). On graphs and in your final summary give the page number of the data or discussion referred to. The summary and conclusions may be left for the report when one is being written.

## V. Report

When a report is required on an experiment it is due at the beginning of the period one week after the experiment was performed. The report is to be written independently by each student in ink (or typewritten) on white, unlined 8 $1 / 2 \mathrm{x}$ ll paper (graph paper for graphs). Each report must have:
A. A cover sheet containing the following information course, experiment title, your name, laboratory period day and hours, group number, date experiment was performed, and date of report.
B. A statement of the purpose of the experiment and a brief summary of how you went about performing it (not detailed), data and observations (if you used data sheets rather than a notebook these may be submitted as they are), sample calculations, tabulated results, graphs, conclusions, and a discussion of the experiment. The discussion section of a report should be more thorough and complete than the corresponding section in the notebook. It may include a discussion of what was learned in doing the experiment, as well as the results and the accuracy of the results. It should also contain a discussion of any points which the instructor may have brought to your attention through questions written on the instruction sheets, and of any other points of interest that may occur to you.

It is customary to use the passive voice in scientific writing (e.g. "The time required for the pendulum to swing through twenty complete cycles was measured...etc.") thus not calling attention to the observer. The following styles are not to be used in a report: $I^{\prime \prime}$ (we) swung the pendulum and..." or "Swing the pendulum and measure the time for twenty complete cycles...". If you quote or paraphrase any outside sources in writing your report (including your own text book) give credit to the original author in a footnote.

## References:

1. Baird, "Experimentation" chapter 7
2. Olson. "Experiments in Modern Physics": section 1.4

Measurement, Probability, and Experimental Errors

## I. Types of Error

Whenever a measurement is made of any physical quantity there is a certain amount of uncertainty in the result. Determination of the amount of uncertainty in a measurement is not usually easy but an attempt should always be made to do so, even if it is no more than an educated guess. Without some estimate of the uncertainties associated with experimental measurements one has no indication of the accuracy of the results and it is difficult to come to any conclusion about what the experiment has shown (or not shown). In all of the experiments which follow in the physics laboratory sequence the student will be expected to make some estimate of the accuracy of his quantitative experimental results.

There are two types of errors which may occur in the measurement process, systematic errors and random errors. Systematic errors tend to make all the observations of one item too small or too large. For example if voltage measurements were taken in an electric circuit using a voltmeter which consistantly read 0.1 volt too high, a systematic error would be present. Other common examples of causes of systematic error are worn weights, clocks which gain or lose time, friction, and personal bias of the observer which causes him to make readings which are consistently high or low. When systematic errors are recognized in an experiment it is often possible to find out how large their effect is and to correct for it. The error in the voltmeter which reads 0.1 volt too high, for example, can be discovered by calibrating the instrument against some sort of standard (accurately known voltage), and a correction of -0.1 volt made to all the readings. Error due to an observer ${ }^{\text {a }}$ bias may be minimized by having another observer make the same measurement independently (bias is best eliminated if each observer knows nothing of the other's result until after both measurements are completed).

Randomerrors result from chance variations in the quantity being measured, in the measuring devices; or in the observer, and are just as likely to produce too large a value as too small. For example, if one measures the diameter of a metal rod several times with a micrometer the readings will probably fluctuate slightly in a non-systematic fashion due to actual differences in the rod's diameter at different positions, variations in pressure when the micrometers jaws are closed, and changes in the observer s estimate of the scale reading. Random errors are present in all measurements: although they may be too small to be noticeable, and they cannot be corrected for because of their random nature.

Determination of Precision
Suppose that several measurements of the same quantity $x$ were made and all systematic error in the measurements eliminated or corrected (assuming this were possible). As discussed above there would still be a certain amount of rnadom fluctuation apparent in the measurements if they are "fime" enough to make it noticeable. If a histogram was plotted showing the number of measurements $N$ faling within different intervals of size $\Delta x$ it might look like that shown in Fig. 1 .

Fig. 1
No. of Measuremerts vs.
Measured Valves $x$


The meaning of the histogram is that one measurement of $x$ fell between 4.6 and 4.7 units, two between 4.7 and 4.8 units, four between 4.8 and 4.9 units, and so forth. The completely symetrical distribution shown usually results only if a large number of measurements are made and if the fluctuations are entirely random. In such cases the envelope of the distribution often has a particular form called a "mormal" or "Gassian" distribution which is represemted by the mathematical equation

$$
\begin{equation*}
y=\frac{1}{\sqrt{2 \pi} \sigma} e^{-(x-x)^{2} / 2 \sigma^{2}} \tag{1}
\end{equation*}
$$

where $\sigma$ is a constant which determines the "sharpness" of the peak (high narrow peaks are characterized by small values of $\sigma$ ). The quantity $\bar{x}$ is the average of the individual
measurements

where $n$ is the total number of measurements and because of the symmetry of the Gassian function $\bar{x}$ corresponds to the most probable value of $x$ obtained from a measurement of $x$ (peak of curve). Thus $\bar{x}$ is the best estimate that one may make of the irue value of $x$ from these measurements.

The individual measurements of $x$ differ from the average or most probable value $\bar{x}$ by an amount dealled the deviation of that measurement

$$
d_{1}=x_{1}-\bar{x}, d_{2}=x_{2}-\bar{x}, \ldots \ldots \ldots
$$

The standard deviation

$$
\sigma=\left[\frac{d_{1}^{2}+d_{2}^{2}+\ldots \ldots d_{n}^{2}}{n \cdot 1}\right]^{1 / 2}=\left[\frac{\Sigma\left(d_{i}\right)^{2}}{n-1}\right]^{1 / 2}
$$

is an indication of the precision of a set of measurements since narrow Gaussian distributions indicate precise measurements with small deviations from the average and a small standard deviation $\sigma$. If large number of measurements is made, $68 \%$ of them will be in the range $\bar{x}+\sigma, 95 \%$ in the range $\bar{x} \pm 2 \sigma$, and $99 \%$ in the range $\bar{x} \pm 3 \sigma_{\mathrm{i}}$ a fact which can be verified by determining the area under a Gaussian curve between the various limits. If after having determined $x$ and ofrom a large mumber of measurements one makes a single measurement $x_{0}$ he then will have about a two thirds chance of getting a value between $\bar{x}+\sigma$ and $\bar{x}-\sigma$ e etcetera.

Although increasing the number of measurements of quantity $x$ would have little effect on the standard deviation $\sigma$ (the scatter of the datal except to give a more accurate picture of what it really is increasing the number of measurements should improve the reliability of the average value $\bar{x}$. It can be shown from statistics that the standard deviation in the mean $\bar{x}$ is given by the equation

which means that there is a $68 \%$ chance that the true value of $x$ will be in range $\bar{x} \pm \sigma$ assuming the distribution is mormal and there are mo systematic errors present. Thus the precision of the mean $\bar{x}$ can be increased ( $\sigma$ peduced) by taking more observations. but the improvement is slow because of the $\sqrt{n}$ factor ( 90 readings only 3 times as good as 10 readings). The final result of a set of measurements may be stated

$$
x=\bar{x} \pm \sigma m
$$

It is quite often useful to represent the standard deviation $\sigma^{m}$ as a percentage of the value $\bar{x}$. The calculation required is:

$$
\text { per cent std. dev. }=(0 \mathrm{~m} / \mathrm{x}) \cdot(100 \%)
$$

Although the normal or Gaussian distribution (equation 1) is very often a good representation of the kind of distribution found in repeated measurements of physical quantities, it should not be assumed that this distribution always gives an accurate description of the results of such measurements. even when a large number of measurements are made. There are a number of cases where the distribution is mon-Gaussian and perhaps even nomasymetrical. For example, if one makes several determinations of the number of nuclei which decay by particle emission in a certain time, he obtains the Poisson distribution

$$
\begin{equation*}
y \propto \frac{\bar{x}^{x}}{x!} e^{-\bar{x}} \tag{2}
\end{equation*}
$$

where $\bar{x}$ is the average number of counts and $y$ is the probability of obtaining $x$ counts in given trial. This distribution is very unsymmetrical about the mean $\bar{x}$ when the number of counts $\bar{x}$ is small but closely resembles a Gaussian distribution with standard deviatiom $\sqrt{\bar{x}}$ when $\bar{x}$ is large.
III. Propagation of Errors If one uses experimental observations, with their associated random errors, to calculate a result, the precision of the result will be determined by the precision of the quantities involved in the calculation. The standard deviation of the result may be determined from those of the separate quantities $\sigma_{m}$. $\sigma$ mq, etc. by keeping in mind the following rules.
A.

```
The standard deviation of the result of addition and/or substraction is the square root of the sum of the squares of the standard deviations of the separate terms.
```

Example: $\quad x_{1}=5.30 \pm 0.20$ units

$$
x_{2}=1.70 \pm 0.10 \text { units }
$$

$$
x_{3}=7.20 \pm 0.01 \text { units }
$$

$$
\begin{gathered}
x_{1}-x_{2}+x_{3}=(5.30-1.70+7.20) \pm\left[(0.20)^{2}+(0.10)^{2}+(0.01)^{2}\right]^{1 / 2} \\
=10.80 \pm 0.22 \text { units }
\end{gathered}
$$

Note that most of the standard deviation in the result comes from the largest standard deviation present in the separate terms ( $0.22 \simeq 0.20$ ) 。
B. The percentage standard deviation in the result of mulitplication and/or division is the square root of the sum of the squares of the percentage std. deviations of the factors.
example: $x_{1}, x_{2}, x_{3}$ as above
$(\% \text { std. dev. })_{1}=\frac{0.20}{5.30} \times 100 \%=3.8 \%$
(\% std. dev.) ${ }_{2}=\frac{0.10}{1.70} \times 100 \%=5.9 \%$
$(\% \text { std. dev. })_{3}=\frac{0.01}{7.20} \times 100 \%=0.1 \%$
$y=\frac{\left(x_{1}\right)\left(x_{2}\right)}{x_{3}}=1.25 \pm$ std. dev.
$(\% \text { std. dev. })_{y}=\left[(3.8)^{2}+(5.9)^{2}+(0.1)^{2}\right]^{\frac{1}{2}}=7.0 \%$
$(\text { std. dev. })_{y}=(.07)(1.25)=0.09$
$y=1.25 \pm 0.09$ units
Note that in this case the largest contribution to the standard deviation in the result comes from that quantity with the largest percentage standard deviation.
C. In case a quantity is raised to the $n$th power its percentage standard deviation is multiplied by $n$.

The process of carrying standard deviations through calculations is useful not only indetermining the precision of the result but also in determining which quantity contributes most to random error in the result. It may be possible to reduce the deviations in this quantity by using more care or different techniques.

## IV. Accuracy of Experimental Results

Determination of the standard deviation in an experimental result will tell you how much uncertainty is present due to random errors, but this is an indication of the ascuracy of the rocult only in the case where systematic errors are nogligible compared to random errors. For example, if in a particular experiment you obtained a percentage standard deviation of $1 \%$ but the instruments used to obtain the measurements were accurate only
to within $5 \%$ (all readings may be too high or low by $5 \%$ ) then the $5 \%$ accuracy is a better indication of the reliability of the results than the $1 \%$. Some attempt should be made by the student to determine the reliability of his results in each experiment. although in some cases this will involve making some educated guesses as to the accuracy with which a particular measuremant may be made with a particular measuring device. In all cases try to eliminate as much systematic error from the measurement as possible within the time available. An experimental result does not agree with a prediction of a theory unless the theoretically predicted result dies within the range given by the experi= mental result plus and minus the probable error: an experiment does not disagree with a theory unless the predicted result lies outside this range.

## V. Significant Figures

The term "significant figures" refers to the digits of a measurement made in the laboratory, including all the certain digits and one additional doubtful one based on the observer s estimate of a fraction of acale division. The numbers which represent data or the results of calculations should always be given with neither more nor fewer significant figures than are justified by the precision of the observations and computations. The number of significant figures in a measurement (or a calcues: lated quantity) may be determined using the following ruies.
(a) The first significant figure is the first non-zero digit.
(b) Zeros which occur between significant digits are considered significant.
(c) Zeros which occur to the right of the last non-zero digit are considered significant when they are to the right of the decimal point (the significance of such zeros to the left of the decimal point is indeterminate).
(d) If numbers having a different number of significant figures are added. substracted, multiplied or divided. the answer is given so as to have the same number of significant figures as the term or factor which has the least.

Examples: . 0001906 has 4 significant figures
10.937 has 5

93,000 has an indeterminate number
$9.3 \times 10^{4}$ has 2
$9.30 \times 10^{4}$ has 3

## VI. Comparison of Results

Sometimes an experimental result is arrived at by two different methods which should both theoretically give the correct result. If there is no reason to believe that one of the results is much more accurate than the other, it might be instructive to see how much difference there is between the two. This difference is usually given in terms of the "percentage difference" which is defined.

$$
\% \text { diff. }=\frac{\text { diff between values }}{\text { average value }}: 100 \%
$$

## References:

1. Young. "Statistical Treatment of Experimental Data: ${ }^{\text {at }}$
2. Barford "Experimental Measurements: Precision. Error and Trutho
3. Bairds "Experimentation: An Introduction to Measurement Theory and Experiment Design"
4. Braddick, "The Physics of Experimental Method"
5. Pugh and Winslow, "The Analysis of Physical Measurements ${ }^{3}$
6. Bevington, "Data Reduction and Error Analysis for the Physical Sciences ${ }^{\text {P }}$

## METHOD OF LEAST SQUARES

One of the fundamental problems that comes up again and again in the laboratory is that of finding, from simultaneous measurements of quantities $y$ and $x$, the dependence of quantity $y$ on quantity $x$ (the dependence of the period of a pendulum on its length for example). Often this dependence is revealed by making a graph of y versus $x$ from the data. However a certain amount of judgement is always involved in making a graph from experimental data since deviations in the measurements usually make it impossible to draw a smooth curve through all the data points. One usually tries to draw a smooth curve among the points in such a way that it appears that the deviations of the points from the line (positive and negative) add up to approximately zera. In other words, in the graph shown below

$$
\left|d_{1}\right|+\left|d_{3}\right|+\left|d_{4}\right|+\ldots \ldots \ldots \approx\left|d_{2}\right|+\left|d_{5}\right|+\ldots \ldots \ldots .
$$

where the deviations here and in the analysis to follow will be assumed to be deviations in $y$ for precisely known values of $x$.


If a high degree of precision is required in the expression relating $y$ to $x$, this method of balancing deviations "by eye" might not be sufficient. In this case a more scientific approach, based on statistics, is followed. It can be shown that the most probable disposition of the line representing the dependence of $y$ on $x$ is that for which the sum of the squares of the deviations of the points from the line is a minimum (hence the name "least squares"

$$
\Sigma\left(d_{i}\right)^{2}=d_{1}^{2}+d_{2}^{2}+d_{3}^{2}+d_{4}^{2}+\ldots . .=\text { a minimum }
$$

This statement is called the "principle of least squares" and it is the basis of a method for finding the relationship between $y$ and $x$ which best fits the data points (for which the sum of the squares of the deviations is a minimum).

Actually the problem of determining the line which "best" fits a set of data points ( $x_{i}, y_{i}$ ) is several different problems. depending on the type of curve which is to represent the relationship between $x$ and $\bar{Y}$. If it has been predetermined from the data or from theory that $y$ depends on $x$ linearly so that $y=A x+B$; the problem becomes one of picking out, from all possible straight lines, the one with values of slope $A$ and intercept B such that the sum of the $d_{i}{ }^{2}$ will be as small as possible. If ( $x_{1}, y_{1}$ ) are the coordinates of the first data point, ( $x 2, y_{2}$ ) the coordinates of the second and so forth, and if it is assumed that the deviations are only in the $y$ measurement for precisely known $x{ }^{\circ} s_{8}$ then

$$
\Sigma\left(d_{i}\right)^{2}=\left(A x_{1}+B-y_{1}\right)^{2}+\left(A x_{2}+B-y_{2}\right)^{2}+\ldots .
$$

If the "best" straight line is that which makes the sum of the squared deviations or a minimum.

$$
\begin{aligned}
& \frac{d\left[\Sigma(d i)^{2}\right]}{d A}=0=2 x_{1}\left(A x_{1}+B-y_{1}\right)+2 x_{2}\left(A x_{2}+B-y_{2}\right)+\ldots \\
& \frac{d[\Sigma(d i) 2]}{d B}=0=2\left(A x_{1}+B-y_{1}\right)+2\left(A x_{2}+B-y_{2}\right)+\ldots \ldots
\end{aligned}
$$

are the conditions which should lend to the "best" values of $A$ and B. These equations may be rewritten:

$$
\begin{align*}
B \Sigma x_{i}+A \Sigma x_{i}{ }^{2}-\sum x_{i} y_{i} & =0  \tag{1}\\
n B & +A \Sigma x_{i}-\Sigma y_{i}=0 \tag{2}
\end{align*}
$$

where $n$ is the number of points.
The method is illustrated below for a set of $\mathfrak{n}=5$ points.

| Point <br> No. | 1 | 2 | 3 | 4 | 5 |
| :---: | :---: | :---: | :---: | :---: | :---: |
| x | 1.00 | 1.90 | 2.60 | 3.20 | 4.00 |
| y | 0.90 | 3.00 | 4.00 | 5.50 | 6.90 |

$$
--3--
$$

A table is made as follows:

| $x$ | $y$ | $x^{2}$ | $x y$ |
| :--- | :--- | :--- | :--- |
| 1.00 | 0.90 | 1.00 | 0.90 |
| 1.90 | 3.00 | 3.61 | 5.70 |
| 2.60 | 4.00 | 6.76 | 10.40 |
| 3.20 | 5.50 | 10.24 | 17.60 |
| 4.00 | 6.90 | 16.00 | 27.60 |

$\Sigma x_{i}=12.70 \quad \Sigma y_{i}=20.30 \quad \Sigma x_{i}{ }^{2}=37.61 \quad \Sigma x_{i} y_{i}=62.20$

Substituting in (1) and (2).

$$
\begin{aligned}
12.70 B+37.61 A & =62.60 \\
5 B+12.70 A & =20.30
\end{aligned}
$$

Solving simultaneously,
$B=-0.989$
$A=1.988$
The equation of the straight line which best fits the data points is

$$
y=1.988 \quad x-0.989
$$

In other words the sum of the squares of the deviations of the points from the straight line is a minimum for a line of slope 1.988 and $y$ intercept -0.989 .

It is generally shown in books on statistics that the standard deviations in these values obtained for the slope a and intercept $B$ may be found using the equations (3 and 4):

$$
\begin{aligned}
& \sigma_{A}=\left[\frac{\sum_{i} d_{i}}{n \sum x_{i}^{2}-\left(\Sigma x_{i}\right)^{2}}\right]^{1 / 2}=\left[\frac{\sum\left(A x_{i}+B-d_{i}\right)^{2}}{n \Sigma x_{i}^{2}-\left(\Sigma x_{i}\right)}\right]^{1 / 2} \\
& \sigma_{B}=\left\{\frac{\left(\Sigma d_{i} 2\right)\left(\sum_{i} 2\right)}{n^{2} \sum x_{i}^{2}-n\left(\Sigma x_{i}\right) 2}\right\}^{1 / 2}=\left\{\frac{\left[\Sigma\left(A_{x_{i}}+B-y_{i}\right)^{2}\right]\left[\Sigma x_{i} 2\right]}{n^{2} \sum x_{i}^{2}-n\left(\sum x_{i}\right)^{2}}\right\}^{1 / 2}
\end{aligned}
$$

A table is made as follows:

| $x$ | $y$ | $x^{2}$ | $x y$ |
| :--- | :--- | :--- | :--- |
| 1.00 | 0.90 | 1.00 | 0.90 |
| 1.90 | 3.00 | 3.61 | 5.70 |
| 2.60 | 4.00 | 6.76 | 10.40 |
| 3.20 | 5.50 | 10.24 | 17.60 |
| 4.00 | 6.90 | 16.00 | 27.60 |

$$
\Sigma x_{i}=12.70 \quad \Sigma y_{i}=20.30 \quad \Sigma x_{i}{ }^{2}=37.61 \quad \Sigma x_{i} y_{i}=62.20
$$

Substituting in (1) and (2).

$$
\begin{aligned}
12.70 \mathrm{~B}+37.61 \mathrm{~A} & =62.60 \\
5 \mathrm{~B}+12.70 \mathrm{~A} & =20.30
\end{aligned}
$$

Solving simultaneously,

$$
B=-0.989 \quad A=1.988
$$

The equation of the straight line which best fits the data points is

$$
y=1.988 \quad x-0.989
$$

In other words the sum of the squares of the deviations of the points from the straight line is a minimum for a line of slope 1.988 and $y$ intercept -0.989 .

It is generally shown in books on statistics that the standard deviations in these values obtained for the slope $A$ and intercept $B$ may be found using the equations (3 and 4):

$$
\begin{aligned}
& \left.\sigma_{A}=\left[\frac{\sum_{i} 2}{n \sum x_{i}^{2}-\left(\Sigma x_{i}\right)^{2}}\right]^{1 / 2}=\left[\frac{\sum\left(A x_{i}+B-d_{i}\right)^{2}}{n \sum x_{i}^{2}-\left(\sum x_{i}\right.}\right]^{2}\right]^{1 / 2} \\
& \sigma_{B}=\left\{\frac{\left(\sum \sum_{i} 2\right)\left(\Sigma x_{i} 2\right)}{n^{2} \sum x_{i}^{2}-n\left(\sum x_{i}\right) 2}\right\}^{1 / 2}=\left\{\frac{\left[\sum\left(A_{x_{i}}+B-y_{i}\right)^{2}\right]\left[x_{i} 2\right]}{n^{2} \sum_{i}^{2}-n\left(x_{i} x_{i}\right)^{2}}\right\}^{1 / 2}
\end{aligned}
$$

In cases where a nonlinear curve is to be fit to a set of data points in such a way as to make $\bar{\Sigma}(d i)^{2}$ a minimum, equations (1): (2): (3), and (4) no longer apply, Often one can get around this difficulty, however. For example, suppose some data points are to be fit with a parabola of the type $y=A x^{2}+B$. If the quantity $X=x^{2}$ is calculatedfor each of the points, the method may then be applied to quantities $y$ and $X$ since $y$ versus $X$ will be a straight line $(y=A X+B)$ even though y versus $x$ is not.

The least squares method is not confined to finding the constants of a straight line however; it can be applied to any kind of curve. For example if one has a set of data points and wants to determine the constants of the "best fit" parabola $y=A X^{2}+B X+C$ he can apply the conditions that minimize $\Sigma(d i)^{2}$ with respect to variables $A_{f} B_{s}$ and $C$ and will obtain the equations:

$$
\begin{array}{ll}
\Sigma x_{i}^{2} y_{i} & =C \Sigma x_{i} 2+B \sum x_{i}^{3}+A \sum x_{i}^{4} \\
\Sigma x_{i} y_{i} & =C \Sigma x_{i}+B \sum x_{i}{ }^{2}+A \Sigma x_{i}^{3} \\
\Sigma y_{i} & =n C+B \Sigma x_{i}+A \Sigma x_{i}{ }^{2}
\end{array}
$$

which may be solved simultareonsly for constants $A$. $B$; and $C$.

## References:

1. Young. "Statistical Treatment of Experimental Data". section 14.
2. Baird. "Experimentation". Appendix 2
3. Barford, "Experimental Measurements" Chapter 3
4. Pugh and Winslow, "The Analysis of Physical Measurements". Chapter 10 .
5. Bevington. "Data Reduction". Chapters 6 and 11
6. Gerhold. "Least-Squares Adjustment of Weighted Data to a General Linear Equation". American Journal of Physics Vol. 37 g p. 156 .


## Interference and Diffraction

According to Huygens's principle, each point along a wavefront may be regarded as a new source of waves. Whenever something obstructs part of the wavefronts, interference between "wavelets" emanating from different parts of the unobstructed wavefronts produce a diffraction pattern which is characteristic of the geometry of the obstruction (or opening in object which blocks the light) and of the wavelength of the light. It is shown in nearly all introductory physics textbooks, for example (see Resnick and Halliday, section 43-1), that when light waves pass through a double silt arrangement like that shown below they interfere constructively and destructively at different positions to form fringes on the screen $S$ such that intensity maximum appear at positions

$$
x=n\left(\frac{D A}{d}\right) \quad n=0,1,2,3 \ldots \ldots .
$$



In somewhat the same way wavelets passing through different parts of a single slit interfere to produce a single slit diffraction pattern with destructive interference causing diffraction minima at angles $\Theta$ such that (see Resnick and Holiday, section 44-2)

$$
a \sin \theta=m \quad m=1,2,3, \ldots \ldots
$$

with maxima approximately half way between (the exact intensity expressions are given in section 44-3), where a is the slit width. A circular aperture of diameter d results in fringes having circular symmetry with the first minimum appearing at a distance from the center such that (see Resnick and Halliday, section 44-5)

$$
\sin \theta=1.22 \lambda / d
$$

## Experiment:

Your light source will be a monochromatic beam from a heliumneon laser having wavelength $\lambda=6328$ 盲. This experiment is somewhat open ended in that you are not told exactly what to do or how to do it. The object is to investigate the nature of interference and diffraction effects. The details are left to you. You might for example put a single or double slit in the beam and determine the spacing or width of the slits, perhaps checking your results with a direct measurement using the optical comparator. You might try to do an experiment which would confirm the constant 1.22 in the expression for the first fringe minimum from a circular aperture or compare the pattern
fringe/of aperture or slits of different sizes. An experimental study of diffraction by rectangular openings or by a repeated pattern of openings (such as found in a seive) might be of interest.

$$
\begin{aligned}
& \begin{array}{r}
\text { SINGLE SLIT } \\
0=54 \mathrm{~cm} \\
x, 9.9 \mathrm{em} \\
x .=1.65 \mathrm{~cm}
\end{array} \\
& \text { H } \\
& 4 \\
& \text { an }
\end{aligned}
$$

$$
\begin{aligned}
& \text { 1IT) PIN MOLE } \\
& \text { + }-23 \mathrm{~cm}
\end{aligned}
$$

$$
b \geq b 0<175
$$

$$
=78000
$$

$$
\begin{array}{r}
C O Q Q O U T V O L \\
C=\frac{n Q}{x}
\end{array}
$$

$$
-2=2+1+1
$$

$$
=207
$$

$$
\frac{(10 y E}{\left.3 \times 10^{-2}\right)}
$$

$$
\left.\frac{3(2-45)(6,33 \times 10}{2+12}\right)
$$

$$
d=\frac{d_{1}+a_{2}+d_{2}+d a+c}{5}
$$

$$
175-75 N 15=0
$$

$$
\frac{x}{2}
$$

$$
2 A x+A(2)
$$

$$
4
$$

$$
\frac{c-\infty x=1}{(2-0+x)(1+2)}
$$

$$
(2)(51)\left(6-33 \times 10^{-7}\right)
$$

$$
\operatorname{cog}^{6}=-(x+1+2)(4)
$$

$$
\frac{(6-3,3 x+0-7)(54)}{(4)}
$$

$$
\frac{1.22 \times 0}{(1 \times 2)\left(6.33 \times\left(0^{-7}\right)(1.04)\right.}
$$

$$
\frac{(1.22)\left(6.33 \times 10^{-7}\right)(1.04)}{23 \times 10^{-2}}
$$

$$
y 0(\square)
$$

$$
=10(8)
$$

$$
=1 p \pi
$$

$$
-\infty
$$

$$
\frac{4}{4}
$$

$$
7
$$

$$
C_{2}=
$$

CIRGULAR APERTUQE

## Conchuslons ano onservations:


FAET THAT LIGHT OISPGAYS
PROPERTIES HAS CLEARLY Hot amathema
SUPER POSITION
N
$\frac{1}{0}$
$\frac{1}{6}$
OF
$1 N$


## 7.

4. The method of least squares determines the most probable line of a certain type (e.g. straight line) through a set of data points.
(a) The line is placed so as to minimize $\qquad$
$\qquad$
$\qquad$
.
(b) Use the method of least squares to determine the slope and $y$ intercept of the straight line graph $I$ versus $V^{2}$ in question 3. How do your results compare with the values taken directly from the graph in question 3 ?

## GRAPHICAL ANALYSIS

Often one of the aims of an experimental investigation is the determination, from measurements made in the laboratory, of how one of two interdependent quantities; $y$, depends on the other, $x$. Graphical methods provide us with a very usefultool in this type of analysis.

## I. Plotting Graphs

Suppose one is interested; for example, in finding in a particular experiment a methemical relationship which expresses the velocity of a moving object $v$ as a function of the time $t$. In this case velocity is the "dependent variable" whose dependency on the "independent variable" time is to be established from the following data.

Time
Velocity (magnitude)

| Time | Velocity (magnitude) |
| :---: | :---: |
| 1.00 | $(\mathrm{~cm} / \mathrm{sec})$ |
| 2.00 | 1.9 |
| 3.00 | 1.9 |
| 4.00 | 3.0 |
| 5.00 | 3.9 |
| 6.00 | 6.5 |

Suppose that in this experiment the time measurements are very precise and their errors can be ignored while the velocity measurements are estimated to have a standard deviation (see instruction sheet on "Measurement, Probability, and Experimental Errors ) of about $\pm 0.30 \mathrm{~cm} / \mathrm{sec}$. The steps to be followed in constructing a graph which illustrates the dependence of velocity $v$ on time $t$ (or any quantity $y$ on another quantity $x$ ) are summarized below.
(a) The dependent variable (quantity whose dependency on the other is to be determined) is plotted vertically (velocity versus time rather than vice versa).
(b) Scales should be chosen which are easy to plot and easy to read and which make the graph large enough to be read easily and accurately (occupying a full page if possible):
(c) Scales usually start at zero but sometimes this would cause the data to be crowded into one part of the graph. In such a case it is a good idea to suppress the zero (start the scale at some value other than zero or show a break in the scale). However, it should be made obvious to someone looking at the graph that the zero has been suppressed.
(d) The graph should have a title and each of the axes should show the quantity plotted along that axis and the numerical scale and units for that quantity.
(e) The experimental points are marked clearly on the graph by drawing a small circle around each of them and drawing an "error line" (in the above example extending $0.30 \mathrm{~cm} / \mathrm{sec}$ above and below the data point) at each point.
(f) Draw the simplest possible smooth line or curve (i.e. the simplest curve is a straight line, the next is a curve whose curvature is always in the same direction and doesn't change magnitude suddenly, etc) among the points, with no more details of shape and curvature than is justified by the size of the estimated errors. If the magnitude of the standard deviations are estimated correctly and the line is drawn correctly the curve should cut about two thirds of the error lines (very roughly).

When these steps are applied to the example of the moving object given above, a graph results such as that shown in the following figure.
Velocity $v \quad$ versus Time $t$


If a graph of dependent variable $y$ versus independent variable $x$ turns out to be a straight line, the dependence of $y$ on $x$ is expressed by the equation

$$
\begin{equation*}
y=a x+b \tag{1}
\end{equation*}
$$

The slope a and $y$ intercept $b$ of the line can be taken directiy from the graph (see part III) thus establishing the relationship between quantity $y$ and quantity $x$ in this experiment.

If the graph of $y$ versus $x$ is curved, however, as it is in the case of the velocity of an object versus the time in part $I_{\text {, }}$ the quantities must be related by some other equation. For example, one might guess that $y$ is related to $x$ according to an equation of the type

$$
\begin{equation*}
y=a x^{n}+b \tag{2}
\end{equation*}
$$

where $n$ might be an integer $-1, \pm 2, \pm 3, \pm 4, \ldots \ldots \ldots$........... ${ }^{2}$ a fraction $\pm 1 / 2, \pm 1 / 3, \pm 1 / 4, \ldots \ldots . . . . . .$. values of $n$ are truly possibilities one should study the graph of $y$ versus $x$ and equation (2). In the case of the velocity versus time graph of part $I$, for example, negative values of $n$ should be immediately discounted since equation (2) would predict a decrease in $y$ for increasing $x$. Fractional values of $n$ are just as unlikely since as $x$ increases, the graph shows $y$ increasing faster and faster (perhaps indicating $n=+2$ or +3 , etc.).

To see if the velocity - time ( $y=v, x=t)$ data for the moving object example of part $\mathbb{I}$ fits equation (2) with $n=+2$ one could graph $Y=v$ versus $X=t^{2}$ from the experimental values of $v$ and the corresponding values of $t^{2}$. If the graph of $Y$ versus $X$ from the data is a straight line, the experimental results fit a relationship
or

$$
Y=a X+b
$$

$$
v=a t^{2}+b \quad \text { (equation } 2 \text { with } n=+2 \text { ) }
$$

where a and b are the slope and intercept of the line. If such a graph was not straight, but was straighter than a graph of versus $t$, then one might try a graph of $Y=v$ versus $X=t 3$ and so on until a straight line was found. The same general procedure could be followed in cases where $n$ is thought to be a fraction or have a negative value. If the data are to be represented by the equation

$$
\begin{equation*}
y=a x^{-1 / 3}+b \tag{3}
\end{equation*}
$$

then a graph of $y$ versus $x^{-1 / 3}$ should yield a straight line.

Another type of relationship between quantities which appears often is

$$
\begin{equation*}
y=A e^{a x} \tag{4}
\end{equation*}
$$

where $A$ and a are positive or negative constants. If equation (4) accurately represents the data, then
$\ln y=a x+\ln A$
or $\quad Y=a x+b$
making the substitutions $Y=\operatorname{Ln} y$ and $b=\ln A$. Therefore if $Y=\ln y$ is plotted vertically against $x$ horizontally, a straight line of slope a and intercept $b=\ln A$ should result. The values of a and $A$ can be determined from this line.
III. Determination of Slope and Intercept

The slope and intercept of a straight line are found as follows: First the $x$ and $y$ coordinates of two widely separated points on the line are determined (note that the points must be widely separated for accuracy and the points are points on the line, not data points). The slope of the line is defined

$$
a=\frac{y_{2}-y_{1}}{x_{2}-x_{1}}
$$

and should have the same value (for a straight line) regardless of what two points are chosen. The y intercept is obtained by extending the line back to $x=0$ and noting the value of $y$ at this point on the line (this is the intercept b).

A more reliable determination of slope, $a_{\text {, }}$ and $y$ intercept b, results when one computes the slope and intercept of the straight line which minimizes the sum of the squares of the deviations of the data points from the line (see instruction sheet on "Method of Least Squares").

## References:

1. Kruglak and Moore, "Basic Mathematics for the Physical Sciences", chapter 7.
2. G. Wootan, Inc.: "Graphs"
3. Ford, "Basic Physics", section 7.6

Physics III Lab.
Fall 1969

Experiment $2{ }^{\text {'Mapping' }}$ an Electric Field in Two Dimensions

## Purpose

To provide a look at equipotential lines in a two dimensional system. Hopefully an electric field may become more real (less abstract: to you.

The 2-D system is a thin layer of slightly-conducting graphite on the surface of a board. The equipotenital lines (each, the locus of points having a given constant potential) correspond to equipotential surfaces in a three dimensional system. You know where to look these up - refresh your memory!

When two points in a conducting body or circuit are at different potentials a current flows between the points. You will use a galvanometer $G$ to determine a condition of zero-current flowi i.e. the condition of zero-potential difference between the points to which the galvanometer is connected. If you have not learned the principles of this instrument, you need only assume that zero-deflection means zero-current in the galvanometer.

Map at least two field-electrode configurations (at least one for the report of each student). Tentatively sketch some field lines ("lines of force").

Take care: some galvanometers have two button switches to connect them: Use first until the deflection is very small, then use the more sensitive 0-button.

Discuss in report the relation between field lines and equipotentials, and the theory behind this relation. Indicate the direction of the E-vector on your plot.

Explain the analogy between your equipotential plot and another kind of map. (Incidentally, can the energy concept be included in your explanation?)

$$
1.5 \mathrm{v}
$$



POTENTIOMETER MEASUREMENT OF POTENTIAL DIFGEFENCE

OBJECT: To calibrate a potentiometer and to become familiar with its use in measuring potential differences.

THEORY: One of the most obvious advantages of using ammeters, voltmeters, etc. in making circuit measurements is their convenience; they can be easily moved about from one part of the circuit to another. For this reason and also because their accuracy is sufficient for the purpose at hand in many instances, they are widely used in measuring the currents, voltages, etc. in a variety of circuit applications. However, it should be remembered that there are different types or voitmeters, for example, having different characteristics and for use under different conditions. One should not assume that all voltmeters, if operating correctly, will automatically read correctly the voltage he wishes to measure in any circuit.

One cause of incorrect meter readings is of course that the instrument is not properly calibrated. This can be quickly remedied by recalibrating it against a standard and making a calibrating graph which shows the correct value for any meter reading.

Often, however, the difficulty is not in the calibration but in the use that is made of a meter. Suppose in the circuit below for example, that you wish to measure the voltage across the 200 ohm resistor with a voltmeter having only 200 ohms resistance.


An ideal voltneter would be one with infinite resistance; if it were placed in parallel with $R$ it would draw no current and the total resistance of the parallel combination would remain 200 ohms (there would be no change in the operation of the circuit due to the introduction of the voltmeter). In the case of a 200 ohm voltmeter, however, the introduction of the meter causes the resistance of the parallel combination to be cut in half and the total current to increase, half going through $R$ and the other half through the voltmeter. The meter will correctiy read the voltage across its terminals, but that voltage is no longer the same as it was when the meter was not present. This difficulty car usually be avoided by taking care that the resistance of the voltmeter be large compared to the total resistance of the circuit between the points to which the meter is connected.

The potentiometer is in effect an "ideal" voltmeter. It draws no current fron the circuit at the instant of measurement and thus doesn't change conditions in the circuit from what they were before its introduction.

The potentiometer circuit is shown on the next page. The power source supplies direct current through a variable resistance to a uniform resistance siide wire CD. The variable resistance may be increased or decreased in order to control the amount of current I flowing in the wire. Since the slide wire is uniform along its length, its resistance per unit length is the same everywhere so that

$$
\begin{equation*}
\frac{V_{T D}}{V_{C D}}=\frac{I R_{T D}}{I R_{C D}}=\frac{I_{T D}}{I_{C D}} \tag{I}
\end{equation*}
$$

Thus if a certain fixed voltage or potential difference $V$ CD is impressed across points CD, the potential difference $V C D$ may vary from zero to $V_{C D}$ depending on the position of the sliding tap T. The laboratory potentiometers are equipped with dials which tell where along the wire point $T$ is for a particular setting of the dial --- a reading of .400 for example means $I_{T D}=.400 I_{C D}$ (decimal point omitted on dial). OPERATION:

The process of measuring an unknown potential difference with a potentiometer is one of comparing the unknown to an accurately known voltage. For example in the circuit shown on the next page suppose an unknown voltage is placed across terminals $X$ and switch $S_{1}$ is to the right. Suppose in addition that it is known that the voltage drop across $C D$ is exactiy $V_{C D}=2.000$ volts. Now suppose that tap $T$ is moved along the wire until a point is found for which there is no noticeatbe deflection of the galvano-
 condition which could only arise if $T$ and $E$ were at the same potential. Therefore at this balance point $V_{X}=V_{E D}=V_{T D}$.
But equation (I) tells us that if $T$ is halfway between $C$ and $D_{2} V_{T D}=1.000$ volts, or if it is 0.600 of the way from $D$ to $C$, $V_{T D}$
$=(0.600)(2.000)=1.200$ volts, etc. Thus $V X$ is determined by noting what fraction of the wire (fraction ofthe 2.000 volts across the wire) it will balance against. However, the process of getting an accurately known voltage (2.000 volts here) across CD must be accomplished frist -- a process called calibration.

The potential difference $V_{C D}$ is controlled by the variable resistances in series with resistance $R_{C D}$. If the variable resistance is increased, the current $I D^{\circ}$ is decreased and thus the voltage $V_{C D}=I R_{C D}$ decreased; decreasing the variable

resistance increases $V_{C D}$. Suppose now that switch $S_{I}$ makes contact with the standard cell (left) which is an accurately known, constant emf. having the value (for exampie) $\varepsilon=1.500$ volts. As before, if at any position of the tap there were no current through the galvanometer when the $R$ or 0 switch is closed, the voltage drop across TD would have to equal $\mathcal{E}$. If one puts the tap $T$ three-fourths of the way from $D$ to $C$, then adjusts the variable resistance until there is no galvanometer current, he is assured of having placed 1.500 volts across TD (3/4 of the wire) and by proportion [equation (1)] of having placed 2.000 volts across CD (the whole wire). The process of calibration becomes one of placing the standard voitage across a certain fraction of the wire in order to have the desired voltage $V_{C D}$ across the whole.

INPORTANT NOTE: In balancing the potentiometer always tap Switch $R$ first until the galvanometer shows no deflection and only then start using switch o. Do not hold either switch down; do not touch switch 0 until balance is made with $R$-- very small currents drawn from the standard cell for a short time will ruin it.

INSTRUCTIONS:
I. A. Calibrate the potentiometer to read a maximum potential $V_{C D}=3.00$ volts using the standard cell mounted on the potentiometer board.
B. Connect a dry cell to terminals $X$ (note correct polarities) and measure its emf. Next, check the calibration again to see if $V$ has changed. If it has, repeat the calibration and measurenent.
C. Set up a circuit as shown below and measure the voltage $V_{B C}$ using a voltammeter and again using the potentiometer.


Assuming the internal resistance $r$ of the 32 volt source to be negligible compared to the other resistances, what should voltage $V_{B C}$ be according to theory (roughly - the source voltage is only approximately 32 volts)? How do the voltages measured using the voltmeter and potentiometer compare with this value? Explain discrepancies and calculate the internal resistance of the voltmeter from these measurements assuming that the meter is calibrated correctly. From your data what is the best value that you can obtain for the actual emf of the "32 volt" source?
II. A. Connect the unknown ("black bow") to terminals $X$ (this contains a source of emf. and an internal resistance) and measure the emf.

Now set the dials of a resistance box on 500 ohms inirst and then connect it across the terminals of the unknown as shown below.


Measure the terminal voltage $V_{T}$ of the "black box" with a 500 ohm resistor drawing current from it. Repeat with $\mathrm{R}=200,100,70,50,40,30,20$, and 10 ohms checking the calibration occasionally (caution: be very careful to have switch $S_{2}$ open while turning dials on the resistance box -an occiciental resistance of less than 5 ohms across the terminals will draw enough current to burn out the box).
B. Plct a graph of the terminal voltage $V_{\mathrm{f}}$ of the "black box" versus the current I it supplies to the series loop shown above. Since

$$
V=\varepsilon-I r=I R
$$

this graph should be a straight line with slope ( $-r$ ) and intercept $\varepsilon$ (if it is not, perhaps $\varepsilon$ is changing as current is drawn from the box). Determine the value of the internal resistance of the unknown from the slope of the line or, if it is curved, from the slope of the tangent at $I=0$. Record the values of $\varepsilon$ and $r$ for the unknown.

EXERCISES:

Given a circuit such as that shows below.


1. Show that potential difference $V=\left(\frac{P_{2}}{P_{I}+F_{2}}\right) \varepsilon$

If $V$ is to be equal to (I/IO) E, what fraction must $\mathrm{R}_{2}$ be of $\mathrm{R}_{1}$ ? This circuit arrangement is called a voltage divider.
2. If $R_{2}=200$ ohms, $R_{I}=1800$ ohm, and $E=100$ volts, What is the smallest resistance that a ( $100 \%$ accurate) voltmeter can have if it is to measure the voltage $V$ with an error of less than 5\%?

$$
\frac{100}{1800 \times R_{+}}=\frac{9.5 v}{R_{X}}
$$

$$
R_{+}=\frac{200 \operatorname{Rin} x}{200+\operatorname{Rin}_{x}}
$$

## HIGH RESISTANCE MEASUREMENT

OBJECT: To measure a high resistance using the ballistic galvanometer.

## CIRCUIT:



## THEORY:

A. Capacitor Discharge

If a capacitor is charged so that charge $q_{0}$ is on each plate and then allowed to discharge through a resistor as shown below, the potential difference across the capacitor must equal that across the resistor (they are across the same two points in the circuit $a b$ ) at all times during the discharge.


$$
\begin{aligned}
& q=\frac{\text { charge on plates at any }}{t i m e} t \\
& i=\text { current at the same time }
\end{aligned}
$$

$$
\begin{align*}
& V_{a b}=\frac{q}{C}=i R \quad i=-\frac{d q}{d t} \\
& \frac{1}{C} q=-R \frac{d q}{d t} \tag{1}
\end{align*}
$$

from which one can derive the expression for the charge still on the plates at any time $t$ after discharging starts

$$
\begin{equation*}
q=q_{0} \in-t / P C \tag{2}
\end{equation*}
$$

or

$$
\begin{equation*}
\ln \left(q / q_{0}\right)=-\frac{1}{R C} t \tag{3}
\end{equation*}
$$

## B. The Ballistic Galvanometer

The moving element of this type of galvanometer consists of a rectangular coil which is suspended between the poles of a magnet by a fine wire. When a charge q passes through the coil, the forces exerted by the magnetic field on the moving charges produce a turning moment or torque on the coil. The torque gives the coil an angular momentum, but the coil has a relatively large moment of inertia so that very little actual motion occurs in the time that it takes for charge $q$ to pass through the coil. The coil continues to rotate however, twisting the suspension wire. The twisted suspension wire now exerts a restoring torque which decreases the angular momentum of the rotation, brings the rotation to a stop at some angle $\theta$, and increases the angular momentum in the opposite direction. The result would be oscillatory motion of angular amplitude $\theta$ as long as no mechanical energy was lost from the system (it is almost frictionless). It is not difficult to show that the angle $\theta$ is proportional to the charge $q$ which passed through the galvanometer coil.

If one wishes to damp the oscillations of the galvanometer it is useful to recall that a coil rotating in a magnetic field generates an induced eme. Short-circuiting the galvanometer terminals completes an external circuit so that this emf can cause a current flow in the low resistance short-circuit. Thus the mechanical energy of the rotating coil is converted to electrical energy as in a generator, and this electrical energy is in turn dissipated as heat by the circuit resistance. The loss of mechanical energy by the system results in a quick damping of the oscillation.

## INSTRUCTIONS:

I. A. Charge the capacitor to the battery voltage by shorting $A$ to $B$. Then short $B$ to $C$ and allow the capacitor to discharge through the galvanometer, noting the deflection $D_{0}$.
B. Recharge the capacitor as before, but this time connect $B$ to $D$ for a measured time $t$ ( 5 seconds, 10 seconds, or whatever proves suitable) and allow the capacitor to discharge through resistance $R$ for that time interval before discharging the remaining charge through the gaivanometer. Again record the galvanometer deflection $D$.
C. Repeat step $B$ for longer and longer time intervals t. Plot a graph of galvanometer deflection $D$ versus discharge time t. How does this graph relate to a graph of the charge $q$ on a discharging capacitor as a function of time?
II. A. Prove that $\ln \left(D / D_{0}\right)=-\frac{1}{R C} t$

- 3-
B. Plot a graph of some function of $D$ versus $t$ which may be reasonably expected to yield a straight line. Determine the value of the resistance $R$ from the slope of this line.*
III. To see if leakage of charge off the capacitor's plates is a factor which might cause error, charge the capacitor once again and let it sit for 5 minutes before discharging through the galvanometer. Compare the deflection due to the charge which remained on the plates with that due to the original charge qo.
* The capacitors have the following values for the different circuit boards:

$$
\begin{array}{ll}
\text { no. } 1 & c=0.88 \mu f \\
\text { no. } 2 & c=0.90 \mu f \\
\text { no. } 3 & c=0.88 \mu f \\
\text { no. } 4 & c=0.88 \mu f \\
\text { no. } 5 & c=0.91 \mu f
\end{array}
$$

## Electrical Conduction in Semiconductors

References: Halliday and Resnick, Physics, part II, Chap. 31. sec. 1-4; example 2; chap. 32. prob. 15 on Wheatstone bridge: Holden, A. "The Nature of Solids", chap. 13.14; Feynman, "Lectures on Physics", Vol. III, chap 14, sec. $1-5 ;$ and Purcell, "Electricity and Magnetism" (Berkeley Physics Course) Chap. 4 .

Object
1 - to investigate the temperature-dependence of resistance of a semiconductor

2 - to use a d.c. bridge circuit as a null-reading
technique.

## Theory

According to quantum mechanics an electron which is moving along in a crystalline solid may only have certain energies, these allowed energies being grouped in bands with gaps of forbidden electron energies between. There is a limit to the number of electrons which can take upenergies within a given band; when this limit is reached the band is said to be filled. The distribution of electrons among the allowed energies is different in different solids but it is found in all pure semiconductors at $T$ equals 00 K , that the lower bands are completely filled and the upper ones completely empty. The highest energy filled band is called the "valance" band and the next higher energy band the "conduction" band. The energy gap $E_{g}$ between these two bands is small ( $\sim l$ electron volt) in these materials (in comparison with insulators).


It can be shown that electrons of completely filled band do not contribute to the conduction process; they can not be given. as a group, a net drift in one direction by applying an electric field. Thus a pure semiconductor at absolute zero would be a perfect insulator (conductivity $=0$ ). since the application of an electric field produces no current. Howevers if one puts some energy into such a system, raising its temperature, some of this energy will be taken on by electrons in the valence band which then (in the energy absorption process) jump the forbidden gap and appear in the conduction band. Thus at finite temperatures there will be some electrons in the conduction band and an equal number of "holes" in the valence band. When an electric field is applied to the sample the electrons in the c. b. are given a drift velocity in the opposite direction to that of the fieldi the val. b. holes act like positive particles and drift in the same direction as the field. The conductivity of a given sample is determined by the numbers of conduction electrons and holes per unit volume and by the mobilities (average drift velocity per unit applied field E of these charge carriers.

$$
\begin{equation*}
\sigma^{-}=n e \mu^{\mu}+p e \mu_{h} \tag{1}
\end{equation*}
$$

where $n=$ concentration of electrons in c.b. $\mathrm{p}=$ conc. of holes in v .b.
e = electronic charge magnitue, hole charge magnitude
$\mu_{h} \mu_{\mathrm{e}}=$ mobility of holes, electrons
Note that increasing the temperature of one of these materials causes more electrons to absorb energy and jump from the valence to the conduction band, increasing the number of conduction electrons $n$ and holes $p$. So raising the temperature of a pure semiconductor increases its conductivity instead of decreasing it as in the case of metals where the number of carriers is not a function of temperature and the controlling factor is the decrease of mobility with increasing temperature. In pure ("intrinsic") semiconductors the relationship between temperature and the number of each type of carrier per unit volume is

$$
\begin{equation*}
n=p=c T^{3 / 2} e^{-E_{g} / 2 k T} \tag{2}
\end{equation*}
$$

where $k$ is the Boltzmann constant, and $c$ is a constant. Note that increasing $T$ increases the number of both types of carriers and that decreasing the energy gap $E_{g}$ has the same effect since this makes it easier for valence band electrons to gain enough energy to jump to the conduction band.

If one substitutes equation (2) into equation (1) and takes the natural logarithm of both sides he obtains

$$
\ln \sigma=-E_{g} / 2 k T+\ln \left[\operatorname{ce}\left(\mu_{e}+\mu_{h}\right) T^{3 / 2}\right]
$$

or in terms of the resistivity $Q$ of the sample,

$$
\ln Q=E_{g} / 2 k T-\ln \left[\operatorname{ce}\left(U_{e}+u_{h}\right) T^{3 / 2}\right] \text {. }
$$

The first term on the right varies much more strongly with temperature than the second; over a limited temperature range one can consider the second term nearly constant. Therefore if a graph of in versus

1/T is plotted, the result should be a straight line of slope ( $E_{g} / 2 k$ ) to a good approximation. In this way the energy gap may be found.

Instructions


The resistance of the sample is measured using a Wheatstone bridge. The sample is placed in a calorimeter containing water, which may be heated.
I. Before coming to lab review the bridge theory and operation. Derive the fact that if no current flows throught the galvanometer in the above circuit when the switch is closed, $\mathbb{R}_{3} / \mathbb{R}_{4}=\mathbb{R}_{1} / \mathbb{R}_{2}$ (from which one of the $\mathbb{R}^{\text {is }}$ may be determined if the other three are known). In the set up you will use. there are two switches. Verify for yourself that when one of these is used $\mathbb{R}_{1}=\mathbb{R}_{2}$, and that use of the other produces a 10 to 1 ratio in these two resistances.
II. Connect the circuit as shown and measure the unknown resistance $R_{3} b y$ adjusting $R_{4}$ until there is no defiection of the galvanometer when you tap the o button on its case. (Start by using the less sensitive $R$ button until you are close to the balance point.) Record $\mathbb{R}_{3}$ and the temperature. Turn on the heater and raise the temperature of the sample about five digrees C. Disconnect heater, wait il $T$ stabilizes, and again measure $\mathbb{R}$ and $T$. Measure four more points or so. Do not allow the temperature to exceed $90^{\circ} \mathrm{C}$ or R to go below 15 ohms. If time permits, measure at water-ice temperature.
III. Determine the energy gap graphically. Discuss errors carefully. Note that the meter you used need not be calibrated. as only zero ("null") deflections were used at bridge balance. This fact removes one source of systematic error, namely an improperly calibrated scale.

Compare your results with those you would expect were the sampee a metal, such as copper. Discuss reasons for any differences.

## Linear and Nonlinear Circuit Elements

Object: This experiment actually has two objects, (1) to demonstrate the use of the method of differences in the determination of the mathematical relationship between two variables and (2) to find out how the current through various circuit elements depends on the potential difference applied to them.

Theory: The resistance of a circuit element is defined as the potential difference across the terminals of the element divided by the current flowing in the element. (If the potential difference is in volts and the current in amperes, then the resistance is in ohms.) In a number of circuit elements. for example the common resistor. the current through the element is directly proportional to the voltage across it. provided that the current is not large enough to cause a significant amount of heating or the voltage large enough to cause electrical breakdown. When the current is proportional to the voltage the resistance $R$ in the equation $I=V / R$ is constant and the element is referred to as being linear since a graph of $I$ versus $V$ would yield a straight line through the origin.

Nonlinear elements are those in which. for one reason or another, the current is not directly proportional to the potential difference across them. One can still define $R=V / I$ for these elements but it is apparent that $\mathbb{R}$ is no longer a constant in this case, but is a function of the current through the element. One example of a monlinear element is an ordinary light bulb which, as the current is increased, heats up more and more, resulting in an increase in the filament resistance. Some other circuit elements are intrinsically nonlinear (without any temperature change or other changes in experimental conditions). such as the vacuum tube diode or solid state diode. These particular examples are not only nonlinear but also non-symmetric $\quad$ having a much higher resistance to current flow in one direction ("backward") than to flow in the other ("forward"). Some diodes are nearly linear in the forward direction and are useful in circuits where it is desirable to have $I$ proportional to $V$ when the current flow is in the formard direction, and negligible current in the backward direction when the voltage tends to cause a flow in that direction (rectification). There are other elements which are symmetric but nonlinear.

The circuit elements investigated in this experiment are (1) a wire wound resistor. (2) a globar resistor which is actually a semiconductor (whose resistance changes with temperature), and (3) a solid state diode. Further information about semiconductors and solid state diodes is contained in the following references:

Shortiey and Williams, "Elements of Physics". 4th edition, pp. 712-14

Feynman. "Lectures on Physics", Vol. III. Chapter 14.

2.

## Instructions:

1. Connect the circuit as shown in the preceding diagram and have the instructor check your connections. Record the voltage across. and the current through, the resistor $\mathbb{E}_{1}$ as the voltage is varied in 0.1 volt steps from 0 to 1.5 volts. The voltage is varied by changing the resistance in series with resistor $\mathbb{R}_{1}$ using the variable resistance rheostat in series with the $6 \overline{8} 0$ or 330 ohm resistor, or by itself. The 22 ohm resistor is never to be taken out of the circuit or electrically bypassed during the
1 entire experiment. When you have obtained your highest voltage reading and current readings reverse the polarity of the power source and the voltmeter and ammeter and record the voltage and current. Does the direction of the current have any effect on the resistance? Plot graphs of $I$ versus $V$, making one direction of current flow and one potential difference polarity positive on the graph and the other negative. Also plot the resistance $R$ of the resistor for positive (one direction) and negative (the other direction) current.
2. Repeat part 1 for the diode R2 (INJOO) and the globar resistor R3 (Workman Electronic Products, model FB9). The 180 ohm resistor should always be left in series with the diode and the voltage across the diode may be varied in 0.05 volt steps from about 0.6 volts to 1.0 volts. Increase $V$ across the globar resistor in 0.2 volt steps between 2 and 5 volts. In the case of the globar resistor you will need to keep readjusting the voltage after you set it since the resistance of the element will continue to change until the temperature levels off at the new value. Record voltage and current readings only after equilibrim is obtained at the required voltage.
3. 

Using the method of differences (see separate instruction sheet) determine power series $I=A+B V+C V^{2}+\ldots$ which approximate the behaviour of these three circuit elements (for one direction of current flow only....the direction in which you have complete data). In each case make a table such as that shown below and compute first, second, etcetera differences. Stop taking differences when the last column shows no trend but only random deviations.

## TABLE I

Potential
Difference
V (volts)
0

First
Current
I (ma)
0

11

38

79
138
$\therefore$
11

27
:
41
$-$
59

Difference
$\Delta I(\mathrm{ma})$
$\because 16$
etc.
Record a decision as to the number of terms that must be kept in each case (resistor, diode, globar resistor) in order to closely approximate the correct relationship between $I$ and $V$ for each of these elements.

Finally, take the three expressions which approximately relate $I$ to $V$ for the three circuit elements (e.g. $I=A+B V+C V^{2}$ ). and substitute equally spaced values of voltage $0, \Delta v, 2 \Delta v, 3 \Delta v$, 4 \&v....into the expressions. Make another differenctable for each of the three expressions such as that shown below. By comparing each table with the corresponding table made from the data determine the constants $A, B, C \ldots$...for that circuit element.
4.

## TABLE II




PENDULUM.S $4 \frac{5^{2}}{4}$
(a)


(3)




$$
\left(9-\frac{9}{d}\right. \text { NO: INOD) }
$$


COMPUTATIONOE O FROM THE PERIOD




4



CONCLUSLONS
$n=\frac{4}{4 x}$
$\mathscr{L} 2 \boldsymbol{L}$

$$
2 y-022=\frac{t r y}{t r y}=4466
$$ $1=\infty-$

                                    \(y=\)
                                    \&
                                    148
                                    B \({ }^{1}\)
    P1
$14 y$
-214

$\cdots$$-$
$\square L C L L T G L O L \leq$

$$
Y=\sqrt{x} x+\sqrt{y} \rightarrow+\Delta=4+4+4
$$

$$
{ }^{*}=\mathrm{va}
$$ $\square+\quad$

EALLING BODIES NOV. 7.968
PURPOSE:TO USE INGENUITY IN THE
QUANTTATIVE STUDY OF FALLIN
OBSETS 3
3
18 $-\pi m=40$ Nowo


## Э18 81




DATA ANALYSIS OF FALLING BODY EXPERIMENT FROM THBEEA

QUADRATTC.... $Y=A(1)+A(2) * X+A(3) \div X \because X$

ERROR
$-0.24601980$
$-0.16561540$ $-0.15379606$ $-0.06056179$ 0.01408742
0.02015156
0.10763064
0.17652463
0.17683359 0.20855748 0.22169633
0.31625008
0.34221878
0.39960241
0.48840099
$-0.49138557$ $-0.38975712$
$-0.40671378$ $-0.34225547$ $-0.29638224$ $-0.26909393$ $=0.16039085$ $-0.12027293$ $-0.0487399$ 0.00420802 0.08857089 0.10434865 0.20154142 0.28014898

PERCENT ERROR
15.41458372
5.23169699
3.07358780
0.85171608
0.14850889
0.1661320 ?
0.71551658
0.96866616
0.81591031
0.82136919
0.75462604
0.94027602
0.89685190
0.93038117
1.01725625
0.9212088
0.66128051
0.62757970
0.4824423
0.38318825
0.3202771
0.17633043
0.12251461
0.04613292
0.00371092
0.07295103
0.08045656
0.14578204
0.19048702

DATA ANALYGIS OF FALLING BODY EXPERIMENT D. $\%$ TH - FROM TABLE +2
1.00000000
2.00000000
3.00000000
4.00000000
5.00000000
6.00000000
7.00000000
8.00000000
9.00000000
10.00000000
11.00000000 12.00000000 13.00000000 14000000000 15.00000000 16.00000001 17.00000001 18.00000001 19.00000001 20.00000001 21.00000001 22.00000001 23.00000001 24.00000001 25.00000001 26.00000001
27.00000001
28.00000001
29.00000001
1.35000000
3.00000000
4.85000000
7.05000000
9.50000000 12.15000000 15.15000000 8.40000000 21.85000001 25.60000001 29.60000001 33.95000003 38.50000002 43.35000002 $48 \cdot 50000002$ 52.85000002 58.55000001 $64 \cdot 40000006$ 70.60000002 77.05000004 83.75000005 90.80000004 98.05000004 $105 \cdot 60000002$ 113.40000006 121.50000005 129.80000013 138.45000004 147.35000008

Al 3
0.1342925325

## ( 12

1.1667180005

A(1)
0.2950092772

STD. DEV. 0.2614851326




$$
\begin{aligned}
& t D / D 0 \quad 100190 \\
& \text { 2. } 11 / 22=1.00=000 \\
& \text { (15)/22.928-1020 } \\
& 101042=122 \text {-1088 }
\end{aligned}
$$

$$
\begin{aligned}
& 20 \text { 1925.81\% -200 } \\
& 301545 a=18{ }^{3}-330
\end{aligned}
$$

$$
\begin{aligned}
& 6011426.588-1.54
\end{aligned}
$$

$$
\begin{aligned}
& 12954 / 2.44 \quad 1,36 \\
& 150492+170
\end{aligned}
$$

$$
\begin{aligned}
& m=\frac{30}{B D}--0 m / \operatorname{coc} \\
& D_{n} D_{0}=\left(1 / 10^{-2}\right) t
\end{aligned}
$$

$$
\begin{array}{ccc}
\text { LR } & \text { Cl } & \text { Cl }
\end{array}
$$

2

$$
\begin{gathered}
V \\
6 \\
67 \\
26 \\
25 \\
66 \\
64 \\
48 \\
44
\end{gathered}
$$


-smo1+67n27t?

$$
y
$$


FORCER RLS. (CONT)

```
*amamen
```

है
$4+4$
0
$x^{2}+8+4+48 d$


ELEGTELGAL GONDUGIION IN
SEMLCONDUCTORS
by Robext Jackson Mavks IT Physics III Group A
Friday; Periods 1-3
Performed: 14/29/69
Due: 12/5/69


## PURPOSE:

To investigate the temperature dependence of a resistance of a semiconductor by using a $D . G$. bridge circuit as a null-reading technique.

## RROCEDURE:



The resistance is measured by the use of a "Wheatstone Bridge". The sample is placed in a calorimetexcontaining water, which may be heated or cooled. Temperature is measured off of the themometer, which is placed near the semiconductox. Resistance is measured by varying it so the galvonometer reads zewo, keeping $i=0$ and making calibration of the galvanometer unnescisisary.
since:

$$
\ln (x)=\mathrm{E}_{\mathrm{g}} / 2 k T-\ln \left(\operatorname{ce}\left(u_{e}+u_{h}\right) \mathrm{I}^{3 / 2}\right)
$$

where:

$$
\begin{aligned}
& r=\text { resistivity } \\
& T=\text { temperature } \\
& \mathrm{E}_{\mathrm{g}}=\text { energy gap } \\
& \mathrm{c}^{\mathrm{G}}=\text { Boltzman } \\
& c=\text { constant (characteristic of semiconcuctor) } \\
& \mathrm{u}_{\mathrm{h}}=\text { mobilite of holes } \\
& u_{e}=\text { mobility of electrons }
\end{aligned}
$$

and $\left(-\ln \left(c e\left(u_{e}+\mathrm{H}_{\mathrm{h}} \mathrm{T}^{3 / 2}\right)\right.\right.$ ) changes minimally in comparison with $\mathrm{E}_{\mathrm{g}} / 2 \mathrm{kT}$, a graph of $\ln (x)$ vs. $1 / \mathrm{I}$ would gield a good approximation of the energy gap $\left(E_{g}\right)$, since its slope would be a good approximation of $z_{g} / 2 k$.

Resistance, however, is much easier to directly measure than is resistivity. One may derive one from the other Erom the formula:

$$
x=R L
$$

where:
R=xesistance
L=number with dimention of length charicteristic of the semiconductor's geometry

The shape of the semiconductor changes minimally with the temperature, but this change is so small it can be disxegarded. Therefore, $I$ can be treated as a constant.

Therefore, if:

$$
x=R L
$$

then:

$$
\ln (x)=\ln (R L)
$$

since:

$$
\begin{aligned}
& \ln (r)=\operatorname{Ig}_{g} / 2 \mathrm{k} T \\
& \ln (R L)=\ln (R)+\ln (L)=E_{g} / 2 k T
\end{aligned}
$$

or:

$$
\ln (\mathrm{R})=\mathrm{E}_{\mathrm{g}} / 2 \operatorname{LR}-\ln (\mathrm{L})
$$

$\operatorname{Ln}(L)$, being a constant, would imply that the $\ln (x)$ vs. $1 /$ graph would have the sme slope as the $\ln (R)$ Vs. $1 / T$ graph, being Eg/2k.

| $\frac{x^{0}}{6} \cdot 3^{\circ}$ | $\frac{\mathrm{T}_{\mathrm{E}} \mathrm{O}_{4}}{-5^{0}}$ | $\frac{\text { R-ohms }}{6 \cdot 5 n}$ | $\frac{\text { In }_{1}(R)}{t_{0}(007}$ | $\frac{1 / 5}{5 \times 10^{-5}}$ |
| :---: | :---: | :---: | :---: | :---: |
| 2.6 | 276 | 595 | 6.89 | .00362 |
| 3.6 | 277 | 586 | 6.84 | .00361 |
| 7.7 | 281 | 539 | 6.54 | .00356 |
| 24.5 | 298 | 221 | 5.45 | . 00336 |
| 42.3 | 315 | 104 | 4.64 | . 00317 |
| 46.5 | 320 | 94.0 | 4.54 | . 00313 |
| 49.2 | 322 | 86.2 | 4.45 | . 00311 |
| 50.0 | 323 | 85.9 | 4.45 | .00310 |
| 65.0 | 338 | 51.3 | 3.94 | .00296 |
| 66.5 | 340 | 49.2 | 3.90 | . 00294 |
| 73.1 | 346 | 40.2 | 3.70 | . 00289 |
| 75.0 | 348 | 37.5 | 3.52 | . 00287 |
| 83.9 | 357 | 29.4 | 3.38 | . 00280 |

LINEAR*OOO Y $=A(1)+A\left({ }^{\circ}\right.$
$E R R O R$
-0.00003100
-0.00002936
-0.00000953
0.00005583
0.00008464
0.00002602
0.00001697
-0.00000430
-0.00001499
-0.00001844
0.00000345
-0.00003395
-0.00004533

PEF


## CALCULATIONS:

From the preceeding datag Eg may be computed.

$$
\begin{aligned}
& m(\ln (R) \text { vs. } 1 / T \operatorname{graph})=\mathrm{E}_{\mathrm{g}} / 21 \mathrm{t} \\
& \mathrm{E}_{\mathrm{g}}=\left(2.33 \times 10^{-4} 3 \times 3 \times 10^{-5}\right) 2\left(1.38 \times 10^{-23}\right) \\
& \mathrm{E}_{\mathrm{g}}=6.43 \times 10^{-27} \text { jouLes }
\end{aligned}
$$

Also, the constant $L$ may be computed, in that the "y" intercept of the $\ln (\mathrm{R})$ vs. $1 / \mathrm{I}$ graph is equal to the $\ln 1 / L_{0}$

$$
\begin{aligned}
& \operatorname{In}(1 / \mathrm{L})=.00205 \\
& I=e^{-.00205} \\
& I=1.00 \text { meters }
\end{aligned}
$$

## CONCLUSTONS:

The energy gap of an onknow semiconductor can be determined by finding the relationship of its resistance and change in temperature, keeping all other facets constent. A graph of $\ln (R)$ vs. $1 / T$ yields a straight line, the slope of which is $\mathrm{Eg}_{\mathrm{g}} / 2 \mathrm{k}$, and whose intercept is equal to the $\log$ of the 1 inear constant by which the resistance is multiplied to yield the resistivity.

In this experiment, the energy gap was found to be equal to be $6.43 \times 10^{-27 t} 8.35 \times 10^{-28}$ joules. and the Iinear constant equal to 1.00 meter.

The greatest amount of exror in this experiment undoubtedly arose from the corvilation of a temperature and a resistance. The temperature read off of the themom meter was not that of the semiconductor, for though the temperature fell or rose slowly, the semiconductor's resistance could be visibly seen changing on the galm vanometer. This temperature exror could have varied. as much as $5^{\circ}$.

The deleting of the last part of the equation, the small change in the semiconductoris geometry, and the possible error in the calibration of the variable resistor ane minimal compared to the temp-res reading error, and may be disxegarded.

Physics Laboratory - General Instructions

## I. Purpose of Laboratory

Laboratory work in physics has two important objectives first, to give the student direct experience with some of the natural phenomena upon which physical principles are based, and second, to develop in the student some understanding of the experimental procedures. It is felt that some experience in the laboratory is necessary to give the student an insight into the methods of physics (or for that matter any experimental science). Without it he would be merely accepting principles as they were handed to him without an undexstanding of the experimental procedures on which they are based.

In the laboratory the student will work with real, rather than ideal, apparatus. This equipment (and the experimenter as well) will be subject to limitations which cause errors that must be taken into account before any conclusions can be drawn from the experimental results. Therefore error analysis is an essent= ial part of all good laboratory work.

Although you will be assigned a certain group of experiments to do this quarter, and in many cases the procedure to be followed in performing the experiment is described in an instruction sheet, it is hoped that the student will use some of his own ingenuity in perfoxming the experiments; it is intended that the instructions be used as an aid to understanding rather than something to be followed mechanically without thought. We also want to encourage students to think about possible experiments that they might do in place of one of the prescribed set. Within the limitations of equipment and time, substitution of an experiment which is more interesting to the individual student is permitted, provided it is a physics experiment and it is cleared with the instructor.

## II. Preparation for an Experiment

In order to perform an experiment thoroughly and accurately in the time allotted, it is necessary to put in some time beforehand thinking about the experiment. If an instruction sheet has been provided it is to be studied carefully before the laboratory period. You should come to the laboratory with as thorough an understanding as possible of what you are going to do during the period and why. This may require that you spend some time in the library, looking up references etcetera.

## III. Performance of the Experiment

An essential part of the method of solving an experimental problem is the preparation of a clear, concise record of the data taken during the performance of the experiment. This record should contain, in a clear and legible form, all the "raw" data and information with which to make corrections (don ${ }^{9}$ t try to make corrections "in your head" while taking data) and also enough explanation of what you are doing and why so that your pages of
data can be analyzed later without confusion or ambiguity. your instructor may require that this record be kept in a permanent notebook or he may ask you to keep this record on data sheets which are later included in a report on the experiment. In either case, all observations should be recorded directly into the notebook or on the data sheets (nothing on scratch paper and later copied) and an estimate of the accuracy of each set of measurements should be made and recorded also. Corrections can be made by crossing out errors with a single line (no erasures) Before leaving the laboratory, the student should do enough calculation and graphical work to ensure that the data collected "makes sense" and there are no gaps in it which need to be filled in before he can continue the analysis without having to make any "wild guesses or assumptions. Your data record must be approved by the instruct or before you leave the laboratory.
IV. Laboratory Notebook (Data Record)

The following are specific suggestions concerning the form of the laboratory record of the experiments.
A. If the instructor has you keep a permanent laboratory notebook it should be one having cross-ruled pages (useful for graphs) and it must be labeled with the following information.

1. On the eront cover in ink:

Physics Laboratory
Your Name
2. Inside the front cover at the top:

Fall (or whatever) Quarter
Lab. day and hours
Group Number
B. For each experiment the student should record the title of the experiment and the date performed at the top of the data record. A very brief (not detailed) description of the procedure followed shoulc precede the data record. which is preferably in tabular form. Label the data carefully with the proper column headings and units. Whenever possible, the type and identifying number of instruments being calibrated or used in measurement should be recorded for later reference.


When a report is required on an experiment it is due at the beginning of the period one week after the experiment was performed. The report is to be written independently by each student in ink (or typewritten) on white, unlined 0 $8 / 2 \mathrm{x}$ ll paper (graph paper for graphs). Each report must have:
A. A cover sheet containing the following information - course, experiment title, your name, laboratory period day and hours, group number, date expeximent was performed, and date of report.
B. A statement of the purpose of the experiment and a brief summary of how you went about performing it (not detailed), data and observations (if you used data sheets rather than a notebook these may be submitted as they are), sample calculations, tabulated results, graphs, conclusions, and a discussion of the experiment. The discussion section of a report should be more thorough and complete than the corresponding section in the notebook. It may include a discussion of what was learned in doing the experiment: as well as the results and the accuracy of the results. It should also contain a discussion of any points which the instructor may have brought to your attention through questions written on the instruction shoets, and of any other points of interest that may occur to you.

It is customary to use the passive voice in scientific writing (e.g. "The time required for the pendulum to swing through twenty complete cycles was measured...etc.") thus not calling attention to the observer. The following styles are not to be used in a report: $I^{\prime \prime}$ (we) swung the pendulum and..." or "Swing the pendulum and measure the time for twenty complete cycles...". If you quote or paraphrase any outside sources in writing your report (including your own text book) give credit to the original author in a footnote.

References:

> 1. Baird, "Experimentation", chapter 7
> 2. Olson, "Experiments in Modern Physics", section 1.4

Measurement, Probability, and Experimenta: Errors

## I. Types of Error

Whenever a measurement is made of any physical quantity there is a certain amount of uncertainty in the result. Determination of the amount of uncertainty in a measurement is not usually easy but an attempt should always be made to do so, even if it is no more than an educated guess. Without some estimate of the uncertainties associated with experimental measurements one has no indication of the accuracy of the results and it is difficult to come to any conclusion about what the experiment has shown (or not shown). In all of the experiments which follow in the physics laboratory sequence the student will be expected to make some estimate of the accuracy of his quantitative experimental results.

There are two types of errors which may occur in the measurement process, systematic errors and random errors. Systematic errors tend to make all the observations of one item too small or too large. For example if voltage measurements were taken in an electric circuit using a voltmeter which consistantly read $0 . l$ volt too high, a systematic error would be present. Other common examples of causes of systematic error are worn weights, clocks which gain or lose time, friction, and personal bias of the observer which causes him to make readings which are consistently high or low. When systematic errors are recognized in an experiment it is often possible to find out how large their effect is and to correct for it. The error in the voltmeter which reads 0.1 volt too high, for example, can be discovered by calibrating the instrument against some sort of standard (accurately known voltage), and a correction of -0.1 volt made to all the readings. Error due to an observer's bias may be minimized by having another observer make the same measurement independently (bias is best eliminated if each observer knows nothing of the other's result until after both measurements are completed).

Random errors xesult from chance variations in the quantity being measured, in the measuring devices, or in the observer, and are just as likely to produce too large a value as too small. For example, if one measures the diameter of a metal rod several times with a micrometer the readings will probably fluctuate slightly in a non-systematic fashion due to actual differences in the rod's diameter at different positions, variations in pressure when the micrometers jaws are closed, and changes in the observer's estimate of the scale reading. Random errors are present in all measurements, although they may be too small to be noticeable, and they cannot be corrected for because of their random nature.
II. Determination of Precision

Suppose that several measurements of the same quantity $x$ were made and all systematic error in the measurements eliminated or corrected (assuming this were possible). As discussed above there would still be a certain amount of rnadom fluctuation apparent in the measurements if they are "fine" enough to make it noticeable. If a histogram was plotted showing the number of measurements $N$ falling within different intervals of size $\Delta x$ it might look like that shown in Fig. 1.

Fig. 1
No. of Measurements vs.


The meaning of the histogram is that one measurement of $x$ fell between 4.6 and 4.7 units, two between 4.7 and 4.8 units, four between 4.8 and 4.9 units, and so forth. The completely symetrical distribution shown usually results only if a large number of measurements are made and if the fluctuations are entirely random. In such cases the envelope of the distribution often has a particular form called a "normal" or "Gaussian" distribution which is represented by the mathematical equation

$$
y=\frac{1}{\sqrt{2 \pi} \sigma} e^{-(x-\bar{x})^{2} / 2 \sigma^{2}}
$$

where $\sigma$ is a constant which determines the "sharpness" of the peak (high, narrow peaks are characterized by small values of $\sigma$ ). The quantity $\bar{x}$ is the average of the individual measurements

$$
\bar{x}=\frac{x_{1}+x_{2}+\ldots \ldots \ldots \cdot\left(\frac{\sum x_{i}}{n}\right.}{n}
$$

where $n$ is the total number of measurements, and because of the symmetry of the Gaussian function $\bar{x}$ corresponds to the most probable value of $x$ obtained from a measurement of $x$ (peak of curve). Thus $\bar{X}$ is the best estimate that one maymake of the true value of $x$ from these measurements.

The individual measurements of $x$ differ from the average or most probable value $\bar{x}$ by an amount d called the deviation of that measurement

$$
d_{1}=x_{1}-\bar{x}, d_{2}=x_{2}-\bar{x}, \ldots \ldots \ldots .
$$

The standard deviation

$$
\sigma=\left[\frac{d_{1}^{2}+d_{2}^{2}+\ldots \ldots . d_{n}^{2}}{n-1}\right]^{1 / 8}=\left[\frac{\sum(d i)^{2}}{n-1}\right]^{1 / 2}
$$

is an indication of the precision of a set of measurements since narrow Gaussian distributions indicate precise measurements with small deviations from the average and a small standard deviation $\sigma_{0}$ If a large number of measurements is made, $68 \%$ of them will be in the range $\overline{\mathrm{x}} \pm \sigma, 95 \%$ in the range $\overline{\mathrm{x}} \pm 2 \sigma$, and $99 \%$ in the range $\mathrm{x} \pm 3 \sigma$, a fact which can be verified by determining the area under a Gaussian curve between the various limits. If after having determined $x$ and $\sigma$ from a large number of measurements one makes a single measurement $x$, he then will have about a two thirds chance of getting a value between $\mathrm{x}+\sigma$ and $\overline{\mathrm{x}}-\sigma$, etcetera.

Although increasing the number of measurements of quantity $x$ would have little effect on the standard deviation $\sigma$ (the scatter of the data) except to give more accurate picture of what it really is, increasing the number of measurements should improve the reliability of the average value $\overline{\mathrm{x}}$. It can be shown from statistics that the standard deviation in the mean $\bar{x}$ is given by the equation

$$
\sigma_{m}=\frac{\sigma}{\sqrt{n}}
$$

which means that there is a $68 \%$ chance that the true value of $x$ will be in range $\bar{x} \pm \sigma^{m}$ assuming the distribution is normal and there are no systematic errors present. Thus the precision of the mean $\bar{x}$ can be increased ( $\sigma_{m}$ reduced) by taking more observations, but the improvement is slow because of the $\sqrt{n}$ factor ( 90 readings only 3 times as good as 10 readings). The final result of a set of measurements may be stated

$$
x=\bar{x} \pm 5 \mathrm{~m}
$$

It is quite often useful to represent the standard deviation $\bar{\sigma} m$ as a percentage of the value $\bar{x}$. The calculation required is:

$$
\text { per cent std. dev. }=(\sigma \mathrm{m} / \overline{\mathrm{x}}) \cdot(100 \%)
$$

Although the normal or Gaussian distribution (equation 1) is very often a good representation of the kind of distribution found in repeated measurements of physical quantities, it should not be assumed that this distribution always gives an accurate description of the results of such measurements, even when a large number of measurements are made. There are a number of cases where the distribution is non-Gaussian and perhaps even non-symmetrical. For example, if one makes several determinations of the number of nuclei which decay by particle emission in a certain time, he obtains the poisson distribution

$$
\begin{equation*}
y \propto \frac{\overline{\bar{x}}^{x}}{x!} e^{-\bar{x}} \tag{2}
\end{equation*}
$$

where $\bar{x}$ is the average number of counts and $y$ is the probability of obtaining $x$ counts in a given trial. This distribution is very unsymmetrical about the mean $\vec{x}$ when the number of counts $\bar{x}$ is small but closely resembles a Gausian distribution with standard deviation $\gamma / \overline{\bar{x}}$ when $\bar{x}$ is large.
III. Propagation of Errors If one uses experimental observations, with their associated random errors, to calculate a result, the precision of the result will be determined by the precision of the quantities involved in the calculation. The standard deviation of the result may be determined from those of the
 following rules.
A. The standard deviation of the result of addition and/or substraction is the square root of the sum of the squares of the standard deviations of the separate terms.

$$
\text { Example: } \quad \begin{aligned}
x_{1} & =5.30 \pm 0.20 \text { units } \\
x_{2} & =1.70 \pm 0.10 \text { units } \\
x_{3} & =7.20 \pm 0.01 \text { units }
\end{aligned}
$$

$$
\begin{gathered}
x_{1}-x_{2}+x_{3}=(5.30-1.70+7.20) \pm\left[(0.20)^{2}+(0.10)^{2}+(0.01)^{2}\right]^{1 / 2} \\
=10.80 \pm 0.22 \text { units }
\end{gathered}
$$

Note that most of the standard deviation in the result comes from the largest standard deviation present in the separate texms $(0.22 \simeq 0.20)$ 。
B. The percentage standard deviation in the result of mulitplication and or division is the square root of the sum of the squares of the percentage std. deviations of the factors.
example: $\quad x_{1}, x_{2}, x_{3}$ as above
$(\% \text { std. dev. })_{1}=\frac{0.20}{5.30} \times 100 \%=3.8 \%$
$(\% \text { std. dev. })_{2}=\frac{0.10}{1.70} \times 100 \%=5.9 \%$
$(\% \text { std. dev. })_{3}=\frac{0.01}{7.20} \times 100 \%=0.1 \%$
$y=\frac{\left(x_{1}\right)\left(x_{2}\right)}{x_{3}}=1.25 \pm$ std. dev.
$(\%$ std. dev. $) y=\left[(3.8)^{2}+(5.9)^{2}+(0.1)^{2}\right]^{1 / 2}=7.0 \%$
$\left(s t d . d e v_{0}\right)_{y}=(.07)(1.25)=0.09$
$y=1.25 \pm 0.09$ units
Note that in this case the largest contribution to the standard deviation in the result comes from that quantity with the largest percentage standard deviation.
C. In case a quantity is raised to the $n$ ph power its percentage standard deviation is multiplied by $n$.

The process of carrying standard deviations through calculations is useful not only indetermining the precision of the result but also in determining which quantity contributes most to random error in the result. It may be possible to reduce the deviations in this quantity by using more care or different techniques.

## IV. Accuracy of Experimental Results

Determination of the standard deviation in an experimental result will tell you how much uncertainty is present due to random errors but this is an indication of the accuracy of the result only in the case where systematic errors are negligible compared to randomerrors. For example, if in a particular experiment you obtained a percentage standard deviation of $1 \%$ but the instruments used to obtain the measurements were accurate only
to within $5 \%$ (all readings may be too high or low by $5 \%$ ) then the $5 \%$ accuracy is a better indication of the reliability of the results than the $1 \%$. Some attempt should be made by the student to determine the reliability of his results in each experiment. although in some cases this will involve making some educated guesses as to the accuracy with which a particular measurement may be made with a particular measuring device. In all cases try to eliminate as much systematic error from the measurement as possible within the time available. An experimental result does not agree with a prediction of a theory unless the theoretically predicted result dies within the range given by the experimental result plus and minus the probable erxor; an experiment does not disagree with a theory unless the predicted result lies outside this range.

## V. Significant Figures

The term "significant figures" refers to the digits of a measurement made in the laboratory, including all the certain digits and one additional doubtful one based on the observeres estimate of a fraction of a scale division. The numbers which represent data or the results of calculations should always be given with neither more nor fewer significant figures than are justified by the precision of the observations and computations. The number of significant figures in a measurement (or a calcu-n lated quantity) may be determined using the following rules.
(a) The first significant figure is the first non-zero digit。
(b) Zeros which occur between significant digits are considered significant。
(c) Zeros which occur to the right of the last non-zero digit are considered significant when they are to the right of the decimal point (the significance of such zexos to the left of the decimal point is indeterminate).
(d) If numbers having a different number of significant figures are added. substracted, multiplied or divided. the answer is given so as to have the same number of significant figures as the term or factor which has the least.

Examples: $\quad .0001906$ has 4 significant figures
10,937 has 5
93,000 has an indeterminate number
$9.3 \times 10^{4}$ has 2
$9.30 \times 10^{4}$ has 3
VI. Comparison of Results

Sometimes an experimental result is arrived at by two different methods which should both theoretically give the correct result. If there is no reason to believe that one of the results is much moxe accurate than the other, it might be instructive to see how much difference there is between the two. This difference is usually given in terms of the "percentage difference" which is defined.

$$
\% \text { diff. }=\frac{\text { diff. between values }}{\text { average value }} \times 100 \%
$$

## Refexences:

1. Young, "Statistical Treatment of Experimental Data"
2. Barford, "Experimental Measurements: Precision。 Error and Truth"
3. Baird, "Experimentation: An Introduction to Measurement Theory and Experiment Design"
4. Braddick, "The Physics of Experimental Method"
5. Pugh and Winslow, "The Analysis of Physical Measurements"
6. Bevington, "Data Reduction and Error Analysis for the Physical Sciences"

## METHOD OF LEAST SQUADES

One of the fundamental problems that comes up again and again in the laboratory is that of finding, from simultaneous measurements of quantities $y$ and $x$, the dependence of quantity $y$ on quantity $x$ (the dependence of the period of a pendulum on its length for example). Often this dependence is revealed by making a graph of y versus $x$ from the data. However, a certain amount of judgement is always involved in making a graph from experimental data since deviations in the measurements usually make it impossible to draw a smooth curve through all the data points. One usually tries to draw a smooth curve among the points in such a way that it appears that the deviations of the points from the line (positive and negative) add up to approximately zero. In other words, in the graph shown below

$$
\left|d_{1}\right|+\left|d_{3}\right|+\left|d_{4}\right|+\ldots \ldots \ldots=\left|d_{2}\right|+\left|d_{5}\right|+\ldots \ldots \ldots
$$

where the deviations here and in the analysis to follow will be assumed to be deviations in $y$ for precisely known values of $x$.


If a high degree of precision is required in the expression relating y to $x$, this method of balancing deviations "by eye" might not be sufficient. In this case more scientific approach, based on statistics, is followed. It can be shown that the most probable disposition of the line representing the dependence of y on $x$ is that for which the sum of the squares of the deviations of the points from the line is a minimum (hence the name "least squares")

$$
\Sigma\left(d_{i}\right)^{2}=d_{1}^{2}+d_{2}^{2}+d_{3}^{2}+d_{4}^{2}+\ldots \ldots=a \text { minimum }
$$

This statement is called the "principle of least squares" and it is the basis of a method for finding the relationship between $y$ and $x$ which best fits the data points (for which the sum of the squares of the deviations is a minimum).

Actually the problem of determining the line which "best" fits a set of data points ( $x_{i}, Y_{i}$ ) is several different problems, depending on the type of curve which is to represent the relationship between $x$ and $y$. If it has been predetermined from the data or from theory that $y$ depends on $x$ linearly so that $y=A x+B$, the problem becomes one of picking out, from all possible straight lines, the one with values of slope $A$ and intercept $B$ such that the sum of the di2 will be as small as possible. If ( $\mathrm{x}_{\mathrm{l}} \mathrm{g}_{\mathrm{i}} \mathrm{y}_{\mathrm{l}}$ ) are the coordinates of the first data point, ( $x_{2}, y_{2}$ ) the coordinates of the second and so forth, and if it is assumed that the deviations are only in the $y$ measurement for precisely known $x$ 's, then

$$
\Sigma\left(d_{i}\right)^{2}=\left(A x_{1}+B-y_{1}\right)^{2}+\left(A x_{2}+B-y_{2}\right)^{2}+\ldots .
$$

If the "best" straight line is that which makes the sum of the squared deviations or a minimum.

$$
\begin{aligned}
& \frac{d\left[\Sigma(d i)^{2}\right]}{d A}=0=2 x_{1}\left(A x_{1}+B-y_{1}\right)+2 x_{2}\left(A x_{2}+B-y_{2}\right)+\ldots \\
& \frac{d\left[\Sigma(d i)^{2}\right]}{d B}=0=2\left(A x_{1}+B-y_{1}\right)+2\left(A x_{2}+B-y_{2}\right)+\ldots
\end{aligned}
$$

are the conditions which should lend to the "best" values of $A$ and B. These equations may be rewritten:

$$
\begin{align*}
B \Sigma x_{i}+A \Sigma x_{i}^{2}-\sum x_{i} y_{i} & =0  \tag{1}\\
n B+E x_{i}-\Sigma y_{i} & =0 \tag{2}
\end{align*}
$$

where $n$ is the number of points.
The method is illustrated below for a set of $n=5$ points.

| Point |  |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| No. | 1 | 2 | 3 | 4 | 5 |
| $x$ | 1.00 | 1.90 | 2.60 | 3.20 | 4.00 |
| $y$ | 0.90 | 3.00 | 4.00 | 5.50 | 6.90 |

A table is made as follows:

| $x$ | $y$ | $x^{2}$ | $x y$ |
| :---: | :---: | :---: | :---: |
| 1.00 | 0.90 | 1.00 | 0.90 |
| 1.90 | 3.00 | 3.61 | 5.70 |
| 2.60 | 4.00 | 6.76 | 10.40 |
| 3.20 | 5.50 | 10.24 | 17.60 |
| 4.00 | 6.90 | 16.00 | 27.60 |

$E \mathrm{xi}=12.70$
$\Sigma y_{i}=20.30$
$\sum x_{i}{ }^{2}=37.61$
$\Sigma x_{i} y_{i}=62.20$

Substituting in (1) and (2).

$$
\begin{aligned}
12.70 \mathrm{~B}+37.61 \mathrm{~A} & =62.60 \\
5 \mathrm{~B}+12.70 \mathrm{~A} & =20.30
\end{aligned}
$$

Solving simultaneously,
$B=-0.989$
$A=1.999$
The equation of the straight line which best fits the data points is

$$
y=1.988 \quad x-0.989
$$

In other words the sum of the squares of the deviations of the points from the straight line is a minimum for a line of slope 1. 988 and $y$ intercept -0.989 。

It is generally shown in books on statistics that the standard deviations in these values obtained for the slope $A$ and intercept $B$ may be found using the equations (3 and 4):

$$
\begin{aligned}
& \sigma_{A}=\left[\frac{\Sigma d_{i} 2}{n \Sigma x_{i}^{2}-\left(\Sigma x_{i}\right)^{2}}\right]^{1 / 2}=\left[\frac{\Sigma\left(A x_{i}+B-d_{i}\right)^{2}}{n \Sigma x_{i}^{2}-\left(\Sigma x_{i}\right]^{2}}\right]^{1 / 2} \\
& \sigma_{B}=\left\{\frac{\left(\Sigma d_{i}^{2}\right)\left(\Sigma x_{i}^{2}\right)}{n^{2} \Sigma x_{i}^{2}-n\left(\Sigma x_{i}\right) 2}\right\}^{1 / 2}=\left\{\frac{\left[\Sigma\left(A_{x_{i}}+B-y_{i}\right)^{2}\right]\left[\Sigma x_{i}^{2}\right]}{n^{2} \Sigma x_{i}^{2}-n\left(\Sigma x_{i}\right)^{2}}\right\}^{1 / 2}
\end{aligned}
$$

In cases where a nonlinear curve is to be fit to a set of data points in such a way as to make $\Sigma(d i)^{2}$ a minimum, equations (1), (2), (3), and (4) no longer apply. often one can get around this difficulty, however. For example, suppose some data points are to be fit with a parabola of the type $y=A x^{2}+B$ 。 If the quantity $x=x^{2}$ is calculatedfor each of the pointsp the method may then be applied to quantities $y$ and $X$, since y versus $X$ will be a straight line $(y=A X+B)$ even though y versus $x$ is not.

The least squares method is not confined to finding the constants of a straight jine, however; it can be applied to any kind of curve. For example, if one has a set of data points and wants to determine the constants of the "best fit" parabola $y=A x^{2}+B X+C$, he can apply the conditions that minimize $\sum(d i)^{2}$ with respect to variables $A_{,} B_{s}$ and $C$ and will obtain the equations:

$$
\begin{array}{ll}
\Sigma x_{i}^{2} y_{i} & =C \Sigma x_{i}^{2}+B \Sigma x_{i}^{3}+A \Sigma x_{i}^{4} \\
\Sigma x_{i} y_{i} & =C \Sigma x_{i}+B \Sigma x_{i}^{2}+A \sum x_{i}^{3} \\
\Sigma y_{i} & =n C+B \Sigma x_{i}+A \Sigma x_{i}{ }^{2}
\end{array}
$$

which may be solved simultareously for constants $A, B$ and $C$.

## References:

1. Young, "Statistical Treatment of Experimental Data"。 section 14.
2. Baird, "Experimentation", Appendix 2
3. Barford, "Experimental Neasurements", Chapter 3
4. Pugh and Winslow, "The Analysis of Physical Nieasurements" Chaptex 10.
5. Bevington, "Data Reduction", Chapters 6 and 11
6. Gerhold, "Least-Squares Adjustment of Weighted Data to a General Linear Equation", American Journal of Physics, Vol. 37, p, 156.


1-7-70

PARAMAGNETIC 1) VERY WEAK ATTRACTION
DIAMAGNETIC 2) " " FTROULSION
FERROMAGNETIC 3) STRONG ATTRACTION
(NO STRONG REPULIION)
2) MECHANISIM OF DIAMAGNETIC EFFECT
-) PARAMAGNETIC CONTAINS PERMANENT 4TOMIC OIPOLES (LENZ'LAW) 3) FERROMACNETIC-Fe, Ni, Co, ETE. MONOAT - TEST $9-70$
MAGNETIC DIPOLE MOMENT $\mu$ METHOD OIRECCTION




$$
\Rightarrow \frac{d^{2} Q}{d t^{2}}+\frac{1}{L C} Q=0
$$

CHECR BY
KIREHMOFF'S VOLTACE LAW:
IVAROUND LOOP $=0$
FON $L=C$ GINCUIT:

$$
\begin{aligned}
E V=0 & =\frac{Q}{E}+L \frac{d t}{d t} \\
& =\frac{d}{E}+L \frac{d t}{d t}\left(\frac{d t}{d t}\right)
\end{aligned}
$$

SAME OUTEOME AS WITM ENEREY CONSIDERATIO N

SOLUINE THE DFFEX, ERUATION
LET TNIAL SOLUTION:

$$
\begin{aligned}
& Q=Q(t=A \operatorname{cin}(\omega t+\phi): A, \omega \phi \text { cons. } \\
& d t \\
& \frac{d}{t}=\dot{d} Q=\dot{d}=\dot{Q}=\omega \cos (\omega t+\phi)
\end{aligned}
$$

SOLUTTON FF $\omega^{2}=1 / L C$
free $\left(\frac{\text { Evess }}{56 E}=H 2\right)=1 / 2 \pi \sqrt{L E}$ $A=Q_{\text {max }} \Rightarrow$ BOUNDY CONAITION


## 4ellotaty <br> $\rightarrow$ - A AREA


DISPLACEMENT CURRENT IN GAPACITOR
IS SAME AS CURRENT IN THE WIRE

[^0]
$E=\vec{E}(\vec{n}, t)=\vec{E}(x, y, z, t)$
$B=\vec{B}(n, t)$
(WYHL WOXA OFन18 30 58 NVO)

\[

$$
\begin{aligned}
& a_{6} \\
& \dot{b} \\
& a_{1}
\end{aligned}
$$
\]

$$
\begin{aligned}
& s P \cdot Z \quad{ }^{3} \Phi \\
& \text { REMEMBER: }
\end{aligned}
$$

(7VOP 10 LnSwedW00 E)
$+$



## W马yOMH1 3ONS 24En! (1 <br> $\phi \vec{A} \cdot d S=\int \vec{\nabla} \cdot \vec{d} d \vec{i}$

0.ais vologniv(e
D10.

-
$\int_{0}^{N:} \vec{\nabla} \cdot \vec{E} d T$
Cuntismos on and
r.

$$
\vec{\nabla} \cdot \vec{E}=P / \varepsilon_{0}
$$

2) $\vec{\nabla} \cdot \vec{B}=0$ EVERYWHERE:

$$
\begin{aligned}
& \phi \vec{B} \cdot d \vec{l}=\mu_{0}\left(\varepsilon_{0} \frac{\partial a_{e}}{\partial t}+i\right) \\
& \Phi E=\int_{E} \vec{E} \cdot d l \\
& i
\end{aligned}=\int_{S} J \cdot d S
$$

TREAT CASE: CARTESIAN COORDINATES


LET PHASE $=$ CONSTANT $=(\omega t-K x)$
OIFFERENTIATE:
$\omega-K \frac{M E}{t}=0$
$\frac{d x}{d E}=\frac{W}{K}=$ CONSTANT $=V$


N + N

3now
r

OL LNIOd Yata 3WIL SI 1

EXPRESS YAS $Y(x, t)$
TRUETRIAE:
$Y=A \operatorname{Ain}(\omega t-k X)$


$$
\begin{aligned}
& \begin{array}{c}
\sin \left(\omega t_{1}-k x\right)=\sin [\omega(t,+T)-k x) \\
\text { TRUE IF } \\
T=2 n \pi, \quad n=:=2,-1,012 \ldots \\
\omega=2 \pi V=2 \pi / T \\
\omega T=2 \pi \\
k=2 \pi / \lambda \\
v=\lambda / T
\end{array} \\
& \begin{array}{l}
\sin [\omega t-k x]=\sin \left[\omega\left(t_{1}+T\right)-k x_{1}\right) \\
=\sin (\omega t-k x+\omega t=-\sin (k(x-v) \\
\beta=f(x \pm v t)
\end{array} \\
& \begin{array}{l}
\quad 1-30-70 \\
\gamma=A \sin k x(\operatorname{con} A \cos k x) \\
Y\left(x^{\prime}\right)=Y\left(x^{\prime}+n \lambda\right) \\
A \sin k x^{\prime}=A \sin \left(k x^{\prime}+k n \lambda\right) \\
\sin A=\sin (A+2 n \pi) \quad n=0, I^{\prime}+2 \\
\therefore K \lambda=2 n \pi \\
\text { OR } K=\frac{3 \pi}{\lambda}=\text { WAVE NUMBER }
\end{array}
\end{aligned}
$$

$$
\begin{aligned}
& \frac{1}{3}=x \\
& 3 a \\
& 1>20 \\
& 3120
\end{aligned}
$$

$$
x
$$

$$
0
$$

51N+1sN0s Syt uq:up

$$
\sum_{n=1}\left(a_{n} \sin n t+b_{n} \cos n t\right)
$$

crastrwns wo Noillott) NOillsodmsers LELECROMAG. - ER ESSURE
Y SOUND - DENSITYLINEAR

$$
\omega=
$$

$$
\Delta \mu z=\frac{1}{\mu z}
$$ $2 \pi$



CASE AND 2 WAVES OF SAME SPEED - FREQUENCY, BUT

DEFERENT PHASE.

$$
\begin{aligned}
& Y_{1}=Y_{\text {max }}^{\sin }(1 N x-\omega t-\phi) \\
& Y_{2}=Y_{\text {max }} \sin (N x-\omega t)
\end{aligned}
$$

$\phi=$ PHASE ANGLE, CONSTANT

$$
\begin{aligned}
& \sin A+\operatorname{in} B=2 \sin \left(\frac{A+B}{2}\right) \cos \left(\frac{A-B}{2}\right) \\
& Y=Y_{1}+Y=2 r_{\max } \sin \left(k X-\cos t-\frac{\phi}{2}\right) \cos \frac{\phi}{2}
\end{aligned}
$$

$\rightarrow$ MODIFIED TRAVELAG WAVE
2-2-70

1) Ex. OF AOOITION OF 2 WAVES TRAVELoG N SAME ORECTVON. SAME

$$
\begin{aligned}
r & =\sin (k x-\omega c) \cdot \sin \left(k x_{2}-\omega t-\phi\right) \\
& =\left(2 \cos \frac{\phi}{2}\right) \sin (k x-\omega t-\phi)
\end{aligned}
$$

WHIG N IS ALSO \& TRAVELING WAVE. THE AMPLITUDE OSPENOS UPON THE PHASE ANGLE ( $=\beta$ )
2.) 5. OF $4001 T I O N$ OF 2 WAVES

TAAVIEING IN OPPOSITE DIRECTION is sAM $\operatorname{Sa}^{2}=\operatorname{in}(k x-\omega t)+\sin (k x+\omega t)$ $=2 \sin \mathrm{kx}$ cos $\omega t$ IS NOT A TRAVELING WAVE

$06-h-2$




AUERAGE POWEA PROPONTIONRL
TO I) IMPLITUDE
2) FREQ S) MASS(W) 4) SPEED
GENERALLY, ALL WAVES POWER
IS DEPENDENT ON (AMNL)
AND FREQ

$$
\therefore \frac{I_{A}}{I_{A}}=\left(\frac{\Gamma_{A}}{\Gamma_{h}}\right)^{2}
$$

$$
t=10 A T \quad P_{\text {EWER }}=\Delta V / \Delta t=10^{4} \text { WATTS }
$$

CONSERVATION OF $U \$ P$ (MOMENTUM) AFTER I DAY, ship ats POFMV

$$
P \propto U \text { OR } P=\frac{U}{C}
$$

FOR ABSORTION OR EDITION

- GEOMETRICAL OPTICS:
$\lambda \Rightarrow$ DIMENSIONS OF SYSTEM
- A

$$
\frac{A I R}{G L A S S} \quad B
$$

HOW DOES WIGHT 60 FROM A TOB?
$C_{\text {CLASS }}=\frac{1}{n}\left(C_{A I R}\right) \quad n=$ INDEX OF REFRACTION
Principle: light Rat TAKES
BATH $A-B$ SUCH THAT, COMPARE? TO NEARBY PATHS, TIME TAKEN IS A MINIMUM

$$
\begin{aligned}
& \Delta U=10^{4} \frac{\text { jOULE }}{5 E \delta}(\Delta t \sec ) \text { ( } \Delta E^{2}(\mathrm{~nm}) \\
& P_{\text {LIgHT }}=\frac{\Delta U}{C}=\mathrm{MLV} \\
& \Delta V=4 \times 10^{-5} \frac{\mathrm{~m}}{\sec }
\end{aligned}
$$



$$
\begin{aligned}
& \frac{d}{t}+\frac{0}{x}=\frac{r}{x}=0 \text { vo } \\
& \theta \tau=l+x \Leftarrow r b 0 \nabla \\
& \theta=\varepsilon+x \Leftarrow \supset 0 Q \nabla
\end{aligned}
$$

$$
\begin{aligned}
& \theta \text { BE SMALL } \\
& \theta \approx \tan \theta \\
& =\frac{a V}{r} ; \alpha=\frac{a v}{0}
\end{aligned}
$$

CONVENSIONS







$$
\therefore R=A_{R} e^{i\left(\omega t+\phi_{A}\right)}
$$

EIND MAGNETUOE OF $R .\left(A_{R}\right)$

1) FINO $A_{R}{ }^{2}$ BY MULT. ह̂̀ BY it's comprex conjugate $A_{k}^{2}=\left(A_{1} e^{i \phi_{1}}+A e^{i \phi_{2}}\right)\left(A e^{-i \phi}+A_{2} e^{\dot{\omega} \phi}\right)$

$$
=A_{1}^{2}+A_{2}^{2}+A_{1} A_{2}\left(e^{i\left(\phi_{1}-\phi_{2}\right.}+e^{i(\phi-\phi)}\right.
$$

$$
\operatorname{sinCE} e^{i x}+e^{-i x}=2 \cos x
$$

$$
A_{R}^{2}=A_{1}^{2}+A_{2}^{3}+A_{1} A_{2} \cos \left(\phi_{1}-\phi_{2}\right)
$$

$$
\cos \left(\phi_{2}-\phi_{1}\right)=\cos \left(\phi_{1}-\phi_{2}\right)
$$

I(NTENSITY) $\propto A^{2}$ (= AMPGTTVDE)

$$
3-2=70
$$

REVIEW OF INTERFERENEE:
TWO WAVES FROM COHERENT
SOURSES INTERFERE CONSTRUC= TVELY IF OPTICAL PATH DIEFERENEE I mitd WHERT M IS AN INTEGER. DESTRUCTIUE OCCURS At $\left(m+\frac{1}{3}\right)$ (180 out of pHASE)
REFLECTED WAVE MS T PHASE CHANGE IF IT TRAVESS IN MEDIUM ( $n_{1}$ ) ANO REFLECTS FROM $n_{2}\left(2 n_{1}\right)$
INTEREERENGE SOURCES


BY CHANGHE M: ONE MAY
DETERMNNE ABL SONTS OF
NEAT TMINES






$$
\begin{aligned}
& m_{s}=1 \text { GRAM } \\
& T=30 \mathrm{NTS} \\
& A=.5 \\
& B=1.0 \\
& h=\frac{-001}{T}=30^{.5} \\
& l=5 \\
& f=
\end{aligned}
$$


) MAGNETIC PROPERTIES OF MATTER-37
A) POLES AND DIPOLES

b) WITHOUT IRON INSIOE:

$$
B_{0}=\mu_{0} n i \quad\left(n=\frac{\text { UNTURLENGN }}{} n(d \ll r)\right.
$$

c) WITH IRON CORE:

1) $B=B_{0}+B_{M}$
2) $B_{M}$ :MAGNETIC INDUCTION DUE TO SPECIMIN
3) $B_{m} \propto M$
d) DOMAINS - LOCAL REGIONS IN WHICH

THERE, 15 PERFECT ALI GNMENTS
e) HYSTEAESIS - NOT RETRACABLE, SUCH AS ROWLAND RING
G) NUCLEAR MAGNETISM
1)
a) $J_{p}=\mu B \sin \theta$

b) $\omega_{p}=\frac{\mu_{B} B}{L_{p}}$

1) $\mu_{p}=1.4 \times 10^{-.26}$ AMD.M2
2) $L_{p}=.53 \times 10^{-34} \mathrm{~J}-5$
3) WITH FIELD BOSE $\perp B$ $\alpha$ IS VERY SMALL
a) $\omega_{0}=\frac{\mu \beta}{L \beta}$
b) $\therefore \mu_{p}=1.410 \times 10^{-36}$ AMp= $\mathrm{m}^{2}$
H) THREE MAGNETIC VECTORS
4) $M$ - MAGNETI EATON $\cdot d_{\omega}=M d l$
a) $($ AdP $)=$ VOLUME OF SLICE
b) DEFINED AS THE MAGNETIC MOMENT DER UNIT VOLUME OF THE CORE MATERIAL
5) $\vec{B}$ - MAGNETIC INDUCTION
a) IN ROW LANA RING, WITH NO CORE; $\oint B \cdot d A=\mu_{0} i$
b) $O R: B\left(2 \pi r_{0}\right)=\mu_{0} N_{0} i_{0}$ WHERE $r_{0}=M E A N$ RADIUS

c) $\therefore$ AMPEERE'S LAW (a) IS NOT VALIO WHEN MAGNETIC MATERIALS ARE PRESENT. THEREFORE MUST BE MORE CURRENT.
$\oint B \cdot d i=\mu_{0}\left(i+i_{m}\right)$
OR $B\left(2 \pi r_{0}\right)=\mu_{0}\left(N_{0} i_{0}\right)+\mu_{0}\left(N_{0} i_{m a}\right)$
d) FOR ROWLAND RING:

$$
\phi\left(\frac{13-\mu o k}{\mu_{0}}\right) \cdot d \&=i
$$

3) H-MAGNETIC FIELD STRENCTH $=\frac{B-N_{0} M}{W_{0}}$
a) AMPERES LAW BECOMES $\oint H \cdot d l=i$ ( $i$ IS TRUE CURRENT, NOT MAGNETISING CURS
4) RELATIONS BETWEEN THESE VECTORS:
a) $B=K_{m} \mu_{0} H$
5) $\left(K_{m}=\right.$ PERMEABILITY CONSTANT)
6) FOR PARA $\frac{1}{3}$ DIAMAGNETIC MATERIALS
b) $M=\left(K_{m}-1\right) H\left(I N\right.$ VACUUM, $\left.K_{m}=1\right) \Rightarrow B=\mu_{0} H$
c) Km $M=0$
7) paramagnetic $>1$
8) OIAMACNEFIC<I
9) TABLE (SUMMARY)
a) IMAGNETIC INDUCTION-B-ALL CURRENTS - HAS QOUNDRY ${ }^{2}$ 2) " FIELO STR. M-TRUE " -TAG. GOMP. CONT. 3) MAGNETIZATION -M -MAG CURR - $=0 \mathrm{NN}$ VACUUM
b) $D E F$. of $B$ : Bo FiquxB=il $\times B$
10) $B=\mu_{0} H+\mu_{0} M$
11) AMPERE'S LAW- $\varnothing H \cdot d \ell=i$ (TRUE CURRENT)
12) $B=k=m_{0} \mu_{0} t$
13) $M=\left(K_{m}-1\right) H$

$$
\begin{aligned}
& \text { RING } \\
& \text { \& B }
\end{aligned}
$$



$\overparen{H}$
C) ELECTROMAGNETIC OSCIGLATIONS-QUANITATINE

1) OERIVATION OF FRERUENCY FUNCTION
a) $u=U_{\theta}+U_{E}=\frac{1}{2} L i^{2}+9 / 2 C$
b) $d v / d t=(d / d t)\left(\frac{1}{2} L i^{2}+9 / 2 c\right)=$
$L i d i / d t+9 / 6 d g / d t=0$
(ASSUMING $R=O \Rightarrow U=$ CONSTANT
c) $i=d q / d t \rightarrow d i / d t=d^{2} q / d t^{2}$
d) $L^{d^{2}} / d t^{2}+\frac{1}{6} q=0$
(ANAGAGOUS TO $m^{d^{2}} x / d t^{2}+k x=0$ FOR MASS-SPRING SYSTEM)
e) SOLUTION OF OIFFERENTIAL EQUATION YIELDS:
$q=q_{\text {max }} \cos (\omega t+\phi)$
2) $\omega$ = ANGULAR FREQUENCY
3) $\phi$ = ARBITRARY PHASE ANGLE
4) FORMULAS FROM DERIVATION:
a) $d q / d t=i=q_{\text {max }} \omega \sin (\omega t+\phi)$
b) $d^{2} q / d t^{2}=d / d t=q_{\text {max }} \omega^{2} \cos (\omega t+\phi)$
c) SUBSTITUTION OF ORIGINAL

DIFEERINTIAL EQUATION YIELDS:
$\omega=\sqrt{1 / L C}$
d) $t=\frac{\pi}{4 m}$
e) $u_{0}=\frac{1 q^{2}}{2 c}=\frac{q^{2}}{2 c} \cos ^{2}(\omega t+\phi)$
f) $U_{b}=\frac{1}{2} L i^{2}=\frac{1}{2} L w^{2} q_{\text {and dir }}{ }^{2}(w t+\phi)$
3) FOR LRC GIRGUIT
$q=9 m e^{-R t / 2 L} \cos \omega^{\prime} t$
(IWITITALGY 9 99max)

2) MECHAN V ELECTRO $\rightarrow E$ VSF

$i_{m}=6 m / R$
LUMPEO AND DISTRIBUTED ELEMENTS

1) EXAMPLES
Q) LUMPED-MASS STRING SYSTEN.
1)K.E. ALL IN MASS
I U AGG IN SPRING
b) DISTRIBUTED-IN ORGAN PIAE
CVARIATIONS IN AIR DENSITY)
2) IN A SMALL AREA OF GAS:
UN $=\frac{1}{2} P V F 2$
a) UK POLENTIHL ENERGY PER VOLUME
b) VGEVELOCITY OF GAS

6

G) ANALOGOUS PLUG AND CHUG TABLE

1) LUMPED SYSTEMS:

MECHANICAL SYST. ELECTROMAGNETIC SYST.
(MASS $\frac{1}{3}$ STRIA)

$$
\begin{aligned}
& U_{k}=\frac{1}{2} m v^{2} \\
& U_{p}=\frac{1}{2} k x^{2} \\
& \omega=\sqrt{k / m}
\end{aligned}
$$

(LC CIRCUIT)

$$
\begin{aligned}
& U_{B}=\frac{1}{2} L i^{2} \\
& U_{E}=\frac{1}{2}(1 / C) q^{2} \\
& W=\sqrt{1 / L C}
\end{aligned}
$$

2) DISTRIBUTED SYSTEMS

MECHANICAL SYST.
(ACOUSTIC CAVITY)

$$
\begin{aligned}
& U_{K}=\frac{1}{2} \rho_{0} V V_{B}^{2} \\
& U_{p}=\frac{1}{2} B\left(\Delta p / \rho_{0}^{2}\right. \\
& \omega_{1}=\frac{3.14}{2} \\
& V=\sqrt{B / p_{0}}
\end{aligned}
$$

ELECTROMAGNETIC SYST.
(ELECTROMAGNETIC CANTY)

$$
\begin{aligned}
& U_{B}=\frac{1}{2 \mu_{0}} B^{2} \\
& U_{E}=\frac{1}{2} \epsilon_{0} E^{2} \\
& \omega_{1}=\frac{1,19}{a} \\
& C=\sqrt{1 / \varepsilon_{0} \mu_{0}}
\end{aligned}
$$

H) INDUCED MAGNETIC FIELDS

1) A CHANGING E FIELD PRONUCES A B FIELCO
2) 2$) \oint E \cdot d l=-d \Phi_{B} / d t$
b) $\oint B \cdot d l=\mu_{0} \varepsilon_{0} d I_{\varepsilon} / d t$
3) AMPERE K LAW: $\oint B \cdot d l^{e} \mu_{0} i$
a) $\therefore$ ENTIRE $\phi B \cdot d i=\mu_{0} \varepsilon_{0} d \Phi_{E} / d t+\mu_{0} i$
b) or $\oint B \cdot d l=\mu_{0}\left(i_{d}+i\right)$
I) DISPLACEMENT CURRENT ( $i d$ )
4) $i_{d}=\varepsilon_{0} d \Phi E / d t$
5) ACROSS CAPACITOR
a) $E=9 / \varepsilon_{0} A \Rightarrow \frac{\partial E}{\frac{t}{t}}=\frac{1}{\varepsilon_{0} A} \frac{d a}{d t}=\frac{1}{\varepsilon_{E} A} i$
b) $i_{d}=\varepsilon_{0} \frac{d \Phi^{F}}{d t}=\varepsilon_{0} \frac{t}{t}(E A)=\varepsilon_{0} A \frac{d E}{d t}$
c) $i_{d}=\varepsilon_{0} A\left(\frac{1}{\varepsilon_{0}} i\right)=i$

J）Maxwell＇s Equations＋Cavity Oscillations
1）

a）FOR SYSTEM：$\oint E \cdot d l=h E(r)$
WHERE E（r）IS THE VALUE OF EATR．
b）$E=0$ ON CAVITY WALLS $(E \perp d l)$
c）$E(r)=-\frac{1}{h} d$ 旦 $B / d t$
d）$E$ is Max WHEN $B=0$
e）$B(r)=\frac{\mu_{0} E_{0}}{2 \pi r} d$ I昱 $d d t=\frac{\mu_{r}}{2 \pi r} i_{d}$
2）THIS B－E interplay occurs in traveling ELECTROMAGNETIC WAVES SUCH AS RAOIO 髙 VISIBLE LIGHT RAYS
3）$B=0$ OUTSIDE CAVITY
III) ELECTROMAGNETIC

WAVES
A) COAXIAL CABLE (TRANSMISSION LINE)
a)

b) WHEN S IS ON B, THE CENTRAL 5 OUTER SURFACE HAVE SANE
c) WHEN 5 IS ON G, THERE IS A POTENTIAL BETWEEN CORE $⿻$ C SURFACE
d) THIS POTENTILE TRAVELS

ALONE LINE ATE
DAT DISTANCE A ALONE CINE

3) ATt $=t$

e) If 5 ALTERNATES BETWEEN - 4. NO b

f) $5,3, b$ REPLACE BY SIN SOINL GENERATOR

B) a traveling wave in Resistivesless TRANSMISION LINE: $\lambda=S / \mathrm{N}$
a) COMMERCIAL; $V=60$ CYCLES $/ \mathrm{sEC}$
b) 1 (COMMERCIAL) $=3000$ MILES
c) HAS NO RESONATE FRELUENCY
C) COAXIAL CARLE-FIELOS C CURRENTS

1) $i_{d}=\varepsilon_{0}$ IE/ ft $\Rightarrow i_{d}$ IS MAX WHEN $E=0$
2) $B=\mu \cdot i / 2 \pi N$ FOR COAx-CABLE
3) RELATIVE MAGNETUDES
a) E $B$ ARE IN PHASE IN TRAVELING WAVE-REACH MAX.
AT SAME DINT
b) E B ARE $90^{\circ}$ OUT OE PHASE IN A STANDING WAVE
D) WAVEGUIDE
4) ד\#4 +
a) $V_{P H}$ (EPHASE SPEED) $=c / \sqrt{1-(2 / 2 q)^{2}}$
b) $V_{\text {gr }}(=$ GROW SPEED $)=c \sqrt{1-(M / 2 a)^{2}}$
c) As $a \rightarrow \infty, V_{P H}=V_{f r}=c$
5) GUIDE WAVELENCTA ( $=\lambda$ g)
a) $\lambda_{f}=V_{\mathrm{PH}} / \mathrm{V}=V_{\mathrm{Pm}} / \mathrm{cId}=\lambda_{8} \mathrm{~V}_{\mathrm{man}} / \mathrm{c}$
b) $\lambda_{g}=\frac{1}{\sqrt{1-(\lambda / 2 a)^{2}}}$
c) THE GUIDE ATTIENN IS THAT EXIBEDED BY CERTAIN FIELD PATTERNS. IT IS LARGER THAN EREE-SPACE $\lambda$
E) RADIATION
6) G -ELECTRIC DIROLE WHOSE V AND THUS P. VARY SIMOSOIDALLY WITH TIME
a) ELECTRIC LINES OF FORCE BREAK OFF IN CLOSED LOOPS. AND GO INTO SPACE C.
b) THELSE ARE ELECTROMAGNETIC WAVES, RADIATION
7) FACTS $\$$ JUNK
a) $c=v \lambda \Rightarrow c=\omega / k$ ( $K=$ WavE \#)
b) $\omega=2 \pi V ; \quad k=2 \pi / \lambda$
F) TRAVELING WAVES + MAXWELLS EQUATIONS
8) IN AN ELECTROMAGNETIC WAVE
a) $B=B_{m} \sin (k x-\omega t)$
b) $E=E m \sin (k x-\omega t)$
9) $\frac{d E}{d x}=\frac{-d \beta}{d t}$
10) $w / k=E_{m} / B_{m}=C \Rightarrow E=c B$
11) $\frac{-\partial \theta}{\partial X}=\mu_{0} \varepsilon_{0} \partial \varepsilon / \partial t$
12) $6 \mathrm{~m} / \mathrm{Bm}_{m}=k / \mu_{0} \varepsilon_{0} \omega=1 / \mu_{0} \varepsilon_{0} C$
13) THE POYNTING VECTOR (s)
14) $S=\frac{1}{\omega_{0}} E \times B$ (MKS $S-\frac{\frac{\omega 4 T T}{m^{2}} \text { ) }}{1+5}$
15) ENERY MOVES IN DIRECTION OF $\stackrel{S}{S}$
16) $d U / d x$ in A BOX ( $(d x) A=d$ VoLUME)
17) $d U=\left[\frac{1}{2} \varepsilon_{0} E C B+\frac{1}{2 \mu_{0}} \frac{B E}{C}\right] A d x$
b) $=\left[\frac{1}{2} \varepsilon_{0} E^{2}+\frac{1}{3 \mu_{0}} B^{2}\right] A d x$
c) $=E B A / \mu_{0} C$
aid
THE DIFFERENTIAL FORM OF MAXWELL'S FIECTR OMA

USUALGY UNIKNOWN.
IS MARO TO COMPUTE
IONS IN GENERAL CASES
$m=\int \rho d T$
$5 s \mathrm{bW}=m$
$\operatorname{tog}=0$ मLISN3a $=0$ (z
sif
c) IT 15 DESIREABLE TO
18) $T=$ VOLUME
 MAXWELL'S EQUATIONS TO
$\overbrace{0}^{\circ} \mathrm{Br}_{0}$
4yo
$\frac{4}{0}$ 2
2
2
2
$\frac{1}{2}$
$\frac{1}{2}$
2
THE OPERATBR DEL

4
$L^{\prime \prime}$
4
$k E_{\mathrm{e}}=$

$1 N 31$
$(a 731$
923
$73 a$
1 SnHI
tn03
$M 1(E$
$353 H \perp$
1 aN
aNW 3
QUATIONS
IAVE EQUA
M MAXWELL
19) E $\oint$
20) $\oint B$.
21) $\oint B$.
22) $\& E$.
23) E AN
$\rightarrow$
24) USING ELECTROST
a) $E=i E_{x}+j E_{y}$ b) $E=-\nabla V$
25) APPLICATIONS
©


II) WAVES IN ELASTIC MEDIA
A) TYPES OF WAVES (MECHANICAL)
1)a) TRANSVERSE-MOTION OF SYSTEM IS $\perp$ TO DIRECTION OF WAVES
b) TUDINAL-MOTION OF SYSTEM IS BACK + FORTH (SPRING)
26) OIMENSIONALLY
a) 1 OMEN. - (SPRING)
b) 2 OMEN: - (WA TEN)
c) 3 OMEN- (SOUND \$ LIGHT) PROPOGATE IN WAVE FRONTS
27) BY SHAPE (3D)
a) PLANE WAVES- $\|\|\|$
b) PASPHERICAL WAVES:
B) EQATIONS \$ JUNK
28) $y=y_{\max } \sin \frac{2 \pi}{\lambda} x$

at a GIVEN TIME
29) AT TIME $T$

$$
Y=Y_{\text {MAX }} \operatorname{Aim} \frac{2 \pi}{\lambda}(x-v t)
$$

3) $T=$ PERIOPETIME FOR WAVE TO

$$
\text { TRAVE G ON E } \lambda \text { ) }=\lambda / V
$$

4) $\therefore y=Y_{\text {max }} \sin 2 \pi\left(\frac{x}{\lambda}-\frac{t}{T}\right)$
5) $k(E$ WAVE \# $=2 \pi / \lambda$
6) $W(=A N G U G A R$ FREQUENCY $)=\frac{2 \pi}{T}$
$\Rightarrow Y=Y_{\text {Max }} \sin (k x-\omega t)$
$\Rightarrow$ FOR A WAVE TRAVELING TO RIGHT
7) $Y=Y_{\text {max }} \operatorname{lin}(k x+\omega t)$

FOR A WAVE TRAVELING TO LEFT
9) $V \in$ PHASE VELOCITY) $=\frac{\lambda}{T}=\frac{W}{R}$
10) AT $E=0$. $\Phi 0$ -
$Y=Y_{m} \sin (k x-\omega t-\beta)$ FOR WAVE
TRAVELING TO THE RIGHT
II) AT A FIXED POINT. SUCH AS $x=\frac{\pi}{R}$ $Y=Y_{M} \sin (\omega t+\phi)$
c) THE SUPERPOSITION PRINCIDLE -

1) OEFN - TWO OR MORE WAVES CAN TRANSVERSE THE SAME SPACE INDEPENDENTLY OF ONE ANOTHER
2) ANY PERIODIE FUNCTION MAY GE EXPRESSED BY THE FOURIER SERIES:
$Y(t)=A_{0}+A_{1} \operatorname{din} \omega t+A_{3} \sin 2 \omega t .$.
$+B_{1} \cos \omega t+B_{2} \cos 2 \omega t \ldots$
a) $W=2 \pi / T$
b) A'S \$ BIL CONSTAN'S ACCORDINE TO THE FUNCTION
D) WAVE SPEEO-V
3) $V=\sqrt{F / W}$ is A STRETCHED STRING
4) $\lambda=V / y$
a) $f=$ FREQUENCY
b) $\mu=$ MASS/ UNIT LENGTH
E) POWER $\$$ INTENSITY IN WAVE MOTION
5) $P=(-F O Y / \partial x) \partial Y / \partial t$

FE TENTION IN THE STRING
2) FOR A SINE WAVE IN A STRINE
a) $P=Y_{m}{ }^{2} k \omega F \cos ^{2}(k x-\omega t)$
b) $\bar{\rho}=\frac{1}{T} \int_{t}^{t+T}$

$$
\begin{aligned}
P d t & =\frac{1}{2} Y_{m}^{2} k \omega_{F} F=2 \pi^{2} k^{2} V_{2} \frac{F}{V} \\
& =2 \pi^{2} Y_{m}^{2} V V^{2} \mu V
\end{aligned}
$$


$\xrightarrow[4]{2}+\frac{3}{2}$ GRAPHICALLY.

$$
\begin{aligned}
& 2 \text { mToDXY UNT Wh }=\lambda \\
& \text { S } 3 \wedge \forall M \text { ONIGNH }
\end{aligned}
$$

$$
\text { SBA } \because M \text { ONIONHLS }
$$

$$
\operatorname{an}-1+\log
$$

$$
\begin{aligned}
& \text { WAVES } \\
& \text { sin } k x
\end{aligned}
$$

उAVM
4

II) SOUND WAVES
A) CLASSIFICATIONS

1) ACCOROINE TO FREQUENEY
a) BELOW 20 HE - INFRASONIC
b) $20-20,000 \mathrm{HE}$ - AUDIBLE
c) ABOVE 20,000 HZ-VLTRAS NI C
2) TYPE OF WAVE
a) TRANSVERSE-WAVE MOUES
$\perp$ TO BODY
b) LONGITUOINAL-(SOUND) - WAVE

MOVES II TO BODY

- P) ROPICATION F SPEED OF LONGITUDINAL WAVE

1) $V(=$ SPEED IN 1 DIMENSIONAL SYSTEM)

$$
=\sqrt{B} p_{0}
$$

a) $B=$ BULK MODULUS OF ELASTICITY) $=-\nabla \triangle P / \Delta V$
b) $p_{0}=$ DENSITY OF FLUID
2) IT IS POSIBLE TO EXPRESS $B$ Ls THE IN TERMS OF INITITIAL PRESSURE

$$
V=\sqrt{\gamma P_{l} / \rho_{0}}
$$

( $\gamma=$ CONSTANT = RATIO OF SPECIFIC
HEATS FOR THE GAS)
C) TRAVELING LONGIT UDINAL WAVES

1) FOR A WAVE TRAVELING TO RIGHT
a) $Y=Y_{m} \cos \frac{2 \pi}{\lambda}(x-v t)$
b) $y=Y_{n} \cos (k x-\omega t)$
c) i) $p=B d y / d x$ (PRESSURE)
2) $\frac{d y}{d x}=-k Y_{m} \sin (k x-\omega t)$
3) $\therefore p=B K Y_{m} \sin (k x-\omega t)$
4) $P$ ( = PRESSURE AMPLITUDE) $=k P_{O} V^{24 m}$
a) $\therefore p=P \sin (k x-\omega t)$
b) OISPLACEMENT IS $90^{\circ}$ OUT OF PHASE WITH.
D) VIBRATING SYSTEMS \$ SOURSES OF SOUND
5) FIXEO STRING CAN RESONATE AT FREQUENCIES

$$
f_{n}=\frac{n v}{2 L}=\frac{n}{2 L} \sqrt{F / \mu} ; n=1,2,3 \ldots
$$

a) $V$ IS SAME FOR ALL FREQUENCIES
b) $f=\sqrt{\frac{E}{\mu}} \frac{1}{2 L}$ is CALLEO FUNOIMENTAL
FREQUENCY
2) FOR "ORGAN" TYPE PIPE (RES. FREQ)

$$
f_{n}=\frac{n v}{2 e} \quad n=1,2,3 \ldots
$$

a) $V$ = SPEED OF LONGITUDINAL WAVES
b) $n=$ \# OF HALF $\lambda$ 's
c) $\ell=$ LENGTH OF THE COLON
d) EVEN NUMBEREO HARMONIES ARE NOT PRESENT WHEN ONE END OF THE PIPE IS CLOSED
E) BEATS

1) OCCURS WHEN 2 SLIGHTLY DIFFERENT F ARE SOUNDED SIMULTANEOUSLY
2) QUANTITATVVELY $\left(Y_{m_{1}}=Y_{m z}\right)$
a) $Y_{1}=Y_{m} \cos 2 \pi+t ; Y_{2}=Y_{m} \cos 2 \pi+t_{2} t$
b) $Y=2 Y_{m} \cos 2 \pi\left(\frac{t-1}{2}\right) t \cdot \cos 2 \pi\left(\frac{f_{2}+2}{2}\right) t$
c) AMPLITUOE VARIES WITH TIME:

$$
f_{\min }\left(f, f_{2}\right) / 2
$$

d) RESULTINE VIBRATION'S FREQVENCY:

$$
f=\left(f_{1}+f_{2}\right) / 2
$$

e) BEAT HAS MAX AMP WHEN:

$$
\cos 2 \pi\left(\frac{f-f_{2}}{2}\right) t=1_{1}
$$


LIGHT
WAVELENGTH
$E_{0} \xi$

NATURE


目

VII) REFLECTION AND REFRACTION
A) GENERAL LAWS

1) THE REFLECTED $\ddagger$ REFRACTED RAYS LIE IN THE PLANE FORMED DY THE INCIDENT RAY.
2) FOR REFLECTION:
a)

b) $\theta_{1}=\theta_{1}^{\prime}$
3) FOR REFRACTION:
a)

b) $\frac{\sin \theta_{1}}{\sin \theta_{2}}=n_{21}$ (INDEX OF REFRACTION)
(1) IN COMPARISON WITH VACVUM,

$$
n>1
$$

(2) $n$ VARIES SLIGHTLY WITH $A$
B) HUYGEN $\frac{1}{5}$ LIGHT

1) HUYGEN'S PRINCIPLE: ALL POINTS ON A WAVEFRONT CAN BE CONSIRED AS POINT SOURCES FOR THE PRODUCTION OF SPHERICAL SECONDARY WAVELETS. AFTER TIME $\tau$ THE NEW POSITION OF THE WAVEFRONT WILL BE THE SURFACE OF FT TANGENCY TO THESE SECONDARY WAVELETS


$$
\lim _{1 \rightarrow d}
$$

dub

$$
+x_{4}+\frac{1}{x}+\frac{1}{0} \pi
$$

4) WORKING GEOMETRICALLY:
a) MIRROR REFLECTS DIRECTLY AT ITS "Y AXIS" CENTER
b) A RAY REFECTED, WHICH WAS II TO AXIS, WILL GO THRU FOCAL POINT
c) A RAY PASSING THRU THE FOCAL POINT PASSES REFLECTS I/ FHRU TO AXIS
5) LATERAL MAGNEFICATION M

$$
m=-i / 0
$$

a) $m \angle O \Rightarrow$ IMAGE IS INVERTED
b) $m L O \Rightarrow$ IMAGE IS ERECT
D) SPHERICAL REFRACTING SURFACE

1) $\frac{n_{1}}{0}+\frac{n_{2}}{l}=\frac{n_{2}-n_{1}}{r}$
2) SIGN CONVENTION
a) $i$ IS POSITIVE IF THE IMAGE (REAL) IS ONTHE R- SIDE OF THE REFRACTING SURFACE. IS NEG. IE IMAGE (VIRTUAL) IS ON VIR - SIDE
b) $r$ IS POSITIVE NF THE CENTER OF CURVATURE LIES ON THE "REAL"SIOE, AND IS NEG. IE C. OF.C. IS ON "VIRTUAL" SIDE
c) OBJECT virtual sloe
THIN LENSES

SINE CONVENTIONS
$\vec{F}_{2}^{\prime}$

$$
\begin{aligned}
& (n-1) \\
& \text { NTIONS }
\end{aligned}
$$



$$
\left(\frac{1}{r^{2}}-\frac{1}{n_{n}}\right)
$$

REAL) LIES ON REAL SIDE
Y REAL SIDE
FORMULA)

$$
\begin{aligned}
& 2 \\
& 0 \\
& w_{j} \\
& j i \\
& 3 i \\
& m i \\
& m
\end{aligned}
$$

Hedmes (tadoctity)

 ponsfir by dinect nubteplicatmon, Eanto by tano that
varely by explict moultppacution, wa above, the Weatr TOENTHYY

$$
\begin{aligned}
& A \times(\vec{B} \times C) \times(A \cdot C) \frac{B}{B}-(A \cdot B) C^{+} \\
& {\left[=\frac{B}{B}(A \cdot C) \quad C \quad\left(\frac{B}{B}\right)\right.}
\end{aligned}
$$

Noter if $A=\bar{B}=\bar{\eta}$, une of Han identht proven the whation fon aul cunt E waed un deniontion of electampagmetor wane aguatim.

By daket mult. We neow fon exarmplo:
quen $p=5 k+2 \mathrm{~b}, \mathrm{~b}=-\mathrm{b}+2 \mathrm{k}$
them

$$
\begin{aligned}
& 5 x+\infty=(5 x+2 n) \times(-3+3 k) \\
& =5 \mathrm{~B} \times(-\boldsymbol{J})+2 \mathrm{~J} \times(-\hat{J})+5 \mathrm{a}+3 \mathrm{~B}+\mathrm{n} \\
& =-5 b+0+5(-\infty)+b a \\
& =6 \hat{b}+15 \hat{b}
\end{aligned}
$$

Pet peothet

$$
A \cdot \vec{B}=A B \cos a
$$

C

Shaw that







 unc eanply pacom.

 How form the scalam (dot) penstants phacalad watap equp above:



Prtagatrag avex a ublemen ab

A + + eperge





CHAPT. 37
1)

$$
\begin{aligned}
& H_{C}=6.4 \times 10^{21} \text { AMP-M } \\
& \mu=N i A \\
& i=\frac{L}{N A} ; N=1 \\
& =\frac{\operatorname{CA}^{4} \times 10^{3} A M=M^{2}}{=\left(6.32 \times 10^{2} m\right)^{2}} \\
& =.0494 x 10^{5} \text { AMDS } \\
& =4.94 \times 10^{7} \text { AMRS }
\end{aligned}
$$

4) a)


$$
\begin{aligned}
B_{a} & =\mu_{0} n i \\
i & =\frac{B_{0}}{\mu_{0}} \\
& =\frac{2 \times 10^{-4} \text { WEQEe-m}(2 H)(.055)}{m^{2}\left(1.26 \times 10^{-6} H E M R Y\right)(400)} \\
& =.00137 \times 10^{2} \text { WEBER} \\
& =.137 \text { AMPS }
\end{aligned}
$$

b) $18 m=40010$

$$
\begin{aligned}
& B_{0}=2 \times 10^{-4} \frac{\text { WEBER }}{m=} \\
& \therefore B=802 \times 10^{-2} \frac{W^{2} E S E}{m^{2}} \\
& R_{g}=8 \Omega \quad N_{3}=50 \\
& B=\frac{1 R}{N A} \Rightarrow q=\frac{B N A}{R} \\
& q=\frac{W h y e m^{2}\left(8.02 \times 10^{-2}\right) T(.055)^{2}(50)}{m^{2}} \\
& =.0477 \times 10^{-2} \text { WEEER } \\
& =4.77 \times 10^{-4} \text { coulomes }
\end{aligned}
$$

F) $U_{p}=108 \times 10^{-3} A M P-M^{2} \quad L=5 \mathrm{~cm}, A=1 \mathrm{Cm}^{2}$

$$
\begin{aligned}
& =4.29 \times 10^{2} \text { 4MP M2 }
\end{aligned}
$$

$b b_{g}=7.6$ AMP-M ${ }^{2}$

$$
\begin{aligned}
& T=4 \times B \\
& =(7.6 \text { Anp-12 })\left(15 \times 10^{4} \text { GAuss }\right) \\
& =114 \text { meserronhos }
\end{aligned}
$$

CHAPT 19; PE 492

1) $Y=Y_{m} \sin (k x-w t)$
a) $Y=Y_{m} \sin k\left(x-\frac{\psi}{k} t\right)$

$$
=Y_{m \min } k(x-v t)
$$

b)

$$
\begin{aligned}
y & =Y_{m \operatorname{Lin}} \omega\left(\frac{k}{\omega} x-t\right) \\
& =Y_{m} \operatorname{An} \omega\left(\frac{x}{V}-t\right)
\end{aligned}
$$

c)

$$
\begin{aligned}
Y & =Y m \operatorname{In} 2 \pi\left(\frac{k}{2 \pi} x-\frac{u}{2 \pi} t\right) \\
& =Y_{m} \operatorname{Lin} 2 \pi(X A-V t)
\end{aligned}
$$

d) $y=Y_{m} \operatorname{Lin} 2 \pi(A, A)$
2)AFOR VIS. L16HT $\quad 4 \times 10^{-7} m<\lambda<7 \times 10^{-7} m$

$$
\begin{aligned}
& V_{\text {LIOHT }}=3 \times 10^{5} \frac{\mathrm{~m}}{\mathrm{SEC}} \\
& V K=\mu \quad \ll A=2 \pi \\
& \frac{\omega}{\nu}=\frac{2 \pi}{A} \\
& \begin{array}{r}
\frac{2 t}{V}=\frac{\frac{\partial}{y}}{V}=\frac{2 t}{y}
\end{array} \\
& \hat{f}=\frac{2 \pi\left(3 \times 10^{8}\right)}{4 \times 10^{-7}}=4.71 \times 10^{15} / 2 \pi 4= \\
& f_{4}=\frac{2 \pi\left(3 \times 10^{5}\right)}{7 \times 10^{2}}=2.69 \times 10^{15} / 2 \pi \frac{2}{2}
\end{aligned}
$$

b)

$$
\begin{aligned}
& 1.5 \times 10^{6}<\&<3.00 \times 10^{6} \\
& \frac{f}{v}=\frac{1}{\lambda} \\
& \lambda=\frac{V}{f} \\
& \lambda_{0}=\frac{5 \times 10^{2}}{1.5 \times 10^{6}}-2 \times 10^{2}=200 \mathrm{~m} \\
& A_{1}=\frac{3 \times 10^{8}}{3 \times 10^{8}} \\
& 5 \times 10^{-9}<\lambda<10^{-11} \mathrm{~m} \\
& f=\lambda_{3} \\
& 5=\frac{3 \times 10^{5}}{5 \times 10^{4}}=1.67 \times 10^{17} \\
& f_{1}=\frac{3 \times 10^{5}}{10^{-11}}=3 \times 10^{10}
\end{aligned}
$$

c)
5)

$$
\begin{aligned}
& f=500+z ; \quad v=350 \frac{\mathrm{~m}}{\mathrm{mec}} \\
& \lambda=V / f=\frac{35}{50}=.7 \mathrm{~m} \\
& \left(\frac{60}{360}\right) \cdot 7=117 \mathrm{~cm} \\
& \lambda=V / 5=-7 \\
& V t=x \Rightarrow(350)\left(10^{-3}\right)=35 \\
& \frac{.35}{7}=.5=\frac{\text { मHASE }}{360} \\
& \text { PHASE }=180^{\circ}
\end{aligned}
$$

14) ATMEt

$$
\begin{aligned}
& Y_{1}=-03 \sin \left(k_{1} x-\omega t\right) \\
& Y_{2}=.04 \sin \left(k_{2} x-\omega t+\frac{\pi}{2}\right.
\end{aligned}
$$

CHAPT.
18)

$$
\begin{aligned}
& f=\frac{n v}{f=} \\
& e=\frac{n v}{\frac{2}{2}}=\frac{n(30)}{2(660)}=\frac{n}{4} \\
& \frac{1}{2}, \frac{3}{4} 1 \mathrm{ME}
\end{aligned}
$$

20) $\eta$

$$
f=\frac{v}{h^{2}}=\frac{v}{2 f}=\frac{331}{2(300)}=550 \mathrm{~m}
$$



11
 SMALL MAGNET ERGO, THE MAGNET COULD
THEORETICAL BROKEN DOWN INTO N ATOMC MAGNETS, WITH A rotl OF $2 N$ POLES

$$
\begin{aligned}
& \mu=01 \text { AMP-M }{ }^{2} \\
& \Phi_{B}=\oint B \cdot d S=0 \text { rags } \\
& =\delta(\beta \cdot Q d b d w)
\end{aligned}
$$


$70^{\circ} ?$
20
BECAUSE THE HUGER THE TEMP THE MORE ATONIC VIBRATION WHICH (IN TURN FURTHER) RANDONIEES THE Oration of ATOMIC DIPOLES.


THE CURRENT IN THE COIL SETS UP A MAGNETIC FIELD, WHICH IN TURN Y LINES UP THE "CELL "DIPOLES WITH THE ATOM.

b) Explain what occurs in the tron for enterons meat IN THAT THERE NS A FINE NUMBER OF CELLS WITHIN THE IRONANINCREASED MAGNETIC FIELD FROM THE INCREASED CURRENT IN THE WINDINGS ONLY HAS THOSE FEW UNLINED CELLS TO ALICNENG
c) On this plot skeren what might be observed th the tron were

WOOD BEING DIAMAGNETIC. WOULD
HAVE NEARLY A CONSTANT B. FOR ANY CURRENT, FOR THERE ARE NO "CELLS" WHICH CAN BE PERMANETLY $A L 1 C N E O(B=C O N S T A N T ; B=O ; B L O) B_{0}>0$



 a cote or
a) pacamaguetre mater eat.

SLIGHTLY INCREASE

SLIGHTER DECREASE

SINCE MOL.
a paramagnetic material (0) (\%s
COMARLD TO A VERY WEAK K FERROMAGNETIC) WOULD INCREASE M, THUS INCREASE L. THE SAME REASONING (WINE WOULD LEAD ONE TO THE FACT, THAT A DIAMAGNETIC MATERIAL, MAGNETICALLY OPPOSED TO A PANAMAS METIC MATERIAL) WOULD SLIGHTLY DECREASE M THUS L


THE NNOUCED CURRENT IN EASE 2 WOULO BE \&REATAR RELATVELY, THE CHAGEE N CNSE THOQ WOULQ LESSN THE INDUEED CURRENT
3) Prog

3

$$
\begin{aligned}
& \left.+j\left(3 x^{2} y^{2}\right) \oplus\right)^{2}\left(x^{2} y\right)
\end{aligned}
$$

chesk rigns
5pa 5) Compactertighivergenee or the answer to probiem 4
0


 express the princhple of corservathon of onergy?

$$
\phi S \cdot d A=-\frac{d U}{d t}
$$

IT SAYS IN EFFEEF THAT THE AMOVNT
19 OF ENECY LEAVINE THE VOLUME
( $=-04$ ) is THE AMGUNT OF ENERGY
THE FORM OF ELECTROMAGMETIC
WHVES ERGO THE
ENERY OERLETION MNSIDE THE
VOLUME IS EQUAL TQ THE
ANOUNR EMITTEO

BoB MARKS 40 AVE $=47$
 PART OE TUE WAVE WILC BE REFBECTED, WHE RENANUEO WLB GO ON NTO THE DEMSER RORE W/IEM A SMALER MMPLTUDE, ANG A DECREMSED FREG
 REELEGED THE AMATY NCREASE

 $y=A_{1} \sin (x-v t)+A_{2} \sin (x+v t)$ with $k=1$ $\frac{d y}{d x}=A_{1} \cos (x-v t)+A_{x} \cos (x+v t)$

$$
\begin{aligned}
& \frac{\partial x}{x} A_{1} \cos \left(x-v \cos A_{x} \cos (x+v t)-A_{2} \sin (x+v)\right. \\
& \frac{\partial^{2} y}{\partial x^{2}}=A_{1} \sin (x-1)
\end{aligned}
$$

$$
\begin{aligned}
& \frac{\partial x}{\partial x^{2}}=A_{1} \sin \left(A \cdot \cot (x-v t)+V A_{2} \cos (x+y t)\right. \\
& \frac{\partial y}{\partial t}=V A+V^{2} \sin (x+v t
\end{aligned}
$$

$$
\frac{\frac{\partial}{\partial t}}{\frac{d^{2}}{\partial t^{2}}=-V^{2} A_{1} \sin (x-v t)-V^{2} A_{2} \sin (x+v t)}
$$

$$
\left[A \cdot \sin (x-v t)-A_{2} \sin \left(x+v t^{2}\right)=\left[\frac{1}{v^{2}}\right]\left[-V A^{2} \operatorname{Ain}(x-v t)+A_{2} \cdot t_{4}\right.\right.
$$



$$
\text { (1) } \begin{aligned}
f^{\prime}=\frac{v^{\prime}}{\lambda^{\prime}}=f\left(\frac{v}{v-v_{s}}\right) & =\frac{v 1070}{\frac{f^{\prime}-f}{f}=\frac{f^{\prime}}{f}-1=\frac{v}{v-v_{s}}-1}
\end{aligned}=\frac{1070-44}{f}=1070-1026=\frac{44}{1026} 1026
$$



BOTH THERE IS NO MASS DIMENSION IN THE EQUATION. THUS THUS WAVE, FROM THE EQUATION IS INDEPENDENT OF MASS

b

$$
V=\sqrt{\frac{T}{\mu}}, V_{T=T, N S 10 N}^{\frac{2 T}{\mu}} \Rightarrow \frac{V_{1}}{V}=\frac{\sqrt{2 T / / \mu}}{\sqrt{T} / \mu}=\sqrt{2}
$$

$$
f=\frac{n v}{2 t}=\frac{(2) k}{2 T(1)}<20
$$

$3 / 6$


AUDIBLE SOUNO=20 TO $20,000 \mathrm{~Hz}$ ULTRASONIC $>20,000$

$$
\begin{aligned}
& \sin \theta_{c}=\frac{n_{2}}{n}=\frac{1}{n} \\
& \frac{R}{\sqrt{H^{2}+R^{2}}}=\frac{1}{R^{2}} \\
& R^{2}=\frac{H^{2}+R^{2}}{n} \text { Ccreless } \\
& R^{2}(n-1)=\frac{H^{2}}{n} \\
& R=\frac{H}{\sqrt{n(n-1)}} \Rightarrow D=\frac{2 H}{\sqrt{n(n-1)}}
\end{aligned}
$$



$$
\int_{2}^{n=1.5} \quad \begin{aligned}
&\left.\frac{1}{b}+\frac{1}{i}=\frac{2}{r}\right)\left(\frac{1}{f}\right. \\
&=(n-1)\left(\frac{1}{r_{1}}-\frac{1}{r_{n}}\right) \\
&\left.=(.5)\left(\frac{1}{36}\right)-0\right) \\
& f=729 \mathrm{~cm}
\end{aligned}
$$


$U$

$$
\begin{aligned}
& \frac{1}{i}=\frac{1}{0}-\frac{1}{f} \\
& \frac{t}{t}=\frac{f-0}{0 f} \\
& \hat{i}=\frac{o f}{(f-0)} \\
& i=0 f(f-0)^{-1} \\
& \begin{array}{l}
\frac{d i}{d o}=\frac{f}{f-0}+\frac{o f}{(f-0)^{2}} \\
\frac{d 0}{d t}=447 \frac{d i}{d t}=44\left(\frac{f}{f-0}+\frac{o f}{(f-0)^{2}}\right)
\end{array} \\
& =44\left(\frac{6}{6-18}+\frac{6(18)}{144}\right) \\
& =44\left(-\frac{1}{2}+\frac{3}{4}\right)=\frac{1}{4}(44)=11 \frac{\mathrm{~cm}}{5 E C}
\end{aligned}
$$

TOWARD THE MIRROR
FROM THE RIGHT $510 E$




1. A staple ship diffraction partexa is observed mo screen placed 2.0 $m$ from the site The discover between the first minimum on one side ot the control maximum and the trust minium on the other side is 2.0 cm . The wavelength of the 1 tIght is Shun A. RINd the with of the str in mo

$$
D=2 \mathrm{~m} \quad Y=1 \times 10^{-z} \quad \lambda=5 \times 10^{-7} \mathrm{~m} \quad m=0
$$

At mines $d \sin \theta^{\circ}=\left(m+\frac{1}{2}\right) \lambda$ m $X$.

$$
\theta \approx \tan \theta 2 \sin \theta=\frac{Y}{D}
$$

$$
\frac{d Y}{d D}=\frac{1}{2} \lambda \Rightarrow \frac{1}{D} d=\frac{\lambda D}{2 Y}
$$


2. The nearest star 1 s approximately 4.3 11ghtyease ( 4.07 x $10^{18}$ m) away from the earth. Suppose this star had planet with an orbit radius about the same ss that of the ear en - 93 milton miles $\left(1.49 \times 10^{1} \mathrm{~m}\right)$. What aperture diameterets necessary for a telescope to be able co just resolve the images of the star and planet? H ow many times larger or smaller is thus chan the $200^{*}$ Palomar telescope? Assume the mages are formed wt th light of wavelength 5000 A.

$$
\sin \theta=1.22 \frac{\lambda}{d}
$$



$$
d=\frac{1.22 \lambda}{\sin \theta}
$$

$$
\tan \theta 2 \sin \theta \simeq \frac{1 \times 49 \times 10^{10}}{4.07 \times 10^{18}}=3.66 \times 10^{-8}
$$



$$
\therefore d=\frac{(1.22)\left(5 \times 10^{-7}\right)}{3.66 \times 10^{-8}}=1.67 \times 10^{\prime}=16.7 \mathrm{~m}
$$



$$
\begin{gathered}
1 " 22.46 \mathrm{~cm} \\
200^{\prime \prime}\left(\frac{2.462 \mathrm{~m}}{1 "}\right)\left(\frac{10^{-2} m}{c m}\right)=4.92 m \\
x=\frac{162}{4.92}=3.4 \quad 3.29
\end{gathered}
$$

THE TELESCOPE WOOL HAVE
TO BE A LETTLE LESS
THAN 3: TIMES AS LARGE




（b）Wound the above film prevent the reflegeton pe similar 110ht Enctilent obliquely？Explain．yES THE NORMA／INCIDENT LIGHT数
 IS THE MOST QRITICAC CONDITION WHICH CAN OE SET FOR SUCH A SHSTENF THUS，OBlIGEE WCHT WOULD BE aLSO BE PREVENTED FROM AEIGCTAC．

莫。



Sharp edges of two ra cor blades taped together． make straight＂scratches en an opaque coathen on a glass plate（see sketres Red light Ancient normally on the plate passes through the two straight slits and forms several red slit imares－ivith centers one centimeter apart on a screen placed 3 maters From the plate．Estimate the thickness of ．each mayor blade from this information．

$$
\begin{aligned}
& \text { LE* X BE WITH OF }
\end{aligned}
$$

$$
\begin{aligned}
& \theta=\sin \theta \simeq \frac{Y}{D} \\
& y=\left(x+\frac{1}{x}\right) \mathrm{cm} ? \\
& d=\frac{\lambda}{2 \sin \theta}=\frac{\lambda 0}{2 Y}=\frac{3 \lambda_{R}}{2 X+1} \\
& \text { ASsuming } \lambda_{R} \simeq 800 \times 10-7 \mathrm{~m} ; ~ X=1 \mathrm{~cm} \\
& d=\frac{3\left(8 \times 10^{-2}\right)}{2 \times 10^{-2}}=12 \times 10^{-5}=.12 \times 10^{-3} \text { units? }
\end{aligned}
$$

Free question－Comment on what has or has not been of value to you ha your phystes courses，垂 you care to take the time．please be critheat．constructive and brief．（ write on back．）
mother：All good thurs must come to an mhd．True．False．mate bet






Relativistic historical
EXPERIMENTS

1) FITEAU A) ARAGO'S EXPERIMENT


FOR A SINGLE LEASE

$$
\frac{1}{f}=(n-1)\left(\frac{1}{R_{1}}-\frac{1}{R_{2}}\right)
$$

(n: el on class)

$$
V_{G}=V_{E} \text { IN GLASS IN LAB }
$$

AT $A ; \quad \frac{1}{f_{a}}=\left(\frac{c}{V_{0}+V} \cdot\right)\left(\frac{1}{R_{1}}-\frac{1}{R_{2}}\right)$
AT $B ; \quad \frac{1}{f_{b}}=\left(\frac{c}{V_{c}-V}-1\right)\left(\frac{1}{R_{1}}-\frac{1}{R_{2}}\right)$

$$
\begin{aligned}
& \frac{f_{a}}{f_{b}}=\frac{\frac{c}{V_{c}-v}-1}{v_{g}+V}+1 \\
& c=3 \times 10^{\circ} \mathrm{m} / \mathrm{sec} \\
& V_{g}=2 \times 10^{\circ} \mathrm{m} / \mathrm{sec} \\
& V=3 \times 10^{4 \mathrm{~m}} / \mathrm{sec}
\end{aligned}
$$

HOWEVER: ALL FOCAL LENGTHS CAME OUT THE SAME
B) ERESNEL-SAYS ARACO'S ERROR



IIDEAL GAS

1




MAXWELLS DISTRIBUTION

(8)






chapr. 3

VELOUTY GE PARTIELE IS CONSTANT
BECAUSE OF AR $\left(F_{B}=F_{8}\right)$
GLOSE SWITCH FORCE EE
UPWARD
DONE OY MULIRAN
E $=$ I. 10 IM COULOMB
IIMSS OF ELECTRON
m, e

\[

\]


EXPERIMENTHL OBSERVATTONS
$T \lambda_{\text {Max }}=C E N S T A N T$

- T 4



## II \| \| \|

*alozds aN17
THOMPSON ATOM MODEO



RUTHERFORD'S ALPHASCATTERING EXP.
SUPPORTED ASSERTATION WITH
MAGNETIC SCATTERING

$d$-IMPACT PARAMETER

$$
\cot \frac{\phi}{2}=\frac{m v^{2} d}{22 \theta^{2}} \cdot 4 \pi \in
$$



AU FOLL

RATO \# $\alpha$ PARTICLES WITH IMPACT
PARAMETERS $\leq d \quad(=n d)$
TO TOTAL $\#$ 's $(=n)$

$$
\begin{aligned}
\Rightarrow \frac{n d}{n} & =\frac{N \pi d^{2} \tau}{\text { UNIT AREA }} \\
& =N \pi d^{2} L
\end{aligned}
$$

RUTHERFORDS MODEL ACCERTED
$R$
$5-18-70$

ASSUMPTIONS
HESE
ORBITS COULDN'T RADIATE

$$
\begin{aligned}
& \text { HICH } \\
& \text { WHEN }
\end{aligned}
$$

$$
561840
$$

$$
\begin{aligned}
& \text { 6v17n Ni f } \\
& 109 \circ \mathrm{El}={ }^{+4} \mathrm{f}
\end{aligned}
$$

$$
\text { VRDDO ATNO AVW SNOY } 12373
$$

$$
\begin{aligned}
& E_{2}-E_{1}=10.2 E_{n} \\
& \text { CONG } n=1 \text { TON } 2
\end{aligned}
$$

FIRST SEEN BY LYMEN

5-21-70



$5 \cdot 25-70$

$$
\begin{aligned}
& \text { GOR NIERLE } A=1.65 \text { A } \\
& \text { TRIEO FOR OIEEERENT P } \\
& \frac{1}{5} \text { STILL WORKS, MAY } \\
& \text { USE ANY ELEMENT } \$ \text { ANY } \\
& \text { PARTICLE. } \\
& \text { MATTER COULD ACT LIRE } \\
& \text { WAVES. }
\end{aligned}
$$



## heISENBERG UNG. PRI.




$11$




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PHYSICS V - TEST 1
    April 24, 1970
```

Name ROBERT I, MARKS
I. In the following complete the statement, choose the best response, or perform the indicated operation. (2 points each)

1. An inertial system is one in which $15 \times N \angle L C$

2. According to maxweil.s equations. an electromagnetic wave in a vacuum propagates at speed where c is the ratio of ELESAIC 7 MAGNETIC FIELDS.
3. The value of c in mks units is $3 \times 10^{\mathrm{s}} \mathrm{m}$
4. What, where, or who is Algol? $\qquad$ A biNARY STAR。
5. As far as classical E $\mathcal{E}$ is concerned, what was the purpose of the aether ? TO PROPOGATE LIGMT .
6. What was Fresnel's suggested explanation of Aragon's null experimental result? THERE WGS E ETHER DRAG THE EARTH DRAGGED THE ETHER.
7. Fresnel's explanation led to the following expression for the velocity $V$ of light in a moving medium:

$$
V=c / n+\left(1-1 / n^{2}\right) v
$$

where v is VELOCITY OF MOVMG MEDMA
8. The classic experiment to detect the earth's motion through the aether was performed byMMCHE SON
$\qquad$ -
9. The result of that experiment was NEGATLE (ON ROTADON $)$
10. Who performed the aether experiment that Einstein considered crucial? $\qquad$ ELEEAU
II. 1. Write down the equations of the Lorentz transformation which transform the coordinates of system $S$ into those of system $S^{\prime}$ which is moving at a constant speed $v$ down the $+x$ axis of $S$. Define any symbols introduced.


( 15 points)

$$
\begin{aligned}
& x=\frac{1}{\sqrt{1-V^{2} / c^{2}}} \\
& t^{\prime}=y\left(t-\frac{1 x}{c^{2}}\right) \\
& x^{\prime}=d \quad\left(x-y^{t}\right) \\
& y=y \\
& z^{\prime}=z
\end{aligned}
$$

2. Derive an expression for $u^{\prime}$, the $x$-component of the velocity of an object measured in the s'system in terms of $u$ and the coordinates in the system $S$.

3. State the two basic assumptions of Einstein's theory of special relativity. ( 10 points)
1) THE SPEED OF LIGHT, AS MEASURED BY AN OBSERVOR, IS INDEPENDENT OF THE VELOCITY DE THE SOURCE 2) NO ABSOLUTE UNHEORM MOTMON CAN BE DETECTED
4. Suppose you are moving with a speed of 0.75 c past a man who picks up an object and then sets it down. If you note that he held it for 9 seconds, how long does he say he held it? (ll points)

$$
\begin{aligned}
& V=.75 C \\
& t=\frac{t^{\prime}}{\sqrt{1-V^{2} / c^{2}}}=\frac{t^{\prime}}{\sqrt{1-562}}=\frac{t^{t}}{\sqrt{44}}=\frac{t}{.665} \\
& t=\frac{9}{463}=13 \times \text { secenas }
\end{aligned}
$$

5. Does the non-relativistic formula $1 / 2 \mathrm{mv}^{2}$ over-estimate or under-estimate the kinetic energy of an object moving at speeds near c (Don't guess !)(10 points) UNDERESTMMAED
THIS FORMULA IS DERIVED

FROM THE CLASSICAL ENAPESSMON F= Ma, WHMC STATES NN EFFECT A MASS MAY ACHEIVE ANY VEACEMTY N TRUTH A MASS TAKES A MUCH LARGEAMOUNT OFENEACY TO PEACH VELOCITIES NEAL $A H$ MANA THE THEY HL HAVE A MUCH LARCENER K THAN MY:
6. A measuring rod is at rest in system $S^{\prime}$. If $S^{\prime}$ moves down the $+x$ axis of $S$ at $0.8 c$ and an observer in $S$ measures the rod to be 6 meters long, how long does an observer in $S^{\prime}$ measure the rod to be ? (15 points)

$\qquad$

PHYSICS V－TEST II


May 14， 1970
Name


I．In the following fill in the blank with the word or phrase that best completes the statement．（2 points each）

1．The Boltzmann constant is defined as


2．According to the rule of Dulong and Petit，the molar specific heat of a solid is $\qquad$

3．An important assumption of our kinetic theory derivation of the gas law was that no forces act on the molecules except DUP QNG COL LISHOMS
$\qquad$。

4．A classical analysis indicates the maximum kinetic energy of photoelectrons will be proportional to $\qquad$

$\qquad$
$\qquad$。

5．The experiment of Zartman and Ko was designed to $\qquad$
$\qquad$。

6．The average kinetic energy per degree of freedom for an ideal gas is $\qquad$。

7．According to the Rayleigh－Jeans law，the amount of energy radiated by a black body at all wavelengths is c res $\lambda^{-4}$

8．The Thompson experiment was designed to measure

9．The magnitude of the electron＇s charge was measured in a classic experiment devised by $\qquad$
MLELELSON

A plot of stopping potential vs．light frequency for the photoelectric effect is a straight line of slope＿

II. (a) Write down the Maxwell distribution of kinetic energy for a gas with $N$ molecules / unit volume. (5 points)

(b) Show by writing down an appropriate integral equation (complete with limits) how you might calculate the average kinetic energy of molecules that have a Maxwell distribution. ( 5 points)

(c) Sketch on the axes below the Maxwell kinetic energy distribution $\boldsymbol{\theta}^{\text {b }}$ (K) vs.K. ( 5 points)

III. (a) If you studied the photoelectric effect in calcium, you would find for light of the indicated frequencies the following stopping potentials:
frequency
$1.18 \times 10^{15} \mathrm{sec}^{-1}$
stopping potential
$0.96 \times 10^{15}$
1.95 volts
0.98
$0.82 \times 10^{15}$
0.50
$0.74 \times 10^{15}$
0.14

5

Plot these data carefully on the attached graph paper. Determine from your plot the
(1) minimum frequency
(2) maximum wavelength
(3) the work function in volts
(4) the slope in joule-sec.
(b) State two results of the photoelectric effect experiment that cannot be explained by classical physics. ( 5 points)
1SUPPORN OF THE EXISTAMEE OE PHETONS
2) LCHT INTENSITY LOES

3

(c) Using an energy diagram such as was used in class, outline briefly Einstein's derivation of the photoelectric equation. (lo points)
IV. (a) On the axes below, label and sketch carefully the theoretical curves predicted by the Rayleigh -Jeans and Planck equations for radiation emitted by a black body at temperature $T$ as a function of wavelength. ( 10 points)

(b) On the plot in (a) sketch also the experimental curve measured for a black body at temperture $\mathrm{T}^{l}$ which is less than T. ( 10 points)
10
(c) Planck showed that the difficulty with classical theory was concerned with the calculation of the average energy per degree of freedom of the oscillators. Write down (don't derive) the classical and quantum expressions for the average energy. (10 points)


CLASSICAL SAID



4 f TMME ${ }^{-1}\left(\mathrm{SEC}^{-1} \times 10^{-15}\right)$
SLOPE = Ve

$$
\begin{aligned}
& e V_{0}=h f-\phi \quad\left(\phi=h f_{0}\right) \\
& V_{0}=\frac{h}{e} f-\phi \rho \\
& f_{0}=-7 \times 10^{+15} 4 z=\text { mivimum } f \\
& V_{0}=\frac{h}{e} f-\frac{h f_{0}}{e}
\end{aligned}
$$

$$
\begin{aligned}
& \text { d) SLORE }=\frac{4.45 v-5 E C}{2.7 \times 10^{115}} \\
& =1.65 \times 10^{-15} \mathrm{~V}-\mathrm{sec} \\
& \frac{6.63 \times 10^{-34} \text { sec }}{4.14 \times 10^{-15} \mathrm{~V}-\mathrm{sic}}= \\
& 1.6 \times 10^{-19} \frac{0}{V} \\
& 1.65 \times 10^{-15} \mathrm{~V}-\sec \frac{\mathrm{d}}{1.6 \times 10^{-19} \mathrm{~V}} \\
& =1.03 \times 10^{4} \mathrm{~J}-5 E C
\end{aligned}
$$

$\qquad$
June 5. 1970

$$
\frac{\text { PHYSICS } V-T E S T \text { III }}{(100 \text { points })}
$$

Name $\qquad$ ROB MARES
(You may find some of the following information useful:
$e=1.6 \times 10^{-19}$ coulombs, $h=6.62 \times 10^{-34}$, joule-seconds, $\mathrm{ca} 3 \times 10^{8} \mathrm{~m} / \mathrm{sec}$.
electron mass $=9.11 \times 10^{-31} \mathrm{kgm}$, one Angstrom $=10^{-10} \mathrm{~m}$ )
I. (a) An electron is accelerated from rest through a potential difference of 250 volts. What is the deBroglie wavelength associated with this electron? (10)

(b) Suppose the potential difference in part (a) were changed to 25,000 volts. The correct answer would then be (circle the right one)

1. 100
2. $1 / 100$

1
3. 10

5. a little less than $1 / 10$
6. a little more than $1 / 10$
times the answer to part (a).
(c) Explain the reason for your choice in part (b).
$V \times \frac{1^{2}}{1 \infty} 1$ Vo x $C^{2}$
$\begin{aligned} 1-1 & \frac{1}{10}=U \text { WOULD CREASE }\end{aligned}$
(d) The electrons in part (a) are to be used in a Davisson and Germer experiment. At what angle would you expect the first maximum of the diffraction pattern to occur if the crystal used has a spacing between planes of 1.985 Angstroms? cAlvad
II. (a) Diagram and label carefully a collision between a high energy photon and a free electron (Compton effect) and write down the equations for conservation of energyand momentum in the collision.

(b) Which of the quantities in the equation in part (a) would be measured in a Compton effect experiment? (5)
 NEQ $\angle 4+C+E$
III. (a) In the space at the right sketch carefully the energy levels for hydrogen (according to Bohr es theory) for $n=1,2$. 3.4 , and 00 . Show $n$ and $E$ values $i n$ lv). (10)
$E_{n}=\frac{E^{2}}{n^{2}}$
$E_{(H)}=13.6 \mathrm{eV}$ 9

(b) On the diagram above shown by arrows the following spectral lines: Lyman $\alpha$, Lyman $\beta$. Balmier $\alpha$.
Balmer $\beta$, and Balmier limit.
(c) From information in the diagram compute the wavelength of the Calmer $Q$ line.

$$
\begin{align*}
h \omega P & =E_{1}-E_{1} \\
& =3.401 .52 \\
h H & =\frac{153}{5}
\end{align*}
$$

IV. (a) Write down the five postulates of Schrodinger's wave mechanics.


ANALOGOUS TO DERINTIOM OF WAVE PRORECATION IN AN E FIELD
(b) The Schrodinger equation is the wave mechanical equivalent of what classical statement?

I) SPECIAL RELATIVITY

1) THE EINSTEIN POSTULATES
2) SPECIFICALEY:
a) ABSOLUTE, UNILOM NOTION GANOF BE DETECTED
b) THE SPEED OF LIGNT IS NOEPEN. OF THE SOURCE
3) SUPPORTED FROM MAXWELLL'S EQuATIONS
4) SUPPORTEO: NO PROOF OF THE EXISTANCE OF ETHER
B)MIENELON-MORELY EXP.

NFL BEAUS EXP. - NOT EXACT ENOUGH
21 MICHELSON'S INFEROMETER

a) OBSERUE INTGREERENEE PATTERN, NOTATE GO; OBSERVE ANOTHER, THUS COMPUTING GUT DRAG IN ETHER
b) NO SHFT OBSERVED UPON ROTATION
3) ABERRATION OF LICHT
a) TELESCOPE MUST DE THTED TO GET MAE OF STAR $1 T$ ANGLE FROM STARS TRUE POSH
bJEARTH OOESNT DRAG ETHER
c) RINEMATICAL CONSEQUENCES OF EINSTEN'S POSTULATES

1) TIME OILATION

$$
\Delta t=\Delta t \cdot \sqrt{1-V^{2} / c^{2}}
$$

2) LENGTH GONTAACTION

$$
L_{0}=L^{\prime} / \sqrt{1-v^{2} / c^{2}}
$$

3) LIFETIMES CHALE LIFE)

$$
T=T \cdot \sqrt{1-v^{3 / c^{2}}}
$$

D) SIMULTANEITY $\frac{1}{\text { I CLOCK GYNCMRONIEATION }}$

1) TWO EVENTS IN A REFENENCE FRAME ARE SIMULTANEOUS IF THE LIGHT SIGNALS FROM THE EVENT REACH AN DOSERVOR HALFWAY BETWEEN THE EVENTS AT THE SANE TIME.
2) EVENTS THAT ARE SIMULTANEOUS IN ONE FLAME ARE NOT IN ANOTHER
3) TWO CLOCKS SEPARATED BY $L_{0}$ 高 SHNEHRONIEED IN REST FAME ANE UNSYNCHRONIEED IN A REFERENCE FRAME MOURNE WITH SPEED V. THE CHASING CLOCK IS AHEAD BY VLo/C
4) PERPENDIGULAR WAVES TRE THE SAME MEASURED IN ANY SYSTEM
E) THE LORENTZ TRANSFONAATION
5) $x=\gamma\left(x^{\prime}+V t^{\prime}\right)$ i $x^{\prime}=\gamma(x-v t)$

$$
\gamma=\frac{1}{V_{0} 1-v^{2} / c^{2}}
$$

2) $t=\gamma\left(t^{2}+V i / c^{2}\right) ; t^{\prime}=\gamma\left(t^{-V} / c^{2}\right)$
3) cceux s YNCMACN ZatyMgA

$$
\begin{aligned}
& d+b x+2(x+x)-x+x-x
\end{aligned}
$$

F2 7ME HELectry Thatherandedtrad

$$
\begin{aligned}
& (-2)+4)^{4}()^{2}+4^{2}+x^{2}+x^{2}+4^{4}
\end{aligned}
$$

e) the peppleq eftect
$\left.1+1(c+y)_{0}\right)^{2 E}=\left(\frac{c x}{y}\right)$
2) PAEGUENCY
$4+10$ ta $d$
by HoYROA MARBNC ANGLE S YETMR sewnes t actershoh

$$
f=\sqrt{1}+\left[y+e^{2}+e^{2}+1\right.
$$

HDRELATHYKHE ME SGATHRA

1) $P=$ mber $r a+a^{2}+c^{2}$
2) REBATHPGPLC $4 A \leq 8 \quad$ M $<4$

$$
m f \sqrt{p} \cdot b^{2} p^{2}=1
$$

3) RELATMUSTVC ENENGY

4) $E=7+h c^{2}=t h+e^{2}+d^{2} E_{0}$
5) $E^{2}=p^{2} c^{2}+E_{0}^{2} \quad\left(b+\left(p-e^{2}+c^{y}\right)^{4}\right)$
6) Far tho cobwidnue pArticess OF EQUALMMES

b) hesp +4.s. Mecemeres $\Delta m-2+m \cdot(x-1)$
CDERPGRAL RHAETRC ENEREY IN
 BAEROB

$$
T * T^{2}+d r
$$

a) ass of ENEREY shate AN
1)MASS \& BMMME ENERGY

1) $14,1.68 \times 10^{-87}+8^{4} 1$

$318 \times e^{-2}+5^{4}+4 x+x^{4}+10=1$
$41+2+2 e^{2}+98+S^{2}+4$


I) THE KMETIG THEORY OF MATTER
A) OISTRIGUTION FUNCTION
2) $\& \mathrm{~F}_{1}=1$ NORMMUIEATION

FUNETIOM
2) AvERACE:

$$
\bar{S}=\frac{1}{N} \sum_{i} S_{i} n_{i}=\sum_{i} S_{i}
$$

3) AVERACE SOUARE:

$$
5^{2}=\sum s_{i}
$$

4) STANDARD DEVIATION
a) $0=\left(5\left(5 x-\frac{5}{5}\right)^{2} f^{2}\right)^{\frac{5}{2}}$

B) PRESSURE OF A CAS

DASSUMPTIONS
D)ARGE (N) MOLEE: MAKING ELASTHE COLLISIONS
b) MOLEE SEPARATED GY LARGE DISTANEES COMP WITH DIA:EXERT NO EORCE ON EACH OTHER
C)NO DREFERED POSITION OR $V$ For MoLEC
2) $P_{x}=2 n m \int_{0}^{\infty} v_{x}^{2} f\left(v_{x}\right) d v_{x}=n m v_{x}^{2}$
d) $n=N /$ Vouvme
b) $M=1155$ OE M LEEULE
3) $V_{x}^{2}=V^{2}=V^{2}=\frac{1}{3} V^{2}$

$$
\therefore P=\frac{1}{3} n^{2}=
$$

$41 P V=\frac{2}{3} N E=3=N R T$
a) $E_{K}=$ AVGRAGE KINETR ENERGY
b) $R=1.99$ onLOK-MOLE=3.31U KNMOLE
5) $E_{k}=\frac{R_{A} T}{N_{A} T}=\frac{3}{2} k T$

K=golTZMIN's constint = R/Na
4) VRMS $=(3 \mathrm{RT} / \mathrm{H})^{\frac{1}{2}}$
$y=$ MASS of A MOCE
7) $C_{V}=\frac{T}{2} R=2.98 \mathrm{CAL/10L}==\frac{\mathrm{L}}{}=$
c) THE MAXWELL-DOLTEMANN OISTRIBUTION

1) $f\left(v_{x}\right)=\sqrt{\frac{m}{2 \pi k}} Q^{-m v_{x} / 2 k T}$
2) $F\left(v_{x}, v_{r} \cdot v_{B}\right)=\left(\frac{m}{2 \pi / k T}\right)^{\frac{3}{2}} e^{-m\left(v_{s}^{2}+v_{4}^{2}+v_{5}^{2}\right) / 2 k T}$
O) OTHER DERIUATIONS (WERE DONE)
E) EQUAPARTITION THEOROM AND

HEAT CAPACITIES OF EASES F SOUDS

1) THE AVERAGE ENERGY OF LKT IS ASSOCMTED WITH EACH CO-OROINATE OR MOMENTUM COMPONENT APAEARINO IN THE ENERGY AS A SQUARED TERM

$$
E=H\left(\frac{1}{2} k T\right)
$$

2) OULONE PETIT: MOLAR MT CAPACTIES OF MOST SOLIDS IS AOOUT
a) $c 4 / 0$ M mole $3 R$
b) TREAK MELEK ATOMS AS ATTATCHED BY SPRINGS (K=SPRNG CONSTAMT)

$$
E=\frac{1 m}{2 m}\left(P_{x}^{2}+P_{y}+P_{z}^{2}\right)+\frac{1}{2}\left(k_{1} x^{2}+k_{2} x^{4}+k_{3} x^{2}\right)
$$

c)AT HGH TEMP, ALL SOLHDS OEEY LAW

4


she for 0
$x-a \quad x^{2}+x^{2}$




AANDEM WhLR PAREREM
III) THE QUANTIEATION OF ELECTRICITY, LIGNT ENERA
A) MEASUREMENT OF ELECTRIC CHARCE

1) FARADAY $=N_{A} C$
2) TOWNSENO-5STMMTED E $1 \times 10^{-1}$ Cocl
3) THOMPSON $\$$ WILSON
4)MHCIRAN t HIS OIL DROP
a)chances occun IN NuLTIPLES OFG
b) LITTLE OIL OROPS IN A CAAACHOK
4) MEASURGMENT OF G/m

D EEEMAN-ROTATITING CHARGE
YELDING ELECTROMMGNETIC WAVES
2) J.U. THOMASON
I) WHEN $O$ IS PLACED 1 TO PATM.

AARTICLES MOUE IN CINGLE:
(1) $R=M V / q B$
(a) $\sum^{2}=D^{2} p^{3}$ Q/2W.

$$
\left(W=N=m V^{2}\right)
$$

b) U.U.THOMPSON EXPERMMENT
$010+$ HEIELDS
3) DLACKBODY PADIATION
4) CLASSICAL (RAYLERGHOJEAANS LAW)

$$
f(\lambda) d A=8 m \lambda^{-d} d \lambda
$$

b) Max PLANEK

OASSuMED $E=\sum_{n} E^{-E n / k T}$

$$
\left(E_{n}=n h f\right) \text { ar nelntrecen }
$$

(2): Ev nclaf(e hc/AkT-1)
$\operatorname{coc}(\lambda)=8 \pi h e \lambda-5 /\left(e^{\circ}\right.$ he/nNTm $)$


$$
h=6.426 \times 10^{-34} 46 E=4.134 \times 10^{-15} V_{\mathrm{g}}
$$

c) THE PHOTOELECTRIC EFFECT
a) LENARD

$$
\begin{aligned}
& \text { (1) eVa }=\frac{1}{2} m V^{2}=h f-\phi \\
& \text { ( } \phi=\text { ENERGY NESCESANY TO REMOVE } \\
& e \text { FROM SURFACE) } \\
& \text { ( } \phi=h f_{0}=h c / \lambda t \\
& \text { a } \lambda_{t}=\text { THRESHOLD ER } \lambda \\
& \text { bic }=1.24 \times 10^{4} \mathrm{eV}-A
\end{aligned}
$$

D) $X$ RAYS THE COMPTON EFFECT I SCATTERING OF \& RAYS BY

FREE ELECTRONS


1) $\lambda_{2} \cdot \lambda_{1}=\frac{h c}{E_{B}}(1-\cos \theta)=\frac{h}{h c}(1-\cos \theta)$
2) 1 MC = COMPTON $A$

$$
\text { . .0243 } 1 \text { for é }
$$

IV）THE NUCLEAR ATOM
A）EmpIrical spectra formulas
CDATA collected on Emission of light BY ATOMS IN A GAS WHEN EXCITED
ELECTRICALLY OR IN A FLAME）
1）Johann balmer－lines could be REPRESENTED ${ }^{3}$ BY（FOR H）

$$
\lambda=b \frac{m^{2}}{m^{2}-4}
$$

$$
m=3,4,5
$$

$$
b)=3.45 \times 10^{-8} \mathrm{~cm}
$$

2）RYOBERG：

$$
B E_{A}=\left(\frac{1}{A}\right)_{\infty}-\frac{R}{(m+U)^{2}}=\text { WAVE }=f
$$

D）R＝RYOBERG＇S CONSTANT $=1.10 \times 10^{5} \mathrm{~cm}^{-1}$
b）$f=\frac{c}{\lambda}=c f$
3）BITE：

$$
f: R\left[\frac{1}{(m+A)^{2}}-\frac{1}{(n+B)^{3}}\right]
$$

a）$m$ ？$\frac{1}{\text { a }}$ ARE NTEGERS
b）$A$ B CONSTANTS
4）ग．l．THOMPSON＂PLUM PuDDinG＂
MODEL ON THE ATOM．UNABLE TOSUPMOH
B）RUTHER FORD SCATTERMNC（STUNENTS GEIGER ⿳亠口冋刂
1）SHOT \＆PARTICLES THRU SEREEN， SOME DEFLECTED OVER $90^{\circ}$
2）DISPROVED THOMPSON＇S ATOM MODEL
C）BOHR MODEL OP THE ATOM
1）RADIATION EMITTED BY ELEETRONS CMANGING ORBITS，RATHER THAN ROTATING

$$
\text { hf } f=W_{\text {ORBIt I }}-W_{\text {ORBIT }}
$$

2）DERIUED：

$$
f=\frac{1}{2} \frac{k z e^{2}}{h}\left(\frac{1}{r^{2}}-\frac{1}{r_{1}}\right)
$$

a）$r_{n}=n^{2} r_{0}$
b）$Z e$＝CHARGE ON NUCLEUS
3）CORRESPONDANCE PRINCINLE，FOR LARGE QuANTUM \＃i（n）cuASICHE S QuaNTUM
4) DENVED:
a) $W_{n}=-\left(\frac{2 \pi x^{2} e^{4}}{h^{2}}\right) \frac{z^{2}}{n^{2}}=-z^{2} \frac{E_{1}}{n^{2}}$ for H $E_{1}=13.6$ el (GMMEAGENERGy)
b) EMERGY LEvEL MGMAM

prices
BRACKET SERIES-INFRA-RED
cALMER SERIES - visible
LYMAN SERIES -ULTRAVIOLET
5) IN AN PATOM
a) $E_{R}=\frac{p}{2 \mu} ; \mu=\frac{m M}{m+M}$ ( $N=$ MASS of NUC.)
b) $\mu$ is CALLED REDVCED MASS
c) REASON FOR VABIATION OF RYOBERG CONSTANT FROM ElEMENT TO ELEMENT
6) BOHR SAID (POSTULATED)
a) (=ANG MOM. OF e) $=n \%$
b) $f$ (DE RADIATION $=\left(\omega_{1}-\omega \cdot\right)_{h}$
D) FORAY SPECTRA

1) CRYSTAL SPECTROMETRY DEVELOPED

BY THE BRACES
2) MosELEY, PLOTTEO $f^{\frac{1}{2}}$ us $z$ (ATOMiC:

$$
f^{\left.\frac{1}{2}=A(z-6)^{2}\right) \text { USING } X-R A Y S}
$$

(A 1 O CONSTANTS FOR EACH LINE)
E) THE FRANCR-HERTE EXP.


1) ELECTRONS ACCELERATED FROM HEATED CATHORE THRU GRIDO ( $v_{0}$ )
2) 



CORRESPONDS TO EXCITATION
OF DHFERENT LEVELS
IN H
F) WLSONGEMMEREIECD GOANTIZATION RULE

$$
\phi p d q=n h
$$

1) $P=$ COMPONENT OF MOMENTVM, SECH A $X X$

D) ELECTRON WAVES
A) THE DE BROGLIE RELATICNS (ELECTRON UAVES)
$1) f=E / h \quad \because A=\frac{h}{P} \quad(\rho=M O M E N H L A)$
2) DEWELOPED BY S SGGROEDNNERR
B)MEASUREMENTS OF ELECTRCN WALELENGTAS
3) MADE BY DAULSOM FGERMER
2)" LELECTRON GUN

$$
\text { MEASURED AS. } 1.67 A^{\circ}
$$

c) CLASSICAL WAVE EQUATIOUS

1) FOR ASTRING:

$$
(\rho A) D^{2} Z^{2}=T \frac{\partial^{2} y}{\partial x^{2}}
$$

ANSLUER

$$
Y(x, t)=y_{0} \cos (k x-\omega t) \text { or } Y(x, t)=r_{0} e^{i(k x-\omega t)}
$$

2) OSEFULL PARAMETERS
a) $k \lambda=2 \pi$
b) $\omega=2 \pi f$
D) WAVE PACKETS
E) ELECTRON WAVE PACKETS
3) $\psi(x, t)$ ( $=$ WAVE PACKET)
$=\int g(k) \cos (k x-\omega t) d k$
(ECK) CONTAINS RANGE OF WAVE NUMBERS CENTERED ABOUT $\mathrm{K}_{0}$ )
4) $V$ (GVELOCHTY OF WAVE PACKET) FP OM
F) THE PROBABILISTIC INTER PRETATION OF THE WAVE FUNCTION (RELATION OF $\psi(x, t)$ TO ELECTRON'S LOCATION)
5) $\varepsilon^{2} \alpha$ PROBABILITY OF PHOTON BEING

IN UNIT A氺EG VOLUME
2) $\psi^{2} \alpha$ PROBABILITY OF ELECTRON BEING

IN UNIT VOLUME
$\left(\psi^{2} d \&\right.$ ELECTRON BEING MN INTERVAL $d X$ )
G) THE UNCERTAINTY PRINCIPLE (METSENBURG)
$\Delta x \Delta \rho \approx \hbar \approx \Delta E \Delta t$
H) particle wave duality

BOHR'S PRiNCIPLE OF COMPLEMENTARITYPARTICLE ASPECTS 言 WAVE ASPECTS COMPLEMENT EACH OTHER
H) CONSEQUENCES OF THE LNGERTAINTT PRINCIPLE
A) PARTILLE in SMALL SPACE MUST HAVE K.E. (not observable for mackoscomi)
B) BUNCHES ELSE
d) SUMMERY
*) electrons and other
PARTICLES DISPLAY WAVE

PROPERTRES
a) $E=h f=h \omega$
b) $p=\eta A=7 h$
2) IF $\lambda$ IS SMALL, $P$ Y LOHT WALE PROPEATISS

CAN BE NEGLECTED.
3) (WAVE FUNCTICN) ${ }^{2}$ © PROQABILTY
D) THE SGHROEDINGER EGOATION IN ONE DIMENSGDN A) PGAUSIGMMY ARGUMENY FOR SOHROEAHGER EGOATRON(E)

1) $\left.j \frac{d y}{d t}=-\frac{\hbar_{m}^{2}}{2^{2} b^{2}}+V<x\right) \notin$

SOLUNK YIELDS:

3) $P(x, t) d x(p R O D A G / L T Y)^{\circ}=4+q d x$

$$
\left(\text { His is COMPLEX CONNCATE OF }^{3}\right)
$$

4) NORMLLEATION CONDITON

$$
\int_{\infty}^{\infty} \psi \times \ell_{\infty}+x=1
$$

a) THE TME MMEPENDENT S.E

1) $\angle E T \quad \psi(x, t)=f(t) \psi(x)$

$$
\Rightarrow f(l)=A c^{-i \omega t}
$$

LEADS TO THE MNDENENDENT SE

$$
-\frac{1}{2 m} \psi(x)+V(x) \psi(x)=E \in \psi(x)
$$

2) DROBABILITY DISTRUGOTIOA

$$
y * \varphi=\not \subset *(x) \not C(x)
$$

3) MOST GENERAL. SOLUTION TO TGHS.E.

$$
\begin{aligned}
& \notin(x, C)=\sum_{n} C_{n} C^{\left.-x\left(E_{n} / T_{1}\right) \sum_{n} C_{x}\right)} \\
& \text { (CNF POSSIBLE LNEAGY OF PAMCICLE) }
\end{aligned}
$$

4) Y(X) MLIST MEET GOLCOMNMG CONDTIONS
5) SATSFY S K
2.BE CONTINCOUS
6) $(X)$ ( $B E$ COMTHUOLS
7) $4(x)-50$ as $x \rightarrow+\infty, 40,4<x$ CAL BL ND RAMLTELGO
C)GNERCY GUANTEATION FROM S.E.

d) $U_{x}=\frac{U_{4}^{6}+V}{1+U_{4} V / c^{2}}=\frac{1.8 c}{1.81}=.9940$
b) $U_{0}=\frac{6 \times 10^{4}}{1+\frac{36 \times 10^{4}}{9 \times 10^{4}}}=\frac{6 \times 10^{4}}{1+4 \times 10^{-}}=\frac{6 \times 10^{4}}{1.00000004}=6 \times 10^{4} \frac{m}{560}$
c) $23 \times 10^{5} 70$
8) 

$$
\begin{aligned}
& X^{2}=\frac{\left(C-V^{2} \lambda^{2}\right.}{C^{2}-V^{2}}=\frac{(C-V)^{2} A^{+}}{(C-V)(C+V)}=\frac{(C-V)^{2}}{C V} \\
& \frac{X^{2}}{X^{2}}=\frac{C-V}{C+V} \\
& C \frac{\lambda^{2}}{\lambda^{2}}+V \lambda^{2}=C-V \\
& C\left(1-\frac{x^{2}}{x^{2}}\right)=V\left(1+\frac{x^{2}}{x^{2}}\right) \\
& \frac{1^{2}}{1^{2}}=(1.22)^{2}-\left(\frac{1}{3}\right)^{2}=\frac{1}{9} \\
& C\left(\frac{d}{8}\right)=V\left(\frac{1 a}{4}\right) \\
& V=8 C
\end{aligned}
$$

B)

$$
\begin{aligned}
& H_{e_{3}}=3.016030 \\
& n^{5}=1.008665 \\
& H_{e_{4}}=4.004695 \\
& E_{n}=\frac{1022075}{4.024}
\end{aligned}
$$

$$
\begin{aligned}
\frac{2.6 \times 10^{-80}}{21.6} \times 10^{2} 70 & =495 \times 10 \\
& =4.55 \times 10^{20}
\end{aligned}
$$

1) 

$$
\begin{gathered}
k^{*} \rightarrow \pi+17 \\
M k=495 \quad M e v / c^{2} \\
M \pi=135 M=4 C^{=} \\
2(135)=270 \\
\frac{498-270}{2}=114 \text { MeV }
\end{gathered}
$$

$$
\begin{gathered}
4^{0} 0^{0} \\
a^{2} \\
\frac{0}{a} \\
a^{2}
\end{gathered}
$$

唯

 $\qquad$
$\qquad$

## Experimental Errors I - Errors in Measurements

The error in the calculated result of an experiment can be as important as the answer itself. It is impossible to come to any conclusion about what the experiment shows or does not show until some estimate is made of the probable magnitudes of the experimental errors and their effects on the result. Every engineer or scientist who analyzes experimental data must have some knowledge of the methods of error analysis.

The following exercise is designed to introduce you to some of the methods used in analyzing experimental errors. The exercise.will be of value to you only if you work through it carefully, following the directions below, and think about it while you are doing it. You will be tested on these methods later.

Each numbered section of the exercise is called a frame. Some frames are information frames and others ask questions which you are to answer before going on. The correct answers to the questions are given below the questions and should be covered by a "mask" (piece of paper) until you have supplied a written answer to the question.

## Dinections:

1. Read each statement and question carefully and write your answer down (on another sheet of paper, not the exercise sheets) before looking at the printed answer.
2. Work carefully and directly through the exercise in sequence without skipping or browsing. If necessary you may refer back to earlier frames at any time.
3. If your answer tums out to be incorrect or incomplete, try to get straightened out by re-reading the pertinent information before going on to the next frame.
4. The term "error" as it is applied in experimental work does not include mistakes in arithmetic or mistakes such as recording the wrong number or reading instruments incorrectly. These mistakes may, and should, be eliminated completely by careful work. There are still, however, two kinds of errors which may be present in a series of measurements:
(a) Systematic errors------any error that causes a measurement to yield a value which is consistently too large or too small. For example, if a voltmeter reads consistently high because of incorrect calibration, it is contributing a systematic error to the experiment.
(b) Random errors-m---an error which results from chance variations in the measuring device, the observation method, or the quantity being measured; there is an equal chance of these variations producing positive or negative errors. For example, there may be random error in reading a voltmeter because the observer does not always observe from the same angle. (There also may be a systematic error if the observer tends to read the meter at an angle to one side of the vertical.)
5. What is the principal difference between systematic and random errors?

Systematic errors produce results which are systematically too high or too low (one or the other). Random errors are due to chance variations and are just as likely to produce a value which is too high as to produce one which is too low.
3. In which of the following cases are the errors systematic and in which are the errors random in nature?
(a) Objects are weighed on a balance that is not first correctly "zeroed".
(b) The volume of a liquid is measured several times by different observers using an accurate eraduated cylinder and noting the level of the liquid on the scale.
(a) Systematic. Readings will be systematically high or low (one or the other).
(b) Random. It will be difficult to line up the eye with the scale and liquid surface the same way every time.
4. In which of the following cases is the main type of error systematic and in which is it random?
(a) A timer is running slow.
(b) Friction is ignored in an experiment in which it is not quite negligible.
(c) A galvanometer (used for measuring small quantities of charge and currents) is being affected slightly by vibrations.
(d) Fluctuations in the line voltage of an electronic measuring device.
(e) The use of weights which are light due to part of theirimass being worn away over the years.
(f) Bias of an observer which causes him to always read a scale from a point to one side of the vertical.
(a) Systematic
(b) $S$
(c) Random
(d) $R$
(e) S
(f) S
5. Often the largest source of error in an experiment is a systematic error. The experimenter should always look for sources of systematic error in his experiment, and try to eliminate them or correct for them as much as possible. For example, one can guard against observer bias in reading a scale by having several different observers make the same reading without knowledge of the others' results, or by constructing a viewing arrangement which makes bias impossible. He can make sure that all his instruments are properly zeroed before using them. He can calibrate his instruments and measuring devices by comparing readings taken with them with readings taken with another instrument which is known to give results which are more likely to be accurate. Systematic errors are sometimes difficult to prevent, but as you become more experienced with correct experimental procedures you will find these errors easier to detect and correct.
6. How might you go about eliminating systematic error from an experiment in which you suspect a timer may not be keeping accurate time, after the experiment is done?

There are probably a number of correct ways. One way would be to calibrate the timer against a timing device that is more likely to be accurate and make the appropriate corrections to all time measurements in the experiment.
7. Even when systematic error is not eliminated from an experiment it is often possible to estimate, at least roughly, how large it might be. When you use a resistance box with resistors which are supposed to be accurate within 1\%, for example, you can assume that if the box has not been mistreated, the resistance valves are within $1 \%$ of the dial settings. A voltmeter which has specifications which state a $5 \%$ accuracy has been calibrated at the factory to have a systematic error of less than $5 \%$ at full scale reading.
8. Suppose you were doing an experiment using a $1 \%$ resistance box and a $5 \%$ voltmeter and were somewhat dissatisfied with the results. Should you go looking for a better resistance box or a better voltmeter?

A more accurate voltmeter, assuming that the voltmeter and resistance box each contribute error in about the same way to the experimental result (more about this later.)
9. Random errors cannot be eliminated from an experiment. One simple way of finding out how large they are, however, is to repeat each measurement several times, trying not to let the result of any measurement be influenced by the results of previous measurements. The variations which show up in repeated measurements of a single quantity are a measure of the uncertainty involved in measuring that quantity, assuming all systematic errors are negligible (i.e., systematic errors which are not negligible compared to this uncertainty have been eliminated). If repeated measurements of a quantity all give the same value, it does not mean that there is no random error"on uncertainty in the measurements, but only that the measurements were not carried to a sufficient number of decimal places to observe random deviations. The experimenter should always estimate fractions of divisions down to the point where the last significant figure contains some uncertainty, if he wants to get maximum precision in his experiment.
10. Suppose a measurement of a particular quantity is repeated 5 times and each time the result is $x=10.0$ units. If the measurements are made with a device which may have a systematic error of $5 \%$ and you wish to increase the accuracy of your measurement, should you try to estimate another decimal place in the readings or calibrate your device using a method having a smaller systematic error?

Since the systematic error of $5 \%$ gives an uncertainty of $\pm 0.5$ in the measurements there is not much point in determining the (much smaller) random error uncertainty by

12. In the following frames let's suppose that systematic errors are negligible compared to random deviations (i.e.g corrections have been made to eliminate nost of the systematic error). If this is the case then the best way to determine a quantity accurately is to repeat the measurement several tirnes and take an average. Individual trials will deviate fron this average (deviation $d_{i}$ of value $\dot{x}_{i}$ from average $\bar{x}$ is $d_{i}=x_{i}-\bar{x}$, but the average is the best value obtainable from these measurements. One way of indicating the amount of uncertainty present in the individual measurements is to calculate the average of the absolute value of the deviations.

where $n$ is the number of individual measurements in the set.
13. Suppose a certain measurement is made 5 times with the result

| $x_{1}=10.90$ units | $x_{4}=10.10$ units |
| :--- | :--- |
| $x_{2}=10.35$ | $x_{5}=10.55$ |
| $x_{3}=10.30$ |  |

What is the average and the average deviation of these measure-

14. The average deviation in a set of repeated measurements of a quantity is one way of specifying the uncertainty present in one measurement. Another way is to determine the standard deviation:
$\sigma=\left[\frac{d_{1} 2+d_{2} 2+d_{3} 2+\ldots . . . d_{n}{ }^{2}}{n-1}\right]^{\frac{1}{2}}$
where $n=$ no. of deviations $d_{i}$
= no. of individual measurements in the set
15. Determine the standard deviation in the measurements of frame 13.
$\sigma=\frac{(0.46)^{2}+(0.09)^{2}+(0.14)^{2}+(0.34) 2+(0.11)^{2}}{5-1}=0.30$
16. If in a large number of trials, the values are distributed about the average value in a way which we describe as a "normal" or "Gaussian" distribution (i.e., density of values drops off symmetrically on either side of the average-_ a bell shaped density curve), then $68 \%$ of the values are between $x+\sigma$ and $x-\sigma$, and $95 \%$ between $x+2 \sigma$ and $x-2 \sigma$. In other words, a single measurement has a $68 \%$ chance of being between $\bar{x}+\sigma$ and $\bar{x}-0$, and a $95 \%$ chance of being between $\bar{x}+20$ and $\bar{\gamma}-20$. Thus the standard deviation $\sigma$ is a measure of the uncertainty present in a single measurement. The smaller the standard deviation, the narrower the limits between which a single measurement is likely to appear.
17. In the situation of frame 13, suppose you are about to make another measurement. Between what limits can you predict that it will appear with a $68 \%$ probability? Between what limits can you predict that it will have a $95 \%$ chance of appearing? (Note: The limits between which it is absolutely certain, that is, $100 \%$ certain, to appear are probably $+\infty$ and : $-\infty$ or 0 and $\infty$ depending on the nature of the measurement.)
$68 \%$ chance of appearing between 10.14 and 10.74
$95 \%$ chance of appearing between 9.84 and 11.04
18. After having made several measurements of a quantity we are not generally as interested in the uncertainty in a single measurement as we are the uncertainty in the average of the set. One way of determining the uncertainty in the average would be to take a large number of data set's and look at the distribution of the averages. These averages are more reliable (i.e. less uncertain) than the individual measurements of a set and therefore won't spread out as far. The standard deviation in the mean (average) $\sigma_{m}$ is a measure of the uncertainty in the average and can be determined from the standard deviation $\sigma$ (uncertainty in a single measurement) by the equation:

$$
\sigma_{\mathrm{m}}=\frac{\sigma}{\sqrt{n}}
$$

where $\sigma_{m}=$ the standard devidtion of the averages of $n=$ the number of measurements or trials in one set.
$\sigma=$ the standard devidtion of the individual trials in one set.
19. Determine the standard deviation in the mean in the situation of frame 13 .

$$
\sigma_{\mathrm{m}}=\frac{0.30}{\sqrt{5}}=0.13
$$

20. The result of repeated measurements of a single quantity is generally stated

$$
x=\bar{x} \pm \sigma_{m}
$$


22. Assuming that the measurements of frame 13 were made using an instrument that could only be read with $5 \%$ accuracy, state the results of these measurements in the form $x=x+$ -.

```
5% of 10.44 is about 0.52
Estimated uncertainty (random and systematic) = 0.6
x = 10.4\pm0.6
```

(Note that one significant figure has been dropped in both numbers because the uncertainty is only estimated and it would be a bit optimistic to carry another place.)

TEST: (Answer the following):

1. Explain the difference between random error and systematic error.
2. Give two examples of each type of erfor.
3. How might one decrease the amount of systematic error in an experiment? Explain using a specific example.
4. Suppose the possible systematic error in a voltmeter reading is estimated at 5 volts and several voltage readings are taken as follows:

| $V$ | 100 | 101 | 101 | 99 | 102 | 100 | volts |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| Trial | 1 | 2 | 3 | 4 | 5 | 6 |  |

State the result of these measurements along with its uncertainty.
5. Suppose the above measurements were made using a highly accurate potentiometer with systematic error of less than 0.1 volt. State the result of these measurements along with its uncertainty.

## EXERCISE

Make a series of measurements of some fundamental quantity (i.e., mass, length, or time) using any measuring devices at hand. Determine the random error standard deviation in the mean $\sigma_{\mathrm{m}}$, and estimate the uncertainty due to systematic. error. State the result of your measurements along with the uncertainty in the result.

The following exercise is designed to introduce you to some methods of determining the uncertainty present in a numerical result which has been calculated from experimental data. Experimental errors are always present in measurements of any kind, and these errors contribute to errors in the results obtained when calculations are made using the measured quantities. The exercise will be of value to you only if you work through it carefully, following the directions below, and think about it while you are doing it. You will be tested on these methods later.

## Directions:

1. The correct answers to the questions are given below the questions and should be covered by a "mask" (piece of paper) until you have supplied a written answer to the question. Read each statement and question carefully and write your answer down (on another sheet of paper, not the exercise sheets) before looking at the printed answer.
2. Work carefully and directly through the exercise in sequence without skipping or browsing. If necessary you may refer back to earlier frames at any time.
3. If your answer turns out to be incorrect or incomplete, try to get straightened out by re-reading the pertinent information before going on to the next frame.

Given $\quad \sigma=\left[\frac{d_{1}{ }^{2}+d_{2}^{2}+d_{3}{ }^{2}+\ldots .+d_{n}^{2} \cdot}{n-1}\right]^{\frac{1}{2}}$

$$
\sigma_{m}=\frac{\sigma}{\sqrt{n}}
$$

answer the following questions.

1. Which one of these two quantities is a measure of the uncertainty, due to random error, in the average of a set of $n$ measurements?
$\sigma_{\mathrm{m}}$ is the standard deviation in the mean (average)
2. Suppose the possible systematic error in a voltmeter reading is estimated at 5 volts and several voltage readings are taken as follows:

| V | 88 | 90 | 89 | 90 | volts |
| :---: | :---: | :---: | :---: | :---: | :---: |
| Trial | 1 | 2 | 3 | 4 |  |

State the result of these measurements along with its uncertainty.

Average $\quad \bar{V}=89.25$ volts
$\sigma=0.96$
$\sigma_{\mathrm{m}}=0.48$
The systematic error of 5 volts is nearly 10 times as great which means that the calculation of the random error standard deviation $\sigma_{m}$ is meaningless.

Answer: $\quad V=(89 \pm 5)$ volts
3. Suppose the measurements of frame 2 were made with a very accum rate potentiometer which is supposed to have less than .01 volt systematic error. State the result of the measurement along with its uncertainty.

$$
V=(89.25 \pm 0.48) \text { volts (systematic error negligible) }
$$

IF YOU HAVE TROUBLE UNDERSTANDING THE PRECEDING ANSWERS, REVIEW "EXPERIMENTAL ERRORS I" OR ASK YOUR INSTRUCTOR ABOUT IT BEFORE GOING ON.
4. In the following frames $x_{1}, x_{2}, x_{3} \ldots \ldots .$. are the results obtained in measuring several different quantities experimentally. The uncertainties in these measurements are represented by $\Delta x_{1}, \Delta x_{2}, \Delta x_{3} \ldots \ldots .$. knowing how to calculate the uncertainty in the result of adding, subtracting, multiplying, dividing, etcetera, two or more quantities whose individual uncertainties are known.

Rule 1: The uncertainty in the result of addition or subtraction is the square root of the sum of the squares of the uncertainties of the separate terms.

$$
\Delta x=\sqrt{\left(\Delta x_{1}\right)^{2}+\left(\Delta x_{2}\right)^{2}+\ldots \ldots \ldots}
$$

5. Given $\begin{aligned} x_{1} & =7.20 \pm 0.30 \\ & x_{2}=2.30 \pm 0.20 \\ & x_{3}=5.10 \pm 0.05\end{aligned}$
determine $x=x_{1}-x_{2}+x_{3} \quad$ (value and uncertainty).

$$
\begin{aligned}
& x=(7.20-2.30+5.10) \pm\left[(0.30)^{2}+(0.20)^{2}+(0.05)^{2}\right]^{\frac{1}{2}} \\
& x=10.00 \pm 0.36
\end{aligned}
$$

(Note that uncertainties don't add directly as would known errors, since there is a possibility of a positive error in $x_{1}$ being partially cancelled by a negative error in $x_{3}$ and so forth.)
6. Rule 2: The percentage uncertainty in the result of multiplication or division is the square root of the sum of the squares of the percentage uncertainties of the factors. In other words in the multiplication or division of $x_{1}, x_{2}$, $x_{3} \ldots . .$. ,

$$
\frac{\Delta x}{x}=\sqrt{\left(\frac{\Delta x_{1}}{x_{1}}\right)^{2}+\left(\frac{\Delta x_{2}}{x_{2}}\right)^{2}+\left(\frac{\Delta x_{3}}{x_{3}}\right)^{2}+\ldots \ldots . \cdot}
$$

7. Given $x_{1}, x_{2}, x_{3}$ as in frame 5, find $\bar{x}=\frac{\left(x_{1}\right)\left(x_{2}\right)}{x_{3}}$ (value and uncertainty).

$$
x=\frac{(7.20)(2.30)}{5.10}=3.25
$$

$$
\begin{aligned}
& \frac{\Delta x}{x}=\sqrt{\left(\frac{0.30}{7.20}\right)^{2}+\left(\frac{0.20}{2.30}\right)^{2}+\left(\frac{0.05}{5.10}\right)^{2}}=0.097 \\
& \quad \Delta x=(.097)(3.25)=0.32 \\
& x=3.25 \pm 0.32
\end{aligned}
$$

8. In a calculation that includes both addition (or subtraction) and multiplication (or division) the calculation for uncertainty $x$ is broken into parts. For example in determining $\Delta x$ for the calculation $x=\left(x_{1}+x_{2}\right) x_{3}$ rule 1 would be used to detemine an uncertainty for the sum of $x_{1}$ and $x_{2}$, and then rule 2 applied to this uncertainty and $\Delta x_{3}$ to determine $\Delta x$ for the product.
9. Using the values for $x_{1}, x_{2}$, and $x_{3}$ given in frame 5 , determine the value and uncertainty in the result of the calculation

$$
x=\frac{x_{1}}{x_{2}}+x_{3}
$$

$$
\begin{aligned}
& x=\frac{7.20}{2.30}+5.10=8.23 \\
& \frac{\Delta x_{12}}{x_{12}}=\sqrt{\left(\frac{0.30}{7.20}\right)^{2}+\left(\frac{0.20}{2.30}\right)^{2}}=.096 \\
& \Delta x_{12}=(.096)(3.13)=0.30 \\
& \Delta x=\sqrt{(0.30)^{2}+(0.05)^{2}}=0.30 \\
& x=8.23 \pm 0.30
\end{aligned}
$$

10. One advantage of doing a calculation such as the above is that it shows which of the measured quantities, $x_{1}, x_{2}$, or $x_{3}$ is contributing most of the uncertainty to the result of the calculation. This in turn tells the experimenter which measurement to concentrate his attention on if he wants to improve the accuracy of his result.
(a) Does the quotient $\frac{x_{1}}{x_{2}}$ or the quantity $x_{3}$ contribute the most uncertainty to ${ }^{\text {\% }} \mathrm{z}$ ? (Look at the calculation for $\Delta \mathrm{x}$.)
(b) Does $x_{1}$ on $x_{2}$ contribute the most uncertainty to $\frac{x_{1}}{x_{2}}$ ? (look at the calculation for $\frac{\Delta \mathrm{K}_{12}}{\mathrm{~K}_{12}}$.)
(c) Which of the quantities $x_{1}, x_{2}$, and $x_{3}$ would you concentrate on if you wanted to improve the accuracy of $x$ ?
(a) $\frac{x_{1}}{x_{2}}$
(b) $x_{2}$
(c) $x_{2}$ first
11. Rule 3: The percentage uncertainty in the result of raising a quantity to the nth power is $n$ times the percentage uncertainty in the quantity. In other words, if $x=x_{1}{ }^{n}$,

$$
\frac{\Delta x}{x}=n \frac{\Delta x_{1}}{x_{1}}
$$

12. If $x_{1}$ is as given in frame 5 , find the value and uncertainty in:
(a) $x_{1}{ }^{3}$
(b) $x_{1}^{\frac{3}{2}}$
(a) $\frac{\Delta x}{x}=3\left(\frac{0.30}{7.20}\right)=0.125$

$$
x=(7.20)^{3}=373
$$

$$
\Delta x=(0.125)(373)=47
$$

$$
x=373 \pm 47
$$

$$
\text { (b) } \begin{aligned}
\frac{\Delta x}{x} & =\frac{4}{2}\left(\frac{0.30}{7.20}\right)=0.0209 \\
x & =(7.20)^{\frac{1}{2}}=2.68 \\
\Delta x & =(0.0209)(2.68)=0.06 \\
x & =2.68 \pm 0.06
\end{aligned}
$$

13. Although the preceding three rules apply to a great many calculations, there is a more general rule from which they are derived, and which applies to other calculations as well.

Rule 4: If $x$ is a function of $x_{1}, x_{2}, x_{3} \ldots \ldots$. , then the uncertainty in $x$ is

$$
\Delta x=\sqrt{\left(\frac{\partial x}{\partial x_{1}}\right)^{2}\left(\Delta x_{1}\right)^{2}+\left(\frac{\partial x}{\partial x_{2}}\right)^{2}\left(\Delta x_{2}\right)^{2}+\ldots \ldots \ldots \ldots \cdot}
$$

14. Suppose a certain angle $\theta$ is measured as
$\theta=(1.00 \pm 0.05)$ radians. What is the value and uncertainty of $\sin \theta$ ?

$$
\begin{aligned}
x & =\sin \theta=0.841 \\
x_{1} & =\theta \\
\Delta x & =\sqrt{(\cos \theta)^{2}(\Delta \theta)^{2}}=(.0540)(0.05)=0.027 \\
x & =\sin \theta=0.841 \pm 0.027 .
\end{aligned}
$$

1. An object travels a distance $L=(10.6 \pm 0.4)$ meters in a time $t=(4.3 \pm 0.2) \mathrm{sec}$. What is the average velocity (value and uncertainty)?
2. Two experimental quantities $\mathrm{x}_{1}=8.23 \pm 0.56$ and $x_{2}=6.54 \pm 0.42$ are to be added. What is the value and uncertainty of $x=x_{1}+x_{2}$ ?
3. What is the value and uncertainty of $x_{2} 2$ in question 2?
4. Given $\theta=(0.25 \pm 0.07)$ radians, find the value and uncertainty of $\cos \theta$ 。

Consider an experiment in which a quantity $x$ is measured $r$. times. Let each measured value of $x$ be denoted by $x_{j}$ where $j=1, \ldots, r$. : can compute the standard deviation of the measured values of $x$ which we will call $\sigma(x)$. $\quad$ e can also compute the standard deviation of the mean value $\bar{x}$ of $x$ which we will call $\sigma_{m}(x)$. This latter computation is simply $\sigma_{m}(\bar{x})=\frac{\sigma(x)}{\sqrt{r}}$ (so we are told). Ve wish to investigate here whether or not $\sigma_{m}(\bar{x})$ is in fact smaller than $\sigma(x)$ by the factor $\frac{1}{\sqrt{\bar{r}}}$. In order to do so we must obtain $\sigma_{m}(\bar{x})$ by some other presumably more direct method.

Suppose we perform the whole experiment stimes, each time obtaining a value $\bar{x}$. Let each value of $\bar{x}$ be denoted by $\bar{x}_{k}$ where $k=1, \ldots .$, . We can compute the mean value of $\bar{x}$ which we willcall $\bar{x}$. Here $\left.\overline{\bar{x}}=\frac{\sum_{k=1}^{s} \bar{x}_{k}}{s}\right]$. Ve can also compute the standard deviation of the values of $\bar{x}$ which we call $\sigma(\bar{x})$. Here $\sigma(\bar{x})=\frac{\left[\sum_{k=1}^{S}\left(\overline{\bar{x}}-\bar{x}_{k}\right)^{2}\right]^{1 / 2}}{\sqrt{s-1}}$.

Obviously $\sigma(\bar{x})$ is a direct determination of $\sigma_{m}(\bar{x})$.
Therefore by doing the experiment $s$ times we can directly measure $\sigma(\bar{x})$ and test whether or not $\frac{\sigma(\bar{x})}{\sigma(x)}$ is approximately $\frac{1}{\sqrt{r}}$.

To carry this study of uncertainty one final stepwhat is the standard deviation of the experimentally determined quantity $\frac{\sigma(\bar{x})}{\sigma(x)}$ ? (1) Uith the data obtained at this point we can at least calculate directly the standard deviation of the quantity $\sigma(x)(b u t$ not of $\sigma(\bar{x}))$. Call this $\sigma[\sigma(x)]$.

To generate data construct a simple pendulum of length ~lm from string and a small weight. Using your wrist watch or a stop watch, measure the time for the pendulum to execute ten full swings. Let this measured quantity be $x_{j}$ ( $t_{i}$ if you wish). Let $r=7$ and $s=5$.

Give some thought and planning to the organization of data tables before you begin collecting data.

## GRAPHICAL ANALYSIS

Often one of the aims of an experimental investigation is the determination, from measurements made in the laboratory, of how one of two interdependent quantities, $y$, depends on the other, $x$. Graphical methods provide us with a very useful tool in this type of analysis.

## I. Plotting Graphs

Suppose one is interested, for example, in finding in a particular experiment a mathemical relationship which expresses the velocity of a moving object $v$ as a function of the time $t$. In this case velocity is the "dependent variable" whose dependency on the "independent variable" time is to be established from the following data.
Time Velocity (magnitude)
$(\mathrm{sec}) \quad(\mathrm{cm} / \mathrm{sec})$
$1.00 \quad 1.9$
2.00
1.9
3.00
3.0
4.00
3.9
5.00
6.5
6.00
11.0

Suppose that in this experiment the time measurements are very precise and their errors can be ignored while the velocity measurements are estimated to have a standard deviation (see instruction sheet on "Measurement, Probability, and Experimental Errors ) of about $\pm 0.30 \mathrm{~cm} / \mathrm{sec}$. The steps to be followed in constructing a graph which illustrates the dependence of velocity $v$ on time $t$ (or any quantity $y$ on another quantity $x$ ) are summarized below.
(a) The dependent variable (quantity whose dependency on the other is to be determined) is plotted vertically (velocity versus time rather than vice versa).
(b) Scales should be chosen which are easy to plot and easy to read and which make the graph large enough to be read easily and accurately (occupying a full page if possible).
(c) Scales usually start at zero but sometimes this would cause the data to be crowded into one part of the graph. In such a case it is a good idea to suppress the zero (start the scale at some value other than zero or show a break in the scale). However, it should be made obvious to someone looking at the graph that the zero has been suppressed.
(d) The graph should have a title and each of the axes should show the quantity plotted along that axis and the numerical scale and units for that quantity.
(e) The experimental points are marked clearly on the graph by drawing a small circle around each of them and drawing an "error line" (in the above example extending $0.30 \mathrm{~cm} / \mathrm{sec}$ above and below the data point) at each point.
(f) Draw the simplest possible smooth line or curve (i.e. the simplest curve is a straight line, the next is a curve whose curvature is always in the same direction and doesn't change magnitude suddenly, etc) among the points, with no more details of shape and curvature than is justified by the size of the estimated exrors. If the magnitude of the standard deviations are estimated correctly and the line is drawn correctly the curve should cut about two thirds of the error lines (very roughly).

When these steps are applied to the example of the moving object given above, a graph results such as that shown in the following figure.

Velocity $V$ versus Time $t$

II. Determination of a Mathematical Relationship

If a graph of dependent variable y versus independent variable $x$ turns out to be a straight line, the dependence of $y$ on $x$ is expressed by the equation

$$
\begin{equation*}
y=a x+b \tag{1}
\end{equation*}
$$

The slope a and $y$ intercept $b$ of the line can be taken directly from the graph (see part III) thus establishing the relationship between quantity $y$ and quantity $x$ in this experiment.

If the graph of $y$ versus $x$ is curved, however, as it is in the case of the velocity of an object versus the time in part $I_{\text {, }}$ the quantities must be related by some other equation. For example, one might guess that $y$ is related to x according to an equation of the type

$$
\begin{equation*}
y=a x^{n}+b \tag{2}
\end{equation*}
$$

where $n$ might be an integer $-1, \pm 2, \pm 3, \pm 4 \ldots \ldots \ldots \ldots$........... $\quad \pm$ fraction $\pm 1 / 2, \pm 1 / 3, \pm 1 / 4, \ldots \ldots . . . .{ }^{2}+\ldots$ decide which values of $n$ are truly possibilities one should study the graph of $y$ versus $x$ and equation (2). In the case of the velocity versus time graph of part $I$, for example, negative values of $n$ should be immediately discounted since equation (2) would predict a decrease in $y$ for increasing $x$. Fractional values of $n$ are just as unlikely since as $x$ increases, the graph shows y increasing faster and faster (perhaps indicating $n=+2$ or, +3 , etc.) 。

To see if the velocity - time ( $y=v, x=t)$ data for the moving object example of part fits equation (2) with $n=+2$ one could graph $Y=v$ versus $X=t^{2}$ from the experimental values of $v$ and the corresponding values of $t^{2}$. If the graph of $Y$ versus $X$ from the data is a straight line, the experimental results fita relationship

$$
\begin{aligned}
& \mathrm{Y}=\mathrm{ax}+\mathrm{b} \\
& \mathrm{v}=\mathrm{a} \mathrm{t}^{2}+\mathrm{b} \quad \text { (equation } 2 \text { with } \mathrm{n}=+2 \text { ) }
\end{aligned}
$$

where a and $b$ are the slope and intercept of the line. If such a graph was not straight, but was straighter than agraph of versus $t$, then one might try a graph of $Y=v$ versus $X=t{ }^{3}$ and so on until a straight line was found. The same general procedure could be followed in cases where $n$ is thought to be a fraction or have a negative value. If the data are to be represented by the equation

$$
\begin{equation*}
y=a x-1 / 3+b \tag{3}
\end{equation*}
$$

then a graph of $y$ versus $x^{-1 / 3}$ should yield a straight line.

Another type of relationship between quantities which appears often is

$$
\begin{equation*}
y=A e^{a x} \tag{4}
\end{equation*}
$$

where $A$ and a are positive or negative constants. If equation (4) accurately represents the data, then

$$
\ln y=a x+\ln A
$$

$$
\text { or } \quad \mathbf{Y}=a x+b
$$

making the substitutions $Y=\operatorname{Ln} y$ and $b=\ln A$. Therefore if $Y=\ln y$ is plotted vertically against $x$ horizontally, a straight line of slope a and intercept $b=1 n A$ should result. The values of a and $A$ can be determined from this line.
III. Determination of Slope and Intercept

The slope and intercept of a straight line are found as follows: First the $x$ and $y$ coordinates of two widely separated points on the line are determined (note that the points must be widely separated for accuracy and the points are points on the line, not data points). The slope of the line is defined

$$
a=\frac{y_{2}-y_{1}}{x_{2}-x_{1}}
$$

and should have the same value (for a straight line) regardless of what two points are chosen. The y intercept is obtained by extending the line back to $x=0$ and noting the value of $y$ at this point on the line (this is the intercept b).

A more reliable determination of slope, $a$, and $y$ intercept $b$, results when one computes the slope and intercept of the straight line which minimizes the sum of the squares of the deviations of the data points from the line (see instruction sheet on "Method of Least Squares").

## References:

1. Kruglak and Moore, "Basic Mathematics for the Physical Sciences", chapter 7 .
2. G. Wootan, Inc. "Graphs"
3. Ford, "Basic Physics", section 7.6

## Introcuction

Fe know from our stuaies of the motion of cbjects in a uniform gravitational field, that when a mass in is thrown honizontally with some initiai veiocity vo itwil foliow a parabolic path (Fig. 1-a) and will fall a distance
$\frac{1}{2} \mathrm{E}\left(\frac{1}{V_{0}}\right)^{2}$ in the the it takes to

(a)
travel a distance horizontally. One can produce a unifom electric fiejd, by comenting a batrery between two paralie? motal plates as in Fig. i-b are 0 . If an electron is in the region betheer the two plates it is subject to a constent electric force, just as mass m in antiont gravitational field is subject to a constant gravitational
force. As a consequence of this, is the electron enters the field with $a$

(b)
velocity $v_{0}$ as indicatej in Fig. $2-6$, it


$y_{1}=\frac{1}{2} \quad \frac{\sqrt{r}}{\operatorname{ma}}\left(\hat{v}_{0}\right)^{2}$ in the time it takes
the electron to travel a distance homizontally. Here e is the charge on the electron, $n$ is the mass of the electron, dis the distance between the plates, ane $y$ is the difference or potertial between the plates. If $\epsilon_{\mathrm{a}} \mathrm{m}, \mathrm{d}, \mathrm{I}$ and $v_{c}$ are all held constant, then $y_{\text {, }}$ is directy proportional to $V$. When the electron ieaves the eiectric fieid Detween the plates, its path becomes a
straight line until it strikes some obstruction such as the screen in Fig. 2 . It is relatively easy to show that

$$
Y=\text { const } V
$$

ponvided $e, m, d, 1, v o$ and $D$ are held constant. Here $Y$ and $D$ are the distances shown in Fig. 2.

In an oscilloscope, electrons are accelerated until they attain some velocity vo and then pess in turn trmough two sets of plates as indicated in Fig. 今̂. TE there is any potential difference between the finst pait of plates, the electrons will be deflected vertically an emount propowtional to this potential difference; if there is any potential difference between the second peir the electrons will be deflectec horizontanly and amount proportionaj to this potential difference.

Fit. 3 shows schematically some of the essential Fectures of the oscilloscope. Any difference of potential applied to the input teminals of either the hooszontal or vertical input teminals is first multiplied by sone constart factor (determined by the setting of the amplifier controls) and then applied to the deflection plates. The vertical and horizontal position controis allow a potential difference to be appiied to the deflection piates even without an input: and thus permit positioning of the electron beam.


1. Plug in the oscilloscope and rotate the power switch clockwise to turn on the oscilloscope. Note the effect that this rotation has on the amount of scale -illumination.
2. Set the horizontal DISPLAY switch to the forizontal Amplifer Sweep Disableg position.
3. Set the TMTENSTTY, FOCUS, and both POSTTION KNOBS so that the white dot of each knob is at its highest point ()
4. As soon as the scope tas wammed up a smail spot should appear mear the center of the screen. Fote the effect of tuming the intensity knob clockwise anc counterciockwise. Set the intensity control so that the spot is barely yisible.
5. Note the effect of the position cortzols. Use them to set the spot to the center of the screen.


6. Reverse the connections at either the dry cell or the input terminals and note that the spot moves com $13_{5} \mathrm{~cm}$. Whenever the top (large) input teminal is highen in potential then the lower (smali) input teminal. the spot is deflected upward and viae versa.
7. Repeat steps 9 and 10 except use the horizontal + input temminals. Note that when the top (iarge) input terminal is hirher in potential than the lower temminal the spot moves to the right and vice versa.
8. Set the vertical sensitivity switch to 0.5 volts/om and reacenter the spot if necessary. Connect the 1.5 volt dry cell to the verticai input. Note that the spot is defiected approwimately 3 cm .
9. Repeat step 12 with different settings or the sensitivity switch, until you are confident you understand the function of this switch. Repeat, using the borizontal + input teminals, and different settings of the horizontal sensitivity switch.
10. Set the vertical sensitivity switch to i volt/on and rotate the sed sensitivity control so that its white dot is approximately vertical Reocenter the spot if necessary. Connect the 1 落volt battery to the 4 yertical input terainals and note that the spot is deflected only about 1 volt, instead of 2.5 cm . Remove the battery, rotate the red sensitivity control knob to its extreme counterclockwise position, re-center the spot, and re-connect the battery. Note that the spot is deflected oniy about 0.5 cm.
11. Pepeat step 13 with different settings of the red sensitivity control knob, until you are confident that you understand the function of this control. Repeat, using the horizontal if input teminala and the homizontal sensitivity control knob.
12. Use your oscilioscope to measure the potential difference between the terminals of the black box provided. Fosition the spot and choose the sensitivity setting so thar the deflection produced by the uninown voltage is as large as possible, but still on scale. Remember that it is only When the red knob sensitivity control is in its extreme clockwise position that the sensitivity is actuaily that marked on the sensitivity suritch.
13. Wire up the dircuit shown below.


scmentert

Stt the horizontal sensitivity suitch at 0.1 voltfom and position the spet so that it is at ( $-4,0$ ). Connect points A and E and note that the spot begins to move to the right. Note that it moves with approximately constant velocity over about the first three centimeters and then slows down as it moves along the rest of its path. Take another lead and short out the capacitor (i.e. connect the lead across the capacitor). Wote that the spot jumps back approximately to (-4.0). Remove the short and the spot again sweeps to the right. Short the capacitor. set the homizontal sensitivity switch to 20 mulem and adjust the position of the spot to ( -5.0 ). Now remove the short and note that the spot moves across the entire screen at very nearly constant velocity. When at aeaches the right end of screen, short the capacitor momentarily and note that the spot jumps back to $(-5,0)$ and then besins moving to the right at constant speed. A circuit which produces this type of norizontal motion of the spot is called a "sweep" circuit. All osciloscopes heve a builto in sweep circuit which may be used to produce this type of horizontal motion of the spot. Disconnect your circuit. Set the biack LEVEL control switch to the AuTO. position.

IB. Set the HORIZONTAL DTSPLAY sWitch to the [SWEEP NORMAE] pOsition. Set the ShEEP TME/CM switch to 1 sec., and $L$ (xd) set the red sweep tinelon control knob to its extseme clockwise position. Note that the spot moves horizontally to the right at a constant speed, then jumps quickly to the left edge of the screen and begins moving to the right again at constant speed. If necessary. adjust the position control so that the spot starts from (-5.0). Use youn clock to detemmine how long it takes the spot to cover 10 cm. It should be close to 10 secs. since the switch is set at 2 sec, which means the spot should take 1 sec. to traverse each centimeter. Set. The SWEEP TIME/Cm switch to .S secs. and use your clock to determine how much time it takes for the spot to move 10 om . It should De very nearly 5 secs., since with the switch at. S, it should take .5 sec . for the spot to travel each cm. Concinue to experinent with different settings of this switch until you are confident that you understand the Function of this switch Note, that when this switoh is set at, say, 2 m sec ( 2 min isec $=2 \times 10^{-3}$ secs.) one can. no longer observe the motion of the spot. All cre observes is the path of the beam as it moves to the right. With this setring, the spot moves to the right such that it takes .002 seconds to traverse ach centineter.
15. Set the SWEEP THEP/cm switch to 1 sec. and zet the red sweep timelom control knob approximately vertically Measure the time for the spot to move 10 cm . (Note that it now requires more than 1 sec. for the spot to cover 1 cm . Repeat with this red control knot set at its extreme counterclockwise position. Continue to experiment with different positions of this control until you are confident that you understand its function. Remember that it is oniy when this controi is at its extreme clockuise position that the sweep speeds are those indicated by the switch position.
20. There are three small black tromposition switches near the top Fight hand edge of the oscilloscope marked SLOPE, COUPLINE AND SOURCE. Set the Eirst of these to its tposition, the second to DC, and the thind to INT. Set the black level control knob so that the white dot is approximately vertical

- Set the Vertical Sensitivity
switch to. 5 volt/cm. Set the SWEEP TMME/Cm switch to 2 mee. Conmect the small box masked pulsif to the vertical input teminals, as indicated below.


The pulser is a device which produces a potential difference between its two terminals which vanies with times as indicated in fig. 2. Part of this pattem should be observed on the sereen of the oscilloscope. It may be necessary to turn up the intensity (brightness) control; and/or to motate the level controi knob siighty dockwise from its vertical position. The function of the level control is to stant the spot Boving horizontaily exactiy $a t$ the fime when the pulses is emitting a pulse. When adjusted correctiy, the pattern appears to be sta= tionary. While the sweep circuit is moving the spot from left to wight at constant speed, the voltage from the puiser is denlecting it verticaly, so one obtains a plot of the pulser voltage as a function of time. The first time the spot moves horizontally acrose the screen, one gets a plot of the first five pulses emitted by the pulseri the next time the spot moves across the screent one gets a plot of the next five pulses emitted by the puiser, ete. Since each group of five pulses enitted look exactiy the same as every other group of five pulses, the pattern appeans to be stationary. Note that if the level control knob is roteted too far ciockwise on counterclockwise from its vertical position, the patem will
chsappear. Wote the appearance of the pattern when the SWEEP. MTME/ ca switch is at 1 m sec. .5 m sec. . 1 m sec. 50 H sec. 20 Hzec . and 10 ksec.
21. Determine the time interval between the pulses emitted by the pulses. Use any converient setting of the SWEEP mimplam sutroh. Make your measurement as precise as possibie by using as much of the horizontal scaia as possible.

Object To compare the behavior of a real pendulum with the presdiction based on a model.

## Discussion

The real pendulum consists of a metal cylinder hung by a string from a fixed support and free to swing in a calibrated arc. A lamp, lens, photomcell, timer, and oscilloscope make it possible to determine both the velocity of the cylinder at the bottom of its swing, and the time for one half a cycle.

The model consists of a particle of mass $m$, hung from a fixed point by a string of length $L$, free to swing in a vertical plane, and free of all dissipative (frictional) forces. If the particle is released from rest from some point such as $P$ (Fig. 1) then it is easy to show the velocity it will have when it reaches 0 is given by

$$
\begin{equation*}
v_{0}=\sqrt{2 g L(1-\cos A)} \tag{1}
\end{equation*}
$$

To calculate the time for the particle to go from 0 to $Q$ is more difficult. We could get a first approximation to this time by using one half of $v_{0}$ as the average velocity over the path $0 \quad Q$, that is

$$
t=\frac{\operatorname{arc} 00}{\left(\frac{v_{0}}{2}\right)}=\frac{L A}{\frac{v_{0}}{2}}
$$

A better approximation would be to first divide the path 0 Q into a number $N$ of equal pieces, labeled

$$
\Delta \Delta_{0}, \Delta \Delta i, \Delta A_{2}, \quad \Delta \Delta \frac{1}{1}
$$

/indicated in Fig. 2. As an approxi-
 mation we could say that the time $t_{0}$ it takes to traverse the distance $\triangle$, would be

$$
\Delta t_{0}=\frac{\Delta \Delta}{\left.v_{0}+v_{1}\right) /=}
$$

where $v_{0}$ is the velocity at point 0 and $v_{1}$ is the velocity at point 1. Similarly, the time to traverse the $i$ th segment $\Delta L_{i}$ would be

$$
\Delta t_{i}=\frac{\Delta \Delta i}{\left(V_{i}-\sqrt{2}+1\right) / Z}
$$

The total time to go from 0 to $Q$ is then

$$
t_{0 \rightarrow \infty}=\sum_{i=0}^{N-i} \Delta t,=\sum_{i=0}^{N-1} \frac{\Delta \Delta i}{\left(V_{i}+v_{i+1}\right) / 2}
$$

This approximation improves as $N$ is made larger, and one can get as close to the correct answer as one desires by making $N$ sufficiently large.

Using the principle of conservation of total mechanical energy one can easily show that

$$
\begin{equation*}
v_{i}=\sqrt{2 g L[\cos (i \Delta \theta)-\cot A]} \tag{2}
\end{equation*}
$$

Since all the $\Delta \Delta$ are the same size and equal to $L=L \frac{A}{N}$ the time to go from 0 to $Q$ may be written


The electronic timer or clock is controlled by a light beam and photo cell. The timing system operates in two modes determined by the position of the "Hold Mode" switch. When this switch is set in the "Gate" position, the clock will run only during the time the light beam is blocked. When this switch is in the "Pulse" position, the clock will start when the beam is first interrupted and will stop the next time the beam is interrupted. If the system is in the "Pulse" mode, and the cylinder is released from a point such as $P$, its first passage through 0 will start the clock, and its next passage through 0 will stop the clock. The clock reading will then be the time the bob takes to go from 0 to $Q$ and back or twice $t$.

The photo cell is a resistive element whose resistance increases as the intensity of the light falling on it decreases. Consequently, if it is placed in a circuit such as shown in Fig. Ja then as the pendulum starts through the light beam, the voltage across $R$ will begin decreasing. This decrease is used to initiate the scope sweep, so that one observes on the scope screen the manner in which the voltage across $R$ varies as the pendulum passes through the light beam. The pattern observed on the screen is shown in Fig. Bb. The positions of the cylinder relative to the light beam that corresponds to points $S$ and $T$ are shown in Fig. 4. Thus, the times,

indicated in Fig. $3 b$ is the time for the cylinder to travel a distance equal to the diameter of the cylinder. Dividing the diameter of the cylinder by $t$ gives $v_{o}$.

Procedure

1. Wire up the photocell and timer as indicated in Fig. 5
2. Read steps 3 through 6 and prepare (in your notebook) a suitable table for recording the data.
3. Set the "Hold Mode" switch to the "Pulse" position.
4. Use the electromagnet to position the cylinder at the point corresponding to $P$ with $A=70^{\circ}$. Press the "Reset" button to set the clock to zero. Release the cylinder, allow it to swing through point 0 to $Q$ and back throngh 0 . Catch the pendulum after it passes through 0 on its way to $P$. Fead and record the clock reading. The clock reading is in miiliscos ( $10^{-3}$ seconds).
5. Repeat steps 4 four more times.
6. Repeat steps 4 and 5 for $A=60,50,40,30,20,10$.

7. Wire u the photocell as indicated in Fig. 6 . Set the three switches marr the top right hand corner of the oscilloscope to - , DC, and IF": Set the "Level" control to 0 , the Sweep time/Cm switch to 1 millisec/cm., and the vertical gain switch to 0.5 volts/cm. (rite aweptime/cm (ontoul krob must be in cul position)
8. Start the pendulum swinging through about a $60^{\circ}$ arc and note if the passage of the bob through the light beam produces a trace like that in Fig. 3 h. If not, adjust the "Level" control and/or the vertical gain until such a trace does occur.
9. Use the electromagnet to position the cylinder at the point corresponding to $P$ with $A=70^{\circ}$. Release the cylinder and note the positions of pointsT. and $S$ of the scope pattern produced by the interruption of the light beam. $\stackrel{+}{+}$ Record the sweep time/cm. setting.
10. Repeat step 9: several more times.
11. Repeat steps 9 and 10 for $A=60^{\circ}, 50^{\circ}, 40^{\circ}, 30^{\circ}, 20^{\circ}, 10^{\circ}$. For each setting, use the most appropriate sweep time/cm setting.

## Analysis

1. Determine the average and average deviation of each set of five time readings taken for each fixed value of $A$. Calculate the average and average deviation of your measurements of the length of the pendulum and the diameter of the cylinder. Use the average deviation as a measure of the uncertainty in the measurement of that quantity.

进 THE Usallowire trace always itorts frome the same puint if the position ioutiols are not distwhod. Yuu cam detwonine point $S$ by

2. For each fixed value of $A$ determine $v_{0}$ from your measurements of the time required for the pendulum to move through a distance equal to the diameter of the cylinder. Calculate the uncertainty in each of these values. (See Experimental Errors II, Section 6 and 7).
3. Calculate $v_{0}$ for each value of $A$, using equation (1). Make a table comparing these calculated $v_{O}$ 's with the experimentally determined $v_{o}{ }^{\prime} s$.
4. Write a FORTRAN program to calculate $t$ from equation (3) for $0 \rightarrow Q$
the values of $A$ used in the experiment. An $N$ value somewhere between 50 and 100 should give adequate precision. Use the average value of your measurement of the length of the pendulum for $L$ and $9.80 \mathrm{~m} / \mathrm{sec}$ for g . Have this program processed by the computing center. (Each group is to write their own program, not merely borrow one from another group.)
5. Make a table comparing the calculated $t$ with the experimentally measured values. $0 \rightarrow Q$
6. Use the principle of conservation of total mechanical energy to calculate the velocity of the particle at point P, assuming it was released from rest at point Q (See Fig. ) . From this expression derive equation (2).


Ref.: Halliday, Resnick, Sect. 32-8
A. Capacitor Discharge

If a capactior is charged so that charge $q_{0}$ is on each plate and then allowed to discharge through a resistor as shown below, the potential difference across the capacitor must equal that across the resistor (they are across the same two points in the circuit ab) at all times during the discharge.

$q=$ charge on plates at
$i=$ current at the same time

$$
V_{a b}=\frac{q}{c}=i R \quad i=\frac{d q}{-d t}
$$

which yields the equation
$\frac{q}{C}=-R \frac{d q}{d t} \quad$ or $\quad \frac{d q}{d t}=-\frac{1}{R C} \quad q$
Equation (1) shows that at any instant the capacitor will be discharging at a rate that is directly proportional to the charge still on its plates, and inversely proportional to the constant (RC). The fact that the discharge rate is determined by the product $R C$ may be used to measure the resistance $R$ indirectly by making measurements on the rate of discharge of a known capacitor $C$ through the resistance element. From equation (1), one can show that the charge $q$ still on the plates of a capacitor which began with charge $q_{0}$ and discharged for a time $t$, is

$$
\begin{equation*}
q=q_{0} e^{-t / R C} \tag{2}
\end{equation*}
$$

(Show that equation 2 is a solution of the differential equation 1 , and show that the time constant $t=R C$ is the time for the charge to fall to fraction $1 / e$ of its initial value $q_{0}$.)
B. The Ballistic Galvanometer

The moving element of this type of galvanometer consists of a rectangular coil which is suspended between the poles of a magnet by a fine wire. When a charge q passes through the coil, the forces exerted by the magnetic field on the moving charges produce a turning moment or torque on the coil. The torgue gives the coil an anzular momentum, but the coil has a relatively large moment
of inertia so that very little actual motion occurs in the time that it takes for charge $q$ to pass through the coil. The coil continues to rotate however, twisting the suspension wire. The twisted suspension wire now exerts a restoring torque which decreases the angular momentum of the rotation, brings the rotation to a stop at some angle $\theta$, and increases the angular momentum in. the opposite direction. The result would be oscillatory motion of angular amplitude $\theta$ as long as no mechanical energy was lost from the system (it is almost frictionless). It is not difficult to show that the angle $\theta$ is proportional to the charge $q$ which passed through the galvanometer coil. (Supplementary note.)

If one wishes to damp the oscillations of the galvanometer it is useful to recall that a coil rotating in a magnetic field generates an induced emf. (Chapter 35, Hall. \& Res.) Short circuiting the galvanometer terminals completes an external circuit so that this emf can cause a current flow in the low resistance short-circuit. Thus the mechanical energy of the rotating coil is converted to electrical energy as in a generator, and this electrical energy is in turn dissipated as heat by the circuit resistance. The loss of mechanical energy by the system results in a quick damping of the oscillation.

## Experiment

The capacitor in the circuit below may be charged by connecting $A$ to $B$. It can then be discharged through resistance $R$ for a measured amount of time $t$ by connecting $B$ to $E$ for this time interval. The charge left on the plates at the end of this time interval may be determined by discharging the remaining charge through the ballistic galvanometer (connect $B$ to $F$ ), which then deflects an amount $D$ which is proportional to the charge $q$ that has passed through it.


One then gets a feeling for the way the capacitor charge decays as a function of time by plotting graphs showing
the galvanometer deflection $D$ as a function of discharge time $t$ and the natural logarithm of the galvanometer deflection as a function of time. The resistance $R$ may be determined from the slope of the latter graph (first prove that $l_{1 /} D=\ln D_{0}-\frac{1}{R C} r$ ).

One thing that should be considered in this experiment is the fact that since $R$ is large there may be other paths of comparable resistance through which the capacitor can discharge (even when $B$ and $D$ are not connected for example). Investigate this possibility before making any claims as to the accuracy of your result for the value of $R$.

Note: The capacitors have the following values for the different circuit boards:

$$
\begin{aligned}
& \text { No. } 1 \\
& \text { No. } 2 \\
& \text { No. } 3 \\
& \text { No. } \\
& \text { No. } 4 \\
& \text { N }=0.88 \\
& \text { No. } 5 \\
& \text { N }=0.88 \\
& C=0.88
\end{aligned}
$$

Supplementary note $\frac{\text { (ref.-Fund of Elec. and Mag., A.F. Kip, }}{\text { E }}$ McGraw-Hill, 1962)

## How It Works

When a short pulse of electric current $i=i(t)$ passes through the coil of a ballistic galvanometer question 1 - What happens?
answer 1 - It deflects through an angle $\theta_{\text {max }}$, corresponding to a deflection (scale reading) $D \propto \theta$
 and there exists a simple relation between $D$ and $Q=f i(t) d t$, the total charge passing through the coil.
question 2 - WHY? (and what is the relation)?
answer 2 - The parts of a galvanometer coil which are in a magnetic field are in a practically uniform radial field, as in Pig. 33-9 in Hall. and Res. (Beware - example 3 in the text discusses a galvanometer as it is often used to measure (steady) currents, not as we use a (ballistic) galvanometer in this lab exercise to measure charge in a current pulse. The mechanism of the instruments are basically the same, however.)

The magnetic torque on a galvanometer coil, when its plane is parallel to the magnetic field, as Fig. 33-9 (by eqn. 33-7)

$$
\tau=N \text { i } A B
$$

When, instead of a steady current $i$, we supply a short pulse of current such as shown above, the angular impulse given to the coil is $\int \tau$ dt, where the integration is over an interval covering the entire time during which current flows.

Integrating, Impulse $=\int \tau d t=N B A . \int i d t=N B A Q$.
This is the special feature of the ballistic galvanometer. It receives an angular impulse that depends only on the total charge pulse flowing through it and not on the way the charge flow varies with time. The pulse is short compared with the natural period of the galvanometer coil. After the impulse ends the motion of the coll is 'ballistic' - under action of the suspension, only. From mechanics we have that this angular impulse gives the coil an angular momentum $I \omega_{0}$, where $I$ is its moment of inertia and $\omega_{0}$ is the angular velocity just after the pulse has passed. Thus $I \omega_{0}=I B A$ Q. The initial angular momentum $I \omega_{0}$ corresponds to an initial kinetic energy $\frac{1}{2} I \omega_{0}$ 2. As the coil turns against the restoring torque of the galvanometer coil suspension, this kinetic energy is converted into potential energy. At the maximum deflection $\theta_{\max }$,
conversion is complete and we may write

$$
K E=\frac{1}{2} I \omega_{0}^{2} \quad \mathrm{PE}=\frac{3}{3} k \theta_{\max }{ }^{2}
$$

where $k$ is the torque constant of the suspension (text, see torsional pendulum). Solving for $\omega_{0}$ and substituting
in $K E=P E$ gives $Q=\left((I k)^{\frac{1}{2}} / \mathrm{NBA}\right) \cdot \theta_{\max }$, or $Q^{\propto D}$.
Tell in your report how you might verify this last relation experimentally.

The potentiometer is a widely used device for measuring D.C. voltages (potential differences). It is not as convenient to use as a voltmeter, but it has the advantage that it is more precise, and it doesn't affect the circuit to which it is connected. The principle of operation of a potentiometer can be understood from the following circuit. A 1 -meter long uniform

wire, $A B$, is stretched along a meter stick and a known current is arranged to flow through this wire by means of a battery, $B^{\prime}$, a variable resistance $R$, and an ammeter. Let us suppose, for simplicity, that the 1 -meter wire has a total resistance of exactly 1 ohm and the current in the wire is exactly 1 ampere. Then, according to Ohms law, the difference of potential between points $A$ and $B$ would be exactly 1 volt, the difference of potential between $A$ and $C$ exactly 0.5 volt, the difference in potential between $A$ and $D$ exactly 0.8 volt , and so on. Since (conventional) current always flows in a resistance from a higher to a lower potential, point $A$ is 1 volt higher in potential than point $B$, is 0.5 volt higher in potential than point $C$, etc.; in fact, $A$ is higher in potential than any other point of the uniform wire. Now suppose we have a battery whose emf is exactly 0.5 volt. Now the positive pole of a battery is always higher in potential than the negative pole by an amount equal to the emf of the battery, so for our battery of emf equal to 0.5 volt, the positive pole is 0.5 volts higher in potential than the negative pole. Suppose we connect the positive pole to point A, as in Fig. 2. After we make this connection, point A and

the positive pole of the battery are at the same potential since no current is flowing in the lead connecting $A$ and the positive pole. The negative pole is always 0.5 volts lower in potential than the positive pole, point $C$ is 0.5 volts lower in potential than point $A$, hence point $C$ and the negative pole of the battery are at the same potential. If we connect the negative pole to point $C$ via a sensitive ammeter, $G$, as indicated by the dotted lines, no current will flow, since current only flows when there is a difference of potential. If the emf of the battery had been 0.8 volts, we could have connected the negative pole to point $D$ and no current would flow in the galvanometer. If we have a battery whose emf is unknown we can move the slider $S$ along the wire until we reach some point, $U$, where the galvanometer reads zero. Then we know immediately that the difference of potential between the battery terminals is exactly that between $A$ and $U$, which can be read directly from the meter stick. This is the principle of operation of a potentiometer.

It should be apparent that with the above arrangement, one could not measure emfs greater than 1 volt, since the difference of potential between the two ends of the wire is only 1 volt. To measure emfs between 0 and 2 volts, we could decrease $R$ until the current in the wire were exactly 2 amps. Now the difference of potential between $A$ and $B$ would be 2 volts, that between $A$ and $C, 1.0$ volt, and that between $A$ and $D, 1.6$ volts. We could proceed exactly as before to determine the unknown emf. It should be clear that by properly choosing the current that is to flow in the uniform wire, one can measure any size emfs.

It turns out that the current in the uniform wire can be adjusted to the desired value more precisely by means of what is called a standard cell (battery) than by using an ammeter. Suppose we have a battery whose emf is known to be exactly 0.500 volts, and suppose we wish to adjust the current in the wire to be exactly 2 amperes. We know that if the current were 2 amperes, the difference of potential between $A$ and the 25 cm point on the wire would be exactly 0.5 volts. We connect the positive pole of our battery to point $A$, and the negative pole via a sensitive ammeter

(galvanometer), $G$, to the 25 cm point of the wire as in Fig. 3 . Now we adjust $R$ until the galvanometer reads zero. When this is the case, the potential difference between $A$ and the 25 cm point must be exactly the same as the difference of potential between the poles of the battery, namely 0.500 volts; hence, the current in the wire must be exactly 2 amperes. Suppose the emf of the standard cell were 1.35 volts. To what point of the wire should the sliding contact $S$ be connected if it is desired to adjust the current in the wire to 2 amperes?

The potentiometer you will use in this experiment differs from the one described above in that the uniform wire, $A B$, is wrapped as a helix around a cylinder which is then enclosed in a plastic case. The position at which the slider $S$ makes contact with the wire is indicated by a dial mounted on the top of the cylinder. When the dial reads zero, the slider is at the end of the wire corresponding to point A, and when the dial reads 1000 , it is at the end corresponding to point $B$, when it reads 500 it is at a point corresponding to point $C$, and so on. The circuit diagram of the entire set up is shown in Fig. 4. The battery B' is actually a regulated power supply which must be plugged in to an AC outlet to function. The variable resistance $R$ is enclosed in a green plastic cylinder and its resistance is varied by turning the knob at one end of the cylinder. The standard cell is a flat cylindrically shaped mercury cell having an emf approximately equal to 1.35 volts. (The exact value is written on the cell). The switch SW simply makes the process of calibrating of the potentiometer (i.e., adjusting the current to the proper value) more convenient. The swtiches marked $C$ and 0 are provided to prevent a large current being drawn from the standard cell. Switch $C$ is used to obtain a preliminary balance, and then swtich 0 is used for the final balance. The proper technique is to tap the switches rather than hold them closed for a long period of time.

## Procedure

1. The resistance of the uniform wire is 100 ohms. If the current in this wire is adjusted to be .020 amperes, then the potential difference between end $A$ and $B$ would be 2 volts. If this were actually the case, then there would be some point of the wire which would differ in potential from $A$ by exactly the standard cell voltage. Calculate the dial reading that would correspond to this point and set the dial to that value (e.g., if your standard cell is 1.350 volts, the dial should be set to 675 ). Rotate switch $\mathrm{SW}_{2}$ to the S.C. side and adjust the resistance $R$ until the galvanometer indicates zero current when switch 0 is closed momentarily. Always use switch $C$ to obtain a preliminary balance before using switch 0 . Your potentiometer is now calibrated, and there is a difference of potential of 2 volts between the two ends of the wire.

2. Rotate SW to the UNK position. Connect a dry cell to the terminals marked $X$, being careful to connect the positive (center) terminal of the battery to the terminal marked $t$. Now move the slider until a point of the wire is reached such that the galvanometer doesn't deflect when 0 is held down momentarily. (Again, use the switch marked $C$ to obtain a preliminary balance). Note the reading on the dial when a balance is obtained, and from this calculate the emf of the dry cell. It should be somewhere in the neighborhood of 1.5 volts.
3. Reverse the connections at the dry cell so that the positive terminal of the battery is connected to the negative terminal of the potentiometer. Using only the $C$ switch, note that it is impossible to find a position of the slider which will produce zero galvanometer deflection. Is the galvanometer deflection smaller when the slider is at the zero end of the scale (point A) or when the slider is at the 1000 div mark (point B).
4. Connect two dry cells in series and then connect the combination to the potentiometer terminals, as indicated below. This combination has an emf of about 3 volts


Using only the $C$ switch, note that it is impossible to find a position of the slider which will produce zero galvanometer deflection. Is the galvanometer deflection smaller when the slider is at the zero end of the scale (point A) or when the slider is at the 1000 div mark (point B).
5. There are essentially only two conditions in which it is impossible to get a balance on a potentiometer; either the point of higher potential of the unknown emf is not connected to the + terminal (point $A$ ) of the potentiometer or the unknown emf is too large. You encountered the first case in step 3 and the second case in step 4. Note that it is possible to determine which condition you have by noting whether the galvanometer deflection increases or decreases as the slider is moved from the zero end (point A) to the 1000 div end (point B). There is a special case of the second condition which is worth noting, namely, the situation in which there is no current in the uniform wire, a condition easily obtained by failing to plug in the power supply. In this case, all emfs will appear to be too high even if they are correctly connegted.

A black box is provided which has an emf somewhere between 0 and 2 volts. The positive terminal is not indicated. See if you can with your potentiometer determine which is the positive terminal. After you have done this, determine the unknown emf.
6. The potentiometer was calibrated in step 1 so that is essentially direct reading. Once you obtain a balance for an unknown emf, you need only to multiply the dial reading by 2 to obtain the value of the unknown emf (of course you have to decide where the decimal point goes). This manner of calibrating the potentiometer is convenient, but not necessary. All that is required is that the current in the uniform wire be constant and of such a magnitude that one can obtain a balance both with the standard cell and with the unknown emf. To illustrate this case, turn the knob which varies $R$ a couple of turns in either direction. There is now some unknown current flowing in the wire. Connect the dry cell that you used in step 2 to the $X$ terminals of the potentiometer, and move the sliding tap until the galvanometer reads zero when switch 0 is held down momentarily. Record the reading of the dial and let us refer to it as $D_{\text {. N }}$. Now rotate switch $\mathrm{SW}_{3}$ to the SC position and adjust the sliding $x a p$ until the galvanometer doesn't deflect when 0 is held down momentarily. Record this dial reading and let us refer to it as $\mathrm{D}_{\mathrm{Sc}}$. Now calculate the emf of the dry cell from

$$
\frac{\text { emf of dry cell }}{D_{x}}=\frac{\text { emf of standard cell }}{D_{s c}}
$$

This computed value should compare favorably with the value for the emf of the dry cell that you calculated in step 2.
7. Using the technique suggested in step 6 , make three more measurements of the emf of the dry cell, each using a different unknown current. Calculate the average of all five values and the standard deviation of the mean.

$$
E \quad R_{2}
$$















$$
E
$$














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OBJECT: To study the conditions necessary for the production of standing waves in a string.

THEORY: When a string of mass per unit length $\mu$ and under tension $T$ is disturbed in a regular way, waves travel in both directions at a speed

$$
\begin{equation*}
c=\sqrt{T / \mu} \tag{1}
\end{equation*}
$$

If these waves come to point in the string where it is not free to move they are reflected and travel back in the opposite direction, combining with any "incoming" waves to produce the resultant displacement of the string at each point. Under certain conditions the waves may be produced in such a way that the incoming and reflected waves are sine curves of wavelength $\lambda$. In this case the resultant shape of the string will also be a sine curve of wavelength $\lambda$. However, this resultant sine wave remains fixed in position longitudinally while the amplitude fluctuates, rather than moving to the right or left as its component traveling waves do. The wave is therefore called a "standing" or "stationary" wave. Successive "snapshots" of the string near a fixed point $P$ might look as shown below.


Certain points $N$ along the string are stationary at all times and are called nodes. Halfway between the nodes are the antinodes $A$ where the vibrations have the largest amplitude. The distance between two successive nodes (or two antinodes) is equal to half a wavelength. In the preceding figure the point $P$ must of course be node since it is not free to move.

Suppose that a string of length L, fixed at both ends and under tension $T$, is disturbed in such a way that sine waves travel along it in both directions and are reflected at each end. Standing waves such as those described above can result only if the wavelength $\lambda$ is of such a magnitude as to make the two fixed ends of the string nodes. This means that standing waves can occur only if they (and their component traveling waves) have a wavelength $\lambda$ such that $L$ equals an integral number of half-wavelengths.

$$
\mathrm{L}=\mathrm{n} \quad \lambda / 2 \quad \mathrm{n}=1,2,3, \ldots .
$$

$o r$

$$
\lambda=2 L / n
$$

Since the speed of the traveling wave components $c$ is related to their frequency and wavelength by the expression $c=f \lambda$, one can say that standing waves may be set up in the string by vibrations of a number of different frequencies

$$
\begin{equation*}
f_{n}=c / \lambda=n c / 2 L \tag{2}
\end{equation*}
$$

Making use of equation (1), the frequencies $f_{n}$ which give standing waves, are related to the tension $T$ in the string by the equation

$$
\begin{equation*}
f=\frac{n}{2 L} \sqrt{\frac{T}{\mu}} \tag{3}
\end{equation*}
$$

For a given tension $T$ the lowest allowed frequency ( $n=1$ ) is called the fundamental or first harmonic, the next highest frequency ( $n=2$ ) the second harmonic, the next ( $n=3$ ) the third harmonic and so forth. If the string is disturbed in a regular but arbitrary fashion the resultant wave is composed of a combination of sinusoidal waves each having one of the allowed frequencies, rather than a standing wave of a single frequency.

In this experiment one end of the string is attached to a rod which vibrates sinusoidally at a single fixed frequency (the amplitude is small enough that the end of the string may be considered fixed). In general, waves are set up with no well defined nodes and small amplitudes but when the tension $T$ is adjusted so that equation (3) is satisfied for an integral value of $n$, standing waves of large amplitude and definite, although not perfect, nodes result. The tension in the string is varied to obtain standing waves for several different values of $n$.


## INSTRUCTIONS:

(1) Arrange the apparatus as shown in the figure.
(2) Vary the tension by adding or subtracting weights at the pulley end. Find and record those values of T (in newtons) for which well defined standing waves of maximum amplitude are set up. For each standing wave determine the distance between a node near the pulley and a node near the vibrating rod and divide this distance by the number of "loops" between these points to obtain $\lambda / 2$. Make a table showing the values of $T$ and $\lambda$ for the standing wave modes. (Note: to obtain some of the smaller tensions that will be required, it will be necessary to use the 5 gram weight hanger provided. Use this smaller hanger only for weights of less than 75 grams.)
(3) With the string under one of the tensions that produced standing waves in (2), cut out a section l meter long and determine its mass in kilograms. Record this mass per unit length $\mu$.
(4) From equation (1) and the fact that the speed of a traveling wave is equal to the product of its frequency and wavelength, show that for a fixed frequency the quantity $\lambda^{2}$ is directly proportional to the tension $T$ in the string. Plot a graph of $T$ versus $\lambda^{2}$ using the corresponding values of $T$ and $\lambda$ from part (2). Determine the frequency from the slope of the curve.

## EXERCISE:

Make estimates of the precision of $T, \mu$, and $\lambda$ and calculate the corresponding precision in the value of $f$ obtained.


Wow does your Data compare with that predleted by theol? a (ample: (Courtesy of Holcomb b Wright).
tara whist produce
max. one me stinove (at)
$.069 T_{5} \quad n=5$
$.176 T_{4} n=4$
-284 Ta $0=3$
$.490 \quad T_{2} \quad n=2$
.127 T. no 1 pumdemestal frequency). Then $\frac{\pi}{4}$ will give the tension required to produce tho second Lanmoxic $(n=2)$.
calculate the tuna and see how it compares will the exp.

$$
\begin{array}{lll}
T_{2}=\frac{12127}{4} & \frac{\text { they }}{48} & .490 \\
T_{3}=\frac{1127}{9} & .125 & .284
\end{array}
$$

po pos see, our initial assumption is wrong. net, arums that a teprica of 1.127 nt. cooceapondo to $n=2$, the second homwtorie.

| .069 | $T_{6}$ | $n=6$ | $T_{3}=\frac{4}{9} T_{2}$ | .500 |
| ---: | :--- | :--- | :--- | :--- |
| .176 | $T_{5}$ | $n=5$ | .490 |  |
| .284 | $T_{4}$ | $n=4$ | $T_{4}=\frac{9}{16} T_{3}$ | .282 |
| .490 | $T_{3}$ | $n=3$ | $T_{5}=\frac{11}{25} T_{4}$ | .180 |
| 1.187 | $T_{2}$ | $n=2$ |  | .284 |
| $?$ | $T_{1} n=1$ |  |  | .176 |

Thin time our assumption is concept WV Lat tension weill produce the fendemextal mode or frequency?
By comparing the values obtained in the laboratory unit those obtained from theory, perhaps a better undenitamding of the experiment uni cenubt.

Introduction
Fig. 1 shows 2 equal point charges, both positive. Points $M$ and $N$ are positioned as indicated. Which point, $M$ or $N$ is at the higher potential? Now refer to Fig. 2. Which point, M or $N$ is at the higher potential?


$$
\text { Fig } 1 .
$$



Suppose current is flowing through a rectangularly shaped conductor as indicated in Fig. 3. As indicated in the diagram (conventional) current is


FIG 3
NO MAGNETIC FIELD

$\varepsilon$

FIG 4

$$
\begin{aligned}
& \text { MAGNETIC FIELD B } \\
& \text { DIRECTED INTO PLANE } \\
& \text { OF FIGURE }
\end{aligned}
$$

flowing from right to left in the sample. Let us assume that in the sample the current is due to positive charges flowing from right to left as indicated. Since like charges repel each other, the positive charges will be arranged approximately uniformly across a given cross-section of the conductor, as indicated. Because of this uniform arrangement, one would expect that two points such as $M$ and $N$ would be at the same potential. Suppose that we now establish a uniform magnetic field of magnitude $B$ which is directed into the plane of the figure. Since the positive charges are moving with a velocity $\vec{v}$ to the left, there will be an additional force on them given by

$$
\vec{F}=q \vec{v} \times \vec{B}
$$

Since $q$ is + , the right hand rule indicates that this force is directed downward. As a result of this magnetic force, the moving charges are no longer distributed uniformly over a given crosssection but will be crowded into the lower portion of the sample as suggested in Fig. 4. As a consequence we would now expect $N$ to
be at a higher potential than M. Experimentally, one finds that for some samples this is indeed the case; however, for other samples just the reverse is true, i.e. one finds $M$ to be at the higher potential, even though the direction of the conventional current in the sample and the direction of
$\vec{B}$ are the same. One concludes that for those samples for which $\mathbb{N}$ is higher in potential than $M$, that conduction in the sample is actually due to positive carriers moving in the direction of the conventional current. For those samples in which $M$ is higher in potential than $N$, one concludes that conduction in the sample is due to negative charges moving opposite in direction to the conventional current. (You should be able to convince yourself that negative charges moving from left to right would also be forced downward when a magnetic field is applied which is directed into the plane of the paper.)

The difference in potential, $V_{N}-V_{M}$, that is produced between two points such as $M$ and $N$ when a sample carrying current is placed in a megnetic field directed at right angles to the current, is called the Hall voltage $V_{H}$. On the basis of the explanation of the Hall effect given above, would you expect the Hall voltage measured for a given sample to increase, decrease, or remain the same if one increased the magnitude of the magnetic field, keeping everything else constant?

Experimentally, it is found that for a given sample

$$
\begin{equation*}
V_{H}=R_{H} \frac{B i}{t} \tag{1}
\end{equation*}
$$

where $B$ is the magnitude of the magnetic field, is is current in the sample, $t$ is the thickness of the sample (see Fig. 5), and $R_{H}$ is a constant, character-
 istic of the sample. $R_{H}$ is called the Hall coefficient of the sample and is considered to be positive if the charge carriers are positive and negative if the charge carriers are negative. As is evident from equation (1), $R_{H}$ can be determined if one measures $V_{H}, i, B$ and $t$. In this experiment we are going to measure $V_{H}$, $i$ and $B$. The thickness $t$ has already been measured carefully and is marked on the sampie holder. Measurements of $B$ and $i$ are reasonably straightforward. The measurement of $V_{H}$ is somewhar complicated due to the fact that experimentally points $M$ and $N$ do not always lie exactly on the same cross-section. Consider for example Fig. 6 which shows points $M$ and $N$ misaligned. With the (conventional) current flowing as indicated by the arrow, point $N$ will be higher in potential than $N$
 even without a magnetic field. If is ent always flows in a conductor from a higher to a lower potential If the carriers are positive and flowing as indicated, then the application of a magnetic field directed into the paper would force the positive carriers downward, and make $N$ even higher in potential than before. It is this increase in the potential difference which
is the true Hall voltage. The true Hall voltage $V_{H}$ is most easily determined by making two measurements of the difference in potential between $M$ and $N$, one with the field directed into the paper and one with the field directed out of the paper. The true Hall voltage then is

$$
\begin{equation*}
V_{H}=\left[\left(V_{N}-V_{M}\right)-\left(V_{N}-V_{M}\right)^{1}\right] / 0 \tag{2}
\end{equation*}
$$

where $V_{N}-V_{M}$ is the first of these two measurements, and $\left(V_{N}-V_{M}^{N}\right)$ is the second of these two measurements. Note that these two quantities may have different signs, depending on the extent of the misalignment of $M$ and $N$.

## Procedure

1. Since you will use a potentiometer to measure $V_{N}-V_{M}$ as well as to determine the current flowing in the circuit, it will be necessary for you to refresh your memory as to its operation and method of calibration. (Reread, if necessary, the experim mental writeup dealing with the use of a potentiometer.) Since the voltages you will measure in this experiment will not exceed 2 volts, it will be helpful to calibrate your potentiometer so that a dial reading of 1000 division corresponds to a potential of 2.000 volts. Make this calibration before proceeding.
2. Your sample is encased in a piece of plastic to protect the leads. Note that points $M$ and $N$ are brought out to green terminals and that the yellow terminals are connected to the ends of the sample.
3. Wire up the circuit as indicated in Fig. 7.



Do not plug in your 6 volt power supply until the instructor oK's your circuit. Set the sliding contact on the rheostat about midway between the two ends. Note that with the switch $S$ thrown to the right the potentiometer is connected across the two ends of the 10 ohm. precision resistor, while, when it is thrown to the left, it is connected to points $M$ and $N$.
4. After the instructor OK's your circuit and you have calibrated your potentiometer, plug in the power supply and throw $S$ to the left. Orient the board holding the sample and permanent magnet
so that the north pole (marked $N$ ) of the magnet is nearest you. The magnetic field $B$ between the poles will now be directed away from you as indicated in Fig. 8. Orient the sample so that it can be

slipped into the field. With the sample in this position but still out of the field, carefully trace the circuit starting from the positive (red) terminal of the 6 volt power supply to determine if the conventional current is flowing from right to left through the sample as you view it. If necessary, reverse the connections to the two yellow terminals so that the current does flow from right to left. Carefully trace the lead from the top end of the sample (call this point M) to see if it goes to the potentiometer terminal marked + or that marked -- Tie a string or rubber band around this lead to remind you it is the lead that goes to M. With the sample still out of the field see if you can detect with your potentiometer any difference of potential between points $M$ and $N$ : Since you do not know in advance whether $M$ or $N$ is at the higher potential, and since it is impossible to obtain a balance on the potentiometer unless the point of higher potential is connected to the + terminal of the potentiometer, it may be necessary to reverse the leads going to the potentiometer, in order to get a reading. If you are able to detect a difference of potential, record this and note whether it is M or N which is at the higher potential. If you are unable to detect any difference of potential between $M$ and $N$, you can conclude that the misalignment is very small, and the information you obtain in step $\underline{5}$ will suffice to determine the sign of the charge carriers.
5. Insert the sample into the field. You should now be able to measure a difference of potential between $M$ and $N$ although again it may be necessary to reverse the leads at the potentiometer to obtain a reading. Once you are able to get a balance, record the reading and again note whether $M$ or $N$ is at the higher potential.
6. From the information you obtained in steps 3 and 4 and the explanation given in the introduction you should be able to decide whether the charge carriers are positive or negative.
7. Throw switch $S$ to the right and determine with the potentiometer the voltage across the 10 ohm resistance (again, it may be necessary to reverse the leads going to the potentiometer to get a balance.) If this voltage is less than 0.5 volts. move the sliding contact on the rheostat a small amount in. the direction which decreases the resistance, and recheck the voltage across the 10 ohns. Continue this process until the voltage across the 10 ohm resistor is somewhere between 0.5 and 1.0 volt* Measure and record this voltage. Throw the swtich $S$ to the left, insert the sample in the field and carefully measure and record $V_{N}-V_{M}$. Now remove the sample from the field, turn it over and reinsert it. Carefully measure and record $\left(V_{N}-V_{M}\right)^{1}$.
8. Move the slider on the rheostat about 1 to $1 \frac{1}{2 \prime \prime}$ in the direction which increases the resistance in the circuit. Leaving it in this position, measure the new voltage across the 10 ohm resistor, and the new values of $\left(V_{N}-V_{M}\right)$ and $\left(V_{N}-V_{M}\right)^{1}$ as indicated in step 6 .
9. Repeat step 7 three more times. It will be helpful perhaps to make a table as follows:

10. Measure the magnetic field $B$ between the poles of the magnet, using the Gauss meter. The instructor will show you how to use this instrument.
11. For each of the trials calculate $V_{H}$ from equation (2) and $R_{H}$ from equation (1). Calculate the average* and the average deviation. (If you prefer, calculate the standard deviation.) The average or standard deviation can be used as a measure of the precision of your measurement of $R_{H}$.

* Or until slider is at the end of its range.

12. In this experiment, $B$ and $t$ were held constant. Equation (1) predicts that $V_{H}$ should be proportional to i. You might like to check this by plotting $V_{H}$ against $i$.
13. Can you explain why equation (2) gives the true Hall voltage?
*A "better" average could be obtained by "weighting" each $R_{H}$ by an amount proportional to the current in the sample. can you see why this is so?

## OBJECT:

To study the conditions which give rise to longitudinal standing waves in an air column.

## THEORY:

Longitudinal waves traveling along a tube are reflected at the ends in the same way as are transverse waves in a string. In this case also, the waves traveling to the right and those traveling to the left may combine to form standing waves having a large amplitude at certain natural (resonant) frequencies. As in the case of the vibrating string, a node should exist at a closed end since the molecules aren't free to vibrate longitudinally. Tf the tube is narrow compared with the wavelength an antinode will occur at an open end. Thus by drawing the possible standing wave patterns, it is seen that the wavelengths of the first, second, and third harmonics are as shown below for a tube closed at one end and open at the other.




The corresponding resonant frequencies (f $=c / \lambda$ ) are therefore

$$
\begin{equation*}
f_{n}=\frac{(2 n-1) c}{4 L} \quad n=1,2,3, \ldots \tag{1}
\end{equation*}
$$

For a tube open at both ends resonant frequencies can be predicted by assuming antinodes at both ends.

Suppose we have a closed tube which may be varied in length by positioning a plug in the tube. If we hold the frequency fixed then according to equation (l) we should have resonance whenever

$$
\begin{equation*}
L=\frac{(2 n-1) c}{4 f} \tag{2}
\end{equation*}
$$

is satisfied for any integral value of $n$. The smallest
value of $L$ is $c / 4 f$, the next larger value is $3 c / 4 f$, the next $5 \mathrm{c} / 4 \mathrm{f}$, etc. Thus if the L values wich produce resonance are plotted against the numbers $1,3,4,7$, etc., the resulting points should lie on a straight line of slope c/4f.

## APPARATUS:



## INSTRUCTIONS:

(1) Record an identifying number for the oscillator used. Arrange the apparatus as shown in the preceding figure. Set the oscillator dial at 1000 cycle/sec. Starting with the plug near the end of the tube farthest from the speaker, move the plug slowly toward the speaker. Note those positions where the sound intensity increases to a maximum. (Note: Just beyond the resonant points the intensity drops rapidly. These sudden changes may be more readily detected than the actual maximums.) Measure carefully the distance these maximum positions are from the speaker end of the tube. Since the location of resonant positions is somewhat a matter of judgment, locate them independently a number of times.
(2) Repeat step (1) with the oscillator set at 1500 , 2000 , and 2500 cps . For one of the frequencies determine if the lengths $L$ measured depend upon the position of the speaker relative to the tube.
(3) Determine the exact frequency of the oscillator at each of the above settings by hooking it up to the E-put meter via the auxiliary terminals on the supply panel. The E-put meter is being used by several other groups, so it will be necessary to coordinate your activities with theirs.
(4) Using the data obtained in the previous steps plot, for each of the four frequencies, a graph of the length of the tube to the resonant positions against the numbers $1,3,5,7,9,11 . \quad$ Plot the distance to the resonant position nearest the speaker above the $l$ on the horizontal scale, the distance to the next resonant point above the 3 , etc. According to equation (2) the points plotted in this manner should lie in a straight line. Draw the most representative lines through your plotted points and determine the slopes and intercepts for each frequency.
(5) From the slopes determined in the previous step and the frequencies found in step (3), determine the velocity of sound in the tube (average the four values) 。
(6) According to equation (2) the graphs plotted in part (4) should go through the origin. Does your experimental data fit equation (2) and, if not, how would you modify equation (2) to make it fit your data? Can you think of any physical reason why equation (2) might need to be modified?

OBJECT: To measure as a function of wavelength the index of refraction of a glass ample.
Ref - H \& R Chap 41 (see example 3)(see prob. 7)
THEORY: The glass sample, in the form of a prism, is illuminated by a beam of light from a source $S$ as shown in Fig. 1. Light is admitted to the collimating tube by a slit and is converted into a parallel beam by a lens at the other end of the tube. This beam then strikes the prism, is refracted at the two surfaces, and enters a telescope which is used to observe the light after refraction.


Fig. 1
The light beam strikes the first surface of the prism at angle of incidence $\theta$ and is refracted twice, leaving the second surface at angle $\theta$ ? with the normal. The angle of deviation $D$ is the angle which the rays emerging from the prism make with the rays which are incident on the first surface. If the prism is rotated so as to vary the angle of incidence $\theta$, the deviation angle D will change and one must move the telescope in order to view the beam. There will be one particular angle of incidence for which the deviation $D$ has a minimum value $D_{m}$ : Using the laws of refraction and geometry, one can show that the index of refraction of the glass is given by
$n=\frac{\sin A+D_{m}}{2}$

The above discussion assumes that the source $S$ emits light waves of only one frequency (i.e. a monochromatic source), the index of refraction calculated from equation (1) being the index of the glass for this frequency. If the source is not monochromatic but contains a number of discrete frequencies, each frequency will be deviated through a different angle $D$ and several beams will emerge from the prism - one for each frequency of light emitted by the source. This occurs because the refraction of light at the two surfaces depends on the index of refraction $n$ of the material (Snell's law) and the index is different for light of different frequencies. One can find the value of $n$ for any of the frequencies emitted by the source, simply by determining the minimum angle of deviation $D_{m}$ for that particular frequency and using equation (1). A plot of the index of refraction of a substance as a function of light frequency $f$ or wavelength (in air) $\lambda$ is called a dispersion curve.

APPARATUS: The apparatus used for this experiment is called a spectrometer. It consists essentially of a collimator which is simply a tube with a slit at one end and a lens at the other, a telescope and a prism table. The telescope and prism table are arranged so they may be rotated independently about a vertical axis. A circular scale permits one to measure the angle through which the telescope is rotated.

To make accurate measurements with a spectometer the following preliminary adjustments must be made:

1. The collimator and telescope must be adjusted so that the beam emerging from the collimator is parallel and so that the objective lens of the telescope brings this light to a focus exactly in the plane of the cross hairs.
2. The axes of the collimator and telescope must be adjusted so they are perpendicular to the axis of rotation.
3. The refracting surfaces of the prism must be made parallel to the axis of rotation.

These adjustments have already been made.

[^1][^2]A large thumb screw directly beneath the collimator and about one inch from the base of the instrument clamps the prism table and scale. A similar thumb screw directly beneath the telescope clamps the telescope. When this is tifhtened, the telescope can still be rotated a small amount by the vernier screw located just above and to the right of the telescope clamping screw. A small thumb screw at the slit end of the collimator adjusts the width of the slit. The four thumb screws mentioned in this description are the only ones that should be manipulated by the student.

## Gaertner Spectrometer

A large thumb screw near the axis of the instrument and about one inch above the divided scale clamps the telescope. A second large thumb screw at the same height but at the base of the telescope support arm permits a fine adjustment of the telescope position. The prism table assembly can be clamped by means of the small thumb screw located near the axis of the instrument and about $1-1 / 2$ inches below the top of the prism table. Note that the prism table assembly rides on a collar which is clamped in position by means of a second small thumb screw. When it is necessary to raise or lower the prism table assembly, both small thumb screws should be loosened and both the collar and prism table assembly moved at the same time. At the slit end of the collimator is a small thumb screw which adjusts the width of the slit. This thumb screw along with the other ones that are mentioned in this description are the only ones that should be manipulated by the student.

INSTRUCTIONS:

1. Look into the telescope and move the small tube containing the eye lens in or out until the cross hairs may be seen distinctly.
2. Loosen the screw which clamps the prism table and rotate the prism table until the apex of the prism points toward the collimator. Re-tighten the screw to keep the prism in this position. Illuminate the slit using the small 110 volt lamp provided. Rotate the telescope into such a position that the image of the slit formed by light reflected from one face of the prism can be seen in the telescope (position B, Fig. 2). If you can't find the image, ask the instructor for assistance---do not move the prism around or adjust the leveling screws. Clamp the telescope in this position and use the fine adjustment screw so as to set the cross hairs exactly on the image of the slit. Read and record this position of the telescope on the circular scale. Loosen the telescope clamping screw and rotate the telescope into $\mathrm{B}^{\prime}$. Reclamp and use the fine adjustment screw to set the cross hairs on this image. Read and record this position of the telescope on the circular scale.


The angle through which the telescope has been rotated from position $B$ to position $B^{\prime}$ is twice the refracting angle $A$ of the prism. Prove this! Since the decision as to when the cross hairs coincide exactly with image is a matter of judgment, it would be wise to determine the readings corresponding to positions $B$ and $B^{\prime}$ several times independently. The average deviation of these individual settings will give you some idea of how precisely you can measure angle A.
3. Loosen the screw which clamps the prism table and rotate the table until the prism is approximately in the position shown in Fig. 1. Mount the Hg vapor lamp so that it will illuminate the slit of the collimator. With the telescope approximately in the position shown in Fig. 1, look through the telescope and adjust the position of the telescope and/or prism table until you see in the telescope a number of colored images of the slit. These will be spaced approximately as shown in Fig. 3 (all of the lines may not be visible for any one position of the telescope, and it may not be possible for some students to see all of the violet lines).
4. Focus your attention on the list line of the spectrum. Keeping this line in view in the telescope, slowly rotate the prism table in the direction which decreases the angle of deviation for this line. Note that there is some position of the prism table for which the angle of deviation is minimum and if the prism table is turned in either direction from this position, the deviation will increase. (With the prism table set in the position for minimum deviation for the line in question, clamp the prism table and carefully adjust the position of the telescope until the cross hairs are centered on the line.) Read the position of the telescope on the circular scale. From here on the procedure will vary depending on which instrument you are using.

FOR THOSE USING EALING INSTRUMENT
5. Leaving the prism table clamped, rotate the telescope until it is directly opposite the collimator. The prism can be moved aside without rotating the table so that you should be able to see a direct image of the slit. Set the telescope carefully so that the cross hairs fall on the image and record the position. The difference be. tween this reading and the first one is the minimum angle of deviation for the line in question.
6. Unclamp the prism table and repeat steps 4, 5 and 6 for each of the other lines in the spectrum of the source. Finally, set up the small white light at the slit and note the spectrum due to this source.

FOR THOSE USING THE GAERTNER INSTRUMENT
5. Unclamp the prism table and rotate the table so that it is in the position of minimum deviation for the second line of the spectrum. Clamp the table and carefully adjust the position of the telescope so that the second line falls on the cross hairs. Record this position of the telescope. Repeat for all the lines in the spectrum. Set up the small white light source at the slit. Note the spectrum produced by this type of source.
6. Lower the prism table assembly after first loosening the two small set screws. Swing the telescope to a position directly opposite the collimator. It should be possible to see a direct image of the slit. Set the telescope so that this image is centered on the cross hairs and record the reading on the circular scale. Raise the prism table assembly to its original height. You can check this by holding the small white light at the eye piece end of the telescope and examining the beam emerging from the front end to determine if it is completely intercepted by the prism face.

## ANALYSIS AND RESULTS:

1. Calculate for each wavelength the index of refraction using equation (1).
2. Draw a dispersion curve showing the index of refraction $n$ in this type of glass for various light wavelengths (in air) $\lambda$ 。
3. For a wavelength of 6000 A , the indices of refraction of dense flint, light flint and crown glass are respectively 1.65, 1.58 and 1.52. Based on your data, of which type of glass is the prism made?


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If a transparent prism is illuminated by a beam of light from a source $S$ as shown in the figure below, the beam strikes the first surface of the prism at angle of incidence $\theta$ and is refracted twice, leaving the second surface at angle $\phi$ with the normal.


The angle of deviation $D$ of the light depends on the angle of incidence $\theta$, and on the frequency of the radiation since the index of refraction is different for different frequencies in "dispersive" media such as that from which the prism is made. It can be shown geometrically (do this) that:

$$
\begin{align*}
& D=\theta+\phi-A \text { or } \phi=D+A-\theta  \tag{I}\\
& A=B=R+I \tag{2}
\end{align*}
$$

Applying Snell's law at each refracting surface:

$$
\begin{aligned}
& R=\arcsin [(\sin \theta) / n] \\
& I=\arcsin [(\sin \phi) / n]
\end{aligned}
$$

Substituting into (2) we get

$$
A=\arcsin [(\sin \theta) / n]+\arcsin [(\sin \phi) / n]
$$

or, bringing all three terms to one side of the equals sign and calling the sum of these three terms $F(n)$ :

$$
F(n)=\arcsin [(\sin \theta) / n]+\operatorname{arsin}[(\sin \phi) / n]-A=0
$$

where equation (l) allows the replacement of angle $\phi$ by $D+$ $A-\theta$. Thus if angles $\theta$ and $A$ are held constant, equation (3) expresses implicity how the deviation angle $D$ depends on the index of refraction $n$ and vice versa. The prism will i have different indices of refraction $n$ for the different frequencies, which will then be deviated by different amounts $D$ for light incident at a fixed angle $\theta$ on a prism of a certain apex angle A.

An equation like (3) which is not solved explicitly for one variable in terms of another (e.g. $D=E(n)$ ) but has.... the variables "mixed up", is called a "transcendental". equation. Solutions to this type of equation may be obtained. by a method of successive approximations called the Newton-

Raphson method (see McCracken, "Fortran with Engineering Applications", Chapter l0). Suppose that you wanted to solve an equation $F(x)=0$ such as that shown in the figure below for the values of $x$ which make $F(x)=0$ (the roots of the equation).


To start the computation a guess $x_{0}$ is made as a rough approximation to the value of the root. The value of $F(x)$ and the slope of the $F(x)$ curve at $x_{0}$ is obtained and a better approximation to the root is Calculated from the equation for the slope

$$
F^{\prime}\left(x_{0}\right)=F\left(x_{0}\right) /\left(x_{0}-x_{1}\right)
$$

or

$$
x_{1}=x_{0}-F\left(x_{0}\right) / F^{\prime}\left(x_{0}\right)
$$

where $x$ is the second approximation to the root. Then, in the same way $F\left(x_{1}\right)$ and $F^{\prime}\left(x_{1}\right)$ are calculated and a better approximation

$$
x_{2}=x_{1}-F\left(x_{1}\right) / F^{\prime}\left(x_{1}\right)
$$

and so forth until there is a negligible increase in precision for any further iterations. In order to use this procedure on equation (3) it is necessary to have an expression for the slope of the $F(n)$ curve, obtained by taking the derivative of equation (3) with respect to $n$ while holding other variables constant.

$$
\begin{equation*}
F^{\prime}(n)=\frac{-(\sin \theta) / n^{2}}{\left[1-\left(\sin ^{2} \theta / n^{2}\right]^{1 / 2}\right.}+\frac{-(\sin \phi) / n^{2}}{\left[1-\left(\sin ^{2} \phi\right) / n^{2}\right]^{1 / 2}} \tag{4}
\end{equation*}
$$

Experiment:
I. Using the clamping serew which locks the prism table; fix the prism in a position such that its apex points toward. the collimator and white light source as shown. (It isn't possible or necessary to align the base of the prism exactly perdendicular to the beam).


Use the fine adjustment screw to set the telescope cross hair s exactly on the reflected image of the slit at $B$ and $B^{\prime}$, recording both positions on the telescope circular scale. Repeat this measurement several times and determine the average values and standard deviations. Prove that the angle through which the telescope is rotated from $B^{5}$ to $B^{\prime}$ is twice the prism apex angle $A$ and calculate $A$ and its standard deviation.
II. Using the mercury vapor lamp source rotate the prism table until it is in the position shown in Fig. l. Adjust the beam slit opening and the position of the telescope and prism table until you see a number of sharp narrow vertical lines of different colors through the telescope. These are images of the slit opening formed by light of the wavelengths given below.

| Yellow | $5791 \AA$ and $5770 \AA \AA^{\circ}$ |
| :--- | :--- |
| Green | $5461 \AA$ |
| Bluegreen | $4916 \AA$ |
| Blue | $4358 \AA$ |
| Violet | $4078 \AA$ and $4047 \AA$ |

By trial and error position the prism so that the green line has the smallest deviation angle D (see Fig. l) that it can have for any prism setting. This "minimum deviation" angle for the green light is the one for which $\theta$ and $\phi$ are equal for light of this wavelength, and these angles can thus be easily determined from the position of the telescope when lined up on the refracted beam, using equation (1). Leave the prism table clamped in the same position ( $\theta$ will be the same for all lines in the refracted spectrum) and record the position of each of the visible lines on the telescope scale. Each reading should be repeated independently at least three times. In order to get a line in the red part of the spectrum change your source to a gas discharge hydrogen or helium tube and determine the position of one of the following lines without changing the prism orientation. (Note: BEWARE OF THE HICH VOLTAGE - SEVERAL THOUSAND VOLTS - USED DIN THE GAS DISCHARGE TUBES - THINK!!

Again keeping the prism in the same orientation as before determine the positions of some lines of unknown wavelength (for example other lines from hydrogen, heliun, cadmium, oxygen, nitrogen or neon). Finally, rotate the telescope until it is directly opposite the collimator and move aside the prism without unclamping the table (Ealing instrument) or lower the prism table assembly after first loosening the two small set screws (Goertner instrument). Record the position of the unrefracted light beam on the telescope scale. Ihis reading will be necessary in order to get the deviation $L$ associated with each of the previous readings. Calculate angle $\theta$ and angle $D$ for each of the lines on which you made measurements.
III. There will be a computer program available in the laboratory for solving by the Newton-Raphson method for the index of refraction that the prism has for light of each of the wavelengths present in the spectral lines. The program repeats the successive approximations to the value of each n until the difference between a computed value and the preceding value is less than .00001.
IV. Plot a graph showing the prism index of refraction as a function of the wavelength of the refracted light. Use the graph to deternine the wavelengths of the Iines for which $\lambda$ is unknown. A more precise way of determining these wavelengths would be to determine the coefficients in a leastsquares fit of a polynomial $y=A+B x+C x^{2}+D x^{3}+\ldots . .$. to the graph of index $n$ versus wavelength $\lambda$ or vice versa. Any wavelength could then be determined by putting the corresponding value for index of refraction $n$ into the equation. The computing center has subroutines for doing such a least-squares fit (see "Scientific Subroutines" section of their manual for subroutines "GDATA, ORDER, MINV, and MULTR")

## INTERFERENCE AND DIFFRACTION

According to Huygens' principle, each point along a wave-. front may be regarded as a new source of waves. Whenever something obstructs part of the wavefronts, interference between "wavelets" emanating from different parts of the unobstructed wavefronts produce a diffraction pattern which is characteristic of the geometry of the obstruction (or opening in object which blocks the light) and of the wavelength of the light. It is shown in nearly all introductory physics textbooks, for example (see Resnick and Halliday, section $43-1$ ), that when light waves pass through a double slit arrangement like that shown below they interfere constructively and destructively at different positions to form fringes on the screen $S$ such that intensity maximum appear at positions


In somewhat the same way wavelets passing through different parts of a single slit interfere to produce a single slit diffraction pattern with destructive interference causing diffraction minima at angles $\theta$ such that (see Resnick and Halliday, section 44-2)

$$
a \sin \theta=m \lambda \quad m=1,2,3, \ldots \ldots \ldots
$$

with maxima approximately half way between (the exact intensity expressions are given in section 44-3), where a is the slit width. A circular aperture of diameter d results in fringes having circular symmetry with the first minimum appearing at a distance from the center such that (see Resnick and Halliday, section 44-5)

$$
\sin \theta=1.22 \lambda / d
$$

Experiment: (Due to a limitation on equipment, half the groups will have to start on part (2) below and do (1) at the end of the experiment.)

The lasers used for a light source are helium-neon gas discharge lasers and put out a beam of wavelength $\lambda=6328 \AA$. The apparatus is set up as shown below.


The converging lens is moved close enough to diverging lens to keep the beam diameter constant from the converging lens to the wall (beam must be wide enough to completely fill the diffraction opening.
(1) You will be given an IBll card. Use the comparator to determine the dimensions of a punch hole and the spacing (center to center) of adjustment holes. Measure the distance from your diffraction openings to the wall.
(2) Put a single punch opening in the beam as the diffraction opening and measure the distances to the various diffraction peaks in the two directions. Repeat with the triple punched opening.
(3) Repeat part (2) with a single narrow slit, a circular opening, and a screen.
(4) Compare theory and experiment using your measurements of parts (1) and (2) which involve the single and triple punch holes in the IBM cards. What is the percentage difference between the length and width of a single rectangle calculated from measurements on its diffraction pattern and the length and width measured directly?
(5) Discuss qualitatively the appearance of the diffraction patterns of the circular opening and the screen of part (3) and determine the slit width, circular hole diameter, and screen mesh spacing from the diffraction pattern measurements. The bright disc at the center of the diffraction pattern for the circular hole is called the airy disk. You may want to compare this disk diameter (really the diameter of the dark ring surrounding the bright center) to the theoretical value given in the text (section 44-5 in Halliday $\mathcal{E}$ Resnick).

Ref. Halliday, Resnick, Sections 45-1,2,3.
In the prism spectroneter experiment, a spectrum of mercury was obtained by allowing light from a nercury lamp to be rerracted by a prism. In this experiment a spectrum of mercury will be produced by diffraction; a groting replaces the prism.

The arrangement is sketched in Figure 1 . If your grating is labelled $B$ you nay be able to observe only the zeroth and first order spectrum. If it is manked $A$, several orders may be visible. The angle $\theta$ at which a given wavelength $\lambda$ is observed is related to the wavelength as

$$
\sin \theta=m \lambda / d
$$

Where $d=$ grating spacing (distance between lines ruled in grating from which replica has been made)
$m=a n$ integer, the "order" of the observed diffraction maximun.

## Procedure

Use iirst order spectrum for $B$ grating, second order if you have A grating. Canerully position the telescope so that the intense violet Hg line is centered on the crossmair intera section and record the position of the tejescope, Repeat for each of the other lines, keeping a tabulam pecord. Then talse readings for the same lines, same order, on the other side of the zeromorder slit inage, Suggested fomat:

$$
\text { Color } \quad \underline{\text { Wavelength }} \quad \text { Reft } \quad \text { Right } \quad \theta=\frac{R_{I-} R_{R}}{2} \quad \text { sin } \theta
$$

marcury
violet 404.7 A Paint violet-blue 4358
bluemgreen 4916 faint
green 5461
yellow 5770
Jellow 2791
IMEON
yellow
Upon finishing the Hg observations, place the neon source (orange) in rront of the slit and take readings on the brightest yellow line in its spectrum.

Ask someone with a grating label ( $A, B$ ) diferent from yours for a look at his IIg spectrum. Does A or B have a smaller grating spacing? (If curious, analyse a few higher order lines.)

Analysis
Using a full sheet of graph paper and scale s which male full. use of the precision of your measurements, plot sin $\theta$ vs. wavelength. Since the girting equation above predicts a linear relationship between sin $\theta$ and $\lambda$, use a ruler to draw in the line which best fits your data Extrapolate on your plot to find the wavelength of the neon yellow line. From the graph, find $d$. Include quantitative estimates of the experimental moertainties in your reported results.

## Optional questions

1. Can a "line which best fits" be found by more sophisticated techniques? Are these warranted here?
2. Two yellow lines in the $H$ g spectrum, separated by about 20 Angstrom whits, are well resolved. In first order (m = 1), hots close can two red Lines be in order that they be resolved. by your grating? (Base your estimate on resolving power $R$ (text) considm ering the Rayleigh criterion, in e, diffraction expects at the grating this ignores other erects, such as diffraction at the entrance slit and lens effects.
3. Look up your prism report and comment on the relative resolering power of the two " dispersing" devices, prism and grating.


To observe the efect of interserence of light neve nad to meanure the wavelength of the light emitted by a monochromme source using Young' experiment.

THEORY:
Tn Young " "double slic" experiment an approsimataly monochromatic light source illuminates a slit, and light from this alit is allowed to fall on two nlits nerarated by - spacing os ahown in the figure below. Is these two blite ese suificiantly narrow, each whu act as goixt source Eending out new wavelut (Huygens' frinciche) townd the sexeen at the right. The illumination infensity atamy point P can be lound by akeximposing the wave arriving at the foint from the bwo lins; however this "adding" of the waves yields diferent restite at difierent foints $P$. For example, consider light wave which reach the point P ohown in the iguxe rom the Ewo lite. The two wave
 greater distance $82=r$ it lage behrat the other at $P$. II the lag amount to one. fwn, three etc. wavelearthe nt point p. adding the two wawes given rise ho wave having an amylitude which is the sum of the amplitudes of the indiwidual weve. This condithon is setersed so as constructive interference and resultm in a high intembity at P. If point $P$ is a point where the fath difference $82-x$ ig much an to make one wave arciving at the point lag the other by an odd number of half-wvelengths. the wave formed by sererposition has almost sero amplitude. Thi condition is referred so as destructive interference and results in very low intenaity at P.


1t It Bhown in number of texts (e.g. Sears and Zemansky that foy points for which the angle 0 名a small, the path aifference $\left(x_{2}-x_{1}\right)$ is given the equation

$$
x_{2}-x_{1}=\frac{d}{D} x
$$

and hence constructive intexterence occurs at poluts $P$ for which
(5) $x-1$ where nw $0,1,2,3 \ldots \ldots$

Tt follows that there wll be bsight intercererem syingen ot the following values of x:

$$
\begin{aligned}
& x_{n}=\frac{x_{2}-\frac{2 D}{d}}{} \\
& m_{1}-\frac{D A}{d} \\
& m_{3}-\frac{3 D \lambda}{d}
\end{aligned}
$$

Whenever o part of the waverronts comine from - point source axe cut of by some obstacle ixinges occur as the result of intexrexence between buygens" wavelets manatine Irom vasious parts of the unobe structed wavex̌roncs. For example a small circular object when placed in the light rrom a point source produces circular resnges. alternatoly bright and dark. The process involved 18 present in every wave but this edect $1 s$ observable only if some part $11^{\circ}$ the wave 1 s cut ors and $1 s$ called difruction.

AEPAMATUS:
A long box with silt at one end and the cilm holder of Polarold Land camera at the other. a monochromatic (approximately) light souxce, double Elits, alne wire, and a razor blade.

Instructions:
(Important Note: Please keep the cameras wrapped up and inside the boxes when not in use. Do not pull open metal shutter until you are ready to expose the illm. In each of the rollowing parts remove the camere and observe the interference pattern through the magnifler lens berore photo graphing the pattern. Rotate the primary slit (directiy in front of source) to that position where the reinges are most distinct (parallel to the doubie slits or the diffracting obstacle). ther replace the camexa and make an exposure. An exposure 1 s staxted by pulling the thin sheet metal diaphragm up to the ine inscxibed on 1t...... not all the way out or 1ts slot. To stop the
esposure push this diaphragm back down. Hote that there axe tharee notches cut in the flat plate to which the film holder 18 dastened. The bottom of the plate and these three notches detexmine rour fixed vertical positions or the camere. Four exposures are to be taken on a single 811 m . Each proup is expected to tate only one gleture but will be allowed one more Is mastake is made on the (1rst.)
(2) Each group will bo givers two double slit vystms with difcerent spacing d between the silta. place the double :llt with the maller spacing In the support near the center of the box. Astex adjusting for the most aistinct reinges degcribed above, make mhee minute and an - lint minute exposure on the same picture (olecerent vertical position of camera).
(2) Using the Bame P11m repent part (1) with the other double silt. Ask the instructor about the procedure for developing and removing the ciln. Actex the film bas been removed rxom the camere coat it mith the applicator provided.
(3) Mount the tilm on a comparatox (ix they are both in use 80 on to steps 4 and 5). Observe the pattern by means of the microscope provided and deternine the distance between Rainges for the two disferent double slita. (On the lasger comparator one sull curn of the matn dial prom duces a carriage displacement of 1 mm .) since the fringes are presumable equally spaced. one can obtain precision in determining the distance between two successive fringes by measuring the distance between two widely separated iringes and dividing by the appropriate number. Make some kind of estimate of the precision or thes determination of the distance between successlve reinges.
(4) fteasure the distance from the silts to the metal diaphragm on the front of the fllm holder. The distance from this diaphragm to the Illm postion is 2.70 cm . Use the compavator to meature the spacing between the siats.
(5) Heplace the double slits with a sine wioe. observe the difraction pattexn with magnifiex. Hate this as sharg as possible by rotating the primary slit a rew degrees to the right or lett. Sketch the pattern in your notebook. Kepert ror the stralght edge (razor blade).

## MHALYSTS GUGSTZONS:

1. Colculate the wavesength of the 1 ght used ron your Hesumphents of $D_{2}$ the distance between successive fringe, and the values or sils spacing $d_{\text {. }}$
2. What accuracy would you ciasm for the wave lergerg Justisy your answex.

The arrangement of equipment is illustrated in Fig. 1.


Fig. 1
$S_{I}$ and $S_{I I}$ are surfaces which are forced to move in simple harmonic motion at a frequency $f$ determined by the oscillator setting.
With the circuit shown, the motions of $S_{I}$ and $S_{I I}$ are exactly $180^{\circ}$ out of phase, so they may be represented mathematically by

$$
\begin{aligned}
& Y_{S I}=A \sin \omega t \\
& Y_{S I I}=-B \sin \omega t
\end{aligned}
$$

Both surfaces, $S_{I}$ and $S_{T I}$ radiate sound waves of frequency $f$. Referring to Eig. 2, the waves sent out in a given direction, say $x$, by source $S_{I}$ may be represented by the equation

$$
Y_{I}=A \sin \omega\left(t-\frac{x}{c}\right)=A \sin (\omega t-k x)
$$

and those sent out by $S_{I I}$ in a direction specified by $X$ as

$$
Y_{I I}=-B \sin \omega\left(t-\frac{X}{C}\right)=-B \sin (\omega t-k X)
$$

Here $k=\frac{\omega}{c}$ and $c$ is the velocity of sound in the medium (air).


Fig. 2

The disturbance (motion) produced by the wave from source I at any point such as $P_{\text {, }}$ a distance $x_{1}$ from $S_{I}$ is simply

$$
\frac{y}{I_{p}}=A \sin \left(\omega t-k x_{1}\right)
$$

Similarly, the disturbance (motion) at $P$ due to the source $S_{I I}$ is given by

$$
\frac{y}{I}_{P}=-B \sin \left(\omega t-k X_{1}\right)
$$

It is generally assumed that at any point such as $P$ where the two wave trains cross, that the disturbance is simply the sum of the motions that each wave would produce separately, ie.,

$$
y_{P}=A \sin \left(\omega t-k x_{1}\right)-B \sin \left(\omega t-k_{X_{1}}\right)
$$

(A superposition principle applies here.) We consider two special cases.. Suppose $P$ is situated so that the difference between $X_{1}$ and $x_{1}$ is some whole number of wave lengths of the sound wave, ${ }^{1} . e .$,

$$
x_{1}-x_{1}=n \lambda=n \frac{c}{f}=n \frac{2 \pi c}{\omega}=n \frac{2 \pi}{k}
$$

where $n=0,1,2,3, \ldots$ If this holds then

$$
\begin{equation*}
x_{1}=x_{1}+n \lambda \tag{1}
\end{equation*}
$$

and for this special case

$$
\begin{aligned}
y_{P} & =A \sin \left(\omega t-k x_{1}\right)-B \sin \left[\omega t-k\left(x_{1}+n \frac{2 \pi}{k}\right)\right] \\
& =A \sin \left(\omega t-k x_{1}\right)-B \sin \left(\omega t-k x_{1}-n(2 \pi)\right) \\
& =A \sin \left(\omega t-k x_{1}\right)-B \sin \left(\omega t-k x_{1}\right) \\
& =(A-B) \sin \left(\omega t-k x_{1}\right)
\end{aligned}
$$

This last equation represents a simple harmonic motion of amplitude ( $A$ - B). If one considers the special case where $P$ is so situated that

$$
\begin{equation*}
x_{1}-x=(2 n-1) \frac{\lambda}{2}=(2 n-1) \frac{\pi}{k} \tag{2}
\end{equation*}
$$

$$
\mathrm{n}=1,2,3
$$

then it is easily shown that the motion at $P$ will be given by

$$
y_{P}=(A+B) \sin \left(\omega t-k x_{1}\right)
$$

This equation represents a simple harmonic motion of amplitude $(A+B)$.

Suppose that the two sources are arranged as in Fig. 3, and a detector (which responds to the resultant (sum) of the disturbances produced by the two waves at any point) is moved along $0^{\prime \prime} 00^{\prime \prime}$.


There should be some points along this path where condition (1) is satisfied and some points for which condition (2) is satisfied. It should be apparent that the point 0 is one of these points for which condition (1) is satisfied. If the first point between 0 and $0^{\prime \prime}$ for which condition (1) is again satisfied has a coordinate $z$ it should be easy to show that

$$
\lambda=\sqrt{D^{2}+\left(z+\frac{d}{2}\right)^{2}}-\sqrt{D^{2}+\left(z-\frac{d}{2}\right)^{2}}
$$

Hence, by measuring $z, d$, and $D$, one could determine $\lambda$, and if one can measure the frequency $f$ one can determine $c$, the velocity of sound in air.

Hints:
(i) The transducers are resonant systems and work best at their resonant frequency which is around 40,000 hertz.
(ii) $A$ and $B$ in the above equations can be made equal by adjusting the slider or resistance R . This makes it possible to get very nearly zero amplitude at points (nodes) for which (1) is satisfied.
(iii) The distance $d$ in Fig. 3 can be changed. You may want to see what effect this has on $z$.
(iv) The transducers radiate most strongly in the forward direction.

Index of Refraction of a Gas

Object: To measure the index of refraction of air using a modified Michelson Interferometer.

Apparatus: The construction of the interferometer is shown in the figure. A single wave from the light source is partially transmitted and partially reflected at 0 ,

giving rise to two waves, one of which travels to mirror $M_{I}$ and back and the other to $M_{2}$ and back. Part of each of these waves reaches the eye and they are brought to a focus at the same spot on the retina. If the wave traveling path $\mathrm{OM}_{1} \mathrm{O}$ arrives at the retina in phase with the wave traveling path $\mathrm{OM}_{2} \mathrm{O}$, that spot will be bright. Another wave originating at a different place on the monochromatic source will travel different paths to the eye and will be focused at a different spot on the retina. The waves from this spot on the source may arrive at the retina in phase or out of phase depending on their paths (i.e. starting point). The result is that one sees a series of ałternately bright and dark circular fringes, the bright fringes caused by waves which arrive in phase and interfere constructively and the dark fringes by waves which are out of phase and interfere destructively.

Theory: The condition that waves (from a single point on the source) traveling paths $\mathrm{OM}_{1} \mathrm{O}$ and $\mathrm{OM}_{2} \mathrm{O}$ arrive at the retina in phase is that their paths differ by an integral number of wavelengths $\lambda$. However the wavelength of an electromagnetic wave of a given frequency $f$ depends on the medium through which it is traveling and can be expressed

$$
\lambda=\lambda_{0} / n
$$

where $\lambda_{0}$ is its wavelength ir vacuum and $n$ is the index of refraction of the medium for a wave of this frequency. Thus if one wanted to express the number of wavelengths in distance $O_{1}$ he would write

$$
N_{1}=\frac{\left(D_{1}-t-2 t_{\mathrm{E}}\right)}{\lambda_{0} / n_{a}}+\frac{t}{\lambda_{0} / n_{1}}+\frac{2 t_{g}}{\lambda_{0} / n_{g}}
$$

where $D$ is the actual distance from 0 to $M, n_{1}$ is the index of refraction fo the gas in the cell, $t_{\mathrm{f}}$ is the thickness of each glass window, $n$ is the index of reffaction of the glass windows, and $n_{a}$ is the ihdex of refraction of air. Similarly the number of wavelengths in distance $\mathrm{OM}_{2}$ is

$$
N_{2}=\frac{D_{2}}{\lambda_{0} / n_{2}}
$$

The difference in the number of wavelengths along paths $\mathrm{OM}_{1} \mathrm{O}$ and $\mathrm{OM}_{2} \mathrm{O}$ is therefore

$$
\begin{aligned}
2 N_{2}-2 N_{1} & =\frac{2 D_{2} n_{a}}{\lambda_{0}}-2\left[\frac{n_{a}\left(D_{1}-t-2 t_{g}\right)}{\lambda_{0}}+\frac{n_{1} t}{\lambda_{0}}+\frac{2 n_{g} t_{g}}{\lambda_{0}}\right] \\
& =\frac{1}{\lambda_{0}}\left[2 n_{a}\left(D_{2}-D_{1}+t+2 t_{g}\right)-2 n_{1} t-4 n_{g} t_{g}\right]
\end{aligned}
$$

Letting $K=2 n_{a}\left(D_{2}-D_{1}+t+t_{g}\right)-4 n_{g} t$
one obtains for the difference in the number of wavelengths along the two paths,

$$
2 N_{2}-2 N_{1}=\frac{1}{\lambda_{0}}\left(K-2 n_{1} t\right)
$$

It is this difference which must be an integral number $m=0,1$, 2, 3, ......... if the two waves are to arrive in phase at the retina. Thus the condition for constructive interference of the two waves is

$$
K-2 n_{1} t=m \lambda_{0}
$$

The index of refraction $n_{1}$ of any gas is a function of its density and hence varies with the pressure of the gas. Suppose
that for a given pressure the above condition is satisfied with $m=m_{1}$ (some integral number) for a particular spot on the retina so that

$$
\begin{equation*}
k-2 n_{1} t=m_{1} \lambda_{0} \tag{1}
\end{equation*}
$$

If $n_{l}$ is allowed to increase slightly by adding a small amount of gas, the two waves will no longer arrive exactly in phase and the spot will be reduced in intensity. When $n_{1}$ reaches a value $n_{1}{ }^{\prime}$ such that

$$
k-2 n_{1}^{\prime} t=\left(m_{1}-1 / 2\right) \lambda_{0}
$$

the spot will be dark, since this is the condition that the two waves will be exactly out of phase. When $n_{l}$ reaches a value $n_{1}$ such that

$$
\begin{equation*}
K-2 n_{1}{ }^{\prime \prime} t=\left(m_{1}-1\right) \lambda_{0} \tag{2}
\end{equation*}
$$

the spot will again be bright since ( $m_{1}-1$ ) is also an integral value of $m$. Subtracting (2) from (1) we obtain

$$
n_{1}^{\prime \prime}-n_{1}=\frac{\lambda_{0}}{2 t}
$$

It follows then that if we start with condition (l) and allow gas to leak in slowly, counting the number of times N the spot changes from bright to dark and back again, we can write for the change in the index of refraction of the gas during this process

$$
\Delta n=N \frac{\lambda_{0}}{2 t}
$$

Therefore if we start with the cell evacuated so that $n_{1}=1$, we can determine the index of refraction of any gas at ${ }^{1}$ any pressure just by counting the number of times the spot changes while the gas is allowed to leak into the cell until the proper pressure is obtained. The wavelength $\lambda_{0}$ of the light in vacuum and the length $t$ of the gas cell must be known.

Instructions

$$
t=(8.000 \pm 0.002) \mathrm{cm}
$$

1. Using a mercury lamp with a filter for the 5461 A green line adjust the mirrors so that bright circular fringes can be seen clearly.
2. All groups should connect their cells and the manometer to the vacuum pump (open stopcocks) and evacuate them at the same time. Note the pressure in the cells as indicated by the difference in level of the two mercury columns in the manometer. Turn off the valve to the manometer.
3. Close off your cell from the pum and oper the stop cock to the controlled leak, alwire, atr to leak sjowly into the cell as you count the new fringes appearing as bright spots in the center of the pattern. Contirue to count until the cell again contains air at atmospheric pressure. Try to estimate the total number of fringes to a fourth of one fringe. Repeat this procedure.
4. Open one cell only to the pump and evacuate this cell, again reading the manometer. Close off the pump, but not the manometer, from the cell, and agan allow air to leak slowly into the cell. When 20 new fringes have appeared in the center of the pattern shut off the leak and read the manometer.

Analysis

1. Calculate the index of refraction of air at atmospheric: pressure and at the pressure obtained in part 4.
2. Estimate the precision of this method.

## LENSES II



The diagram shows that putting the object near the focal point of a converging lens causes rays to come out of the lens nearly parallel to each other at angle A to the axis. One of these goes through the center of the eye lens undeflected and forms an image on the retina. If the magnifier were not present, the object rays would go through the eye lens undeflected at an angle B. SHOW that the angular magnification for this case is the distance of the object from the eye divided by the focal length of the lens (this is only true for distances in the range of a foot or so from the eye, where some rays from the lens at angle A would be able to reach the pupil of the eye). Angular magnification is the ratio $A / B$, but for small angles we can take this ratio to be (tan A/tan B) and this is what one should use in the proof above.


## EXPERIMENT

For each of the two shortest focal length lenses, put a $3 \times 5$ card at approximately the focal length of the lens, and look through the lens at the card. Adjust the lens-to-card distance until you see a sharp image of the lines on the card you may want to draw. one or two additional lines on the card). Measure the magnification
by looking with both eyes at both the card and the lens and determine how much larger the line spacing appears as viewed in the lens. Compare this value to that given by the theory for each lens used.

## THE TELESCOPE

In this part of the experiment, you will study the simplest types of telescopes, made out of two lenses. At the end of the telescope nearest the object to be viewed is placed a lens (called the objective lens) with a long focal length. The other lens is close to the eye of the observer and is called the eyepiece. The distance between the lenses is the sum of their focal lengths. If the eyepiece is a converging lens, we have an 'astronomical telescope', but if the eyepiece is a diverging lens we have a 'Galilean telescope' or 'opera glass'. Your task in this part of the experiment is to construct one astronomical telescope and one Galilean telescope. For each telescope, you should measure the magnification. To do this, you will have to figure a way to measure the apparent enlargement of the object. You may want to do this by taking the cover of a book and viewing it in the telescope until a particular line of print becomes just legible. The distance of the object from your eye can then be compared to the distance from your naked eye for the same line of print to be barely readable. Or you may want to draw equally spaced lines on the blackboard and view the lines simultaneously, one eye looking directly at the blackboard, and the other looking through the telescope. From the relative sizes, one can judge magnification. Be sure to measure the distance between lens and compare this to the algebraic sum of their focal lengths (the diverging lens focal length is negative).

The theory of the telescope is that rays come in from a distant object and are focused at the focal point of the objective lens. This point is also the focal point of the eyepiece, so the rays come out of the eyepiece parallel to themselves. The eye focuses these parallel rays on the retina. The diagram shows the geometry for an astronomical telescope. The ratio of the outgoing angle $B$ to the incoming angle $A$ is called the angular magnification and for small angles this turns out to be equal to the ratio of the focal lengths:

$$
\operatorname{magnification}=\left(f_{\text {objective }}\right) /\left(f_{\text {eyepiece }}\right)
$$

Compare the results of your magnification measurements to values given by this formula for both the Galilean and astronomical telescopes.

$\tan A=\frac{H}{f_{0}}$
$\tan B=\frac{H}{f_{e}}$
$\frac{\tan B}{\tan A} \xlongequal{=}$ angular magnification $=\frac{f_{o}}{f_{e}}$

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[^1]:    DO NOT TOUCH THE PRISM. DO NOT ADJUST THE POSITION OF THE COLLIMATOR LENS, NOR THE OBJECTIVE LENS OF THE TELESCOPE. DO NOT DISTURB THE LEVELING SCREWS OF THE PRISM TABLE, TELESCOPE OR COLLIMATOR.

[^2]:    Before using the instrument, please read the following description. The first applies to the spectrometers made by Ealing Co., the second to those made by Gaertner.

