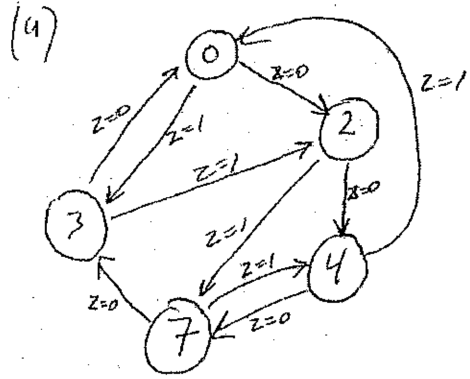
Ph.D. Preliminary Examination — Fall 2007 Solutions

Problem 1 (Core: CompE-ECE2030) — Code Number:__________

(a)



(b)

	ZABC	$A^+B^+C^+$
0	0000	010
1	0001	XXX
2	0010	100
3	0011	000
4	0100	111
5	0101	XXX
6	0110	XXX
7	0111	011
8	1000	011
9	1001	XXX
10	1010	111
11	1011	010
12	1100	000
13	1101	XXX
14	1110	XXX
15	1111	100

(c)

ZA \ BC	00	01	11	10
00	(0)	X (1)	(3)	1 (2)
01	1 (4)	X (5)	(7)	X (6)
11	(12)	X (13)	1 (15)	X (14)
10	(8)	X (9)	(11)	1 (10)

$D_A = A^+ = B\bar{C} + \bar{Z}A\bar{C} + ZAC$

$(\bar{Z}A\bar{B})\ (ZA\bar{B})$

ZA \ BC	00	01	11	10
00	1 (0)	X (1)	(3)	(2)
01	1 (4)	X (5)	1 (7)	X (6)
11	(12)	X (13)	(15)	X (14)
10	1 (8)	X (9)	1 (11)	1 (10)

$D_B = B^+ = \bar{Z}A + Z\bar{A} + \bar{Z}\bar{B}$

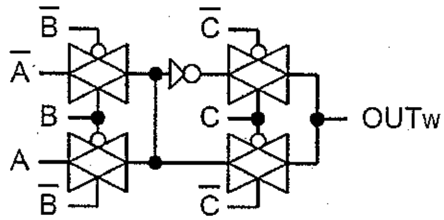
$(\bar{A}\bar{B})$

ZA \ BC	00	01	11	10
00	(0)	X (1)	(3)	(2)
01	1 (4)	X (5)	1 (7)	X (6)
11	(12)	X (13)	(15)	X (14)
10	1 (8)	X (9)	(11)	1 (10)

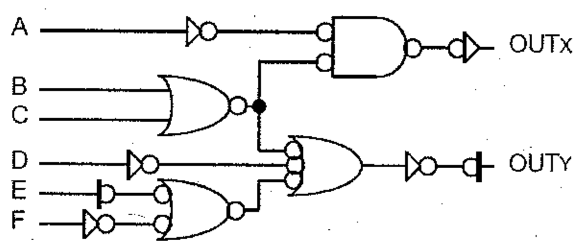
$D_C = C^+ = \bar{Z}A + Z\bar{A}\bar{C}$

Problem 2 (Core: CompE-ECE2030) **Code Number:__________**

For each part below, determine the behavior and write as a Boolean expression.

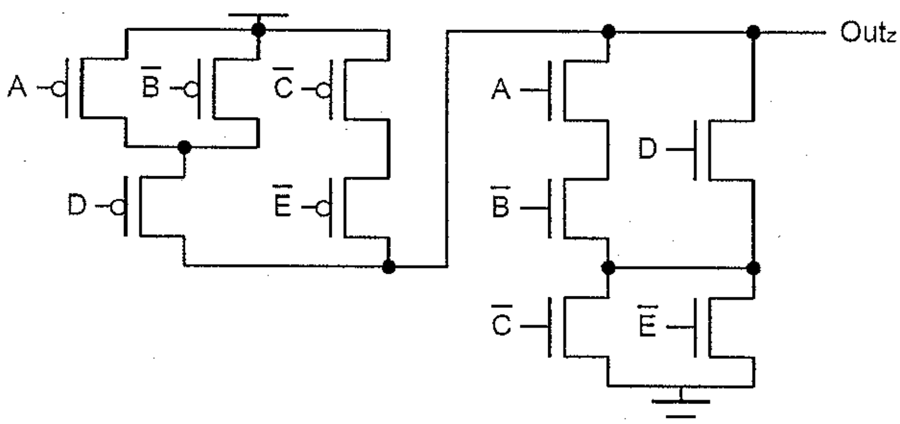


OUT_W = A XOR B XOR C



OUT_X = $A\cdot(B+C)$

OUT_Y = $\overline{B+C+D+\overline{E}+F}$



OUT_Z = $(\overline{A}+B)\cdot\overline{D}+C\cdot E$

Problem 3 (Core: CompE-ECE3055) Code Number:_________

Please fill hex for data signals and binary for control signals and 'x' for don't care.

Input to the "PC" in the IF stage

__0x4F010AB0___________________

"Read Data 1" in the ID stage

____0x8_________________________

Output of the MUX3 in the EX stage

__0x8 __________________________

"Read Data 2" in the ID stage

____0x2D2C2B2A ________________

"PCSrc" in the IF stage

__b'1___________________________

EX[3:1] (3 bits) in the EX stage

____b'110_______________________

OpMux1[1:0] in the EX stage

__b'10__________________________

OpMux2[1:0] in the EX stage

____b'01________________________

M[2:0] in the MEM stage

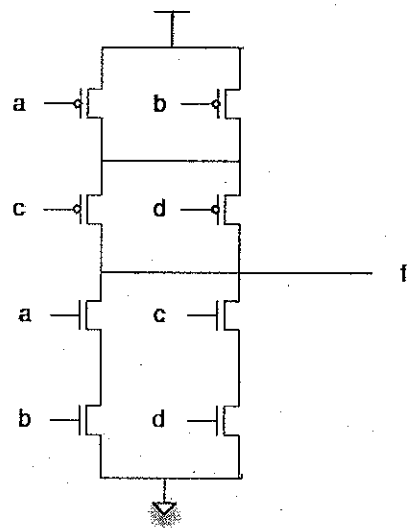
__b'1xx_________________________

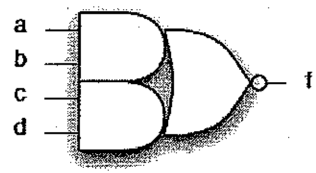
WB[1:0] in the WB stage

____ b'11_______________________

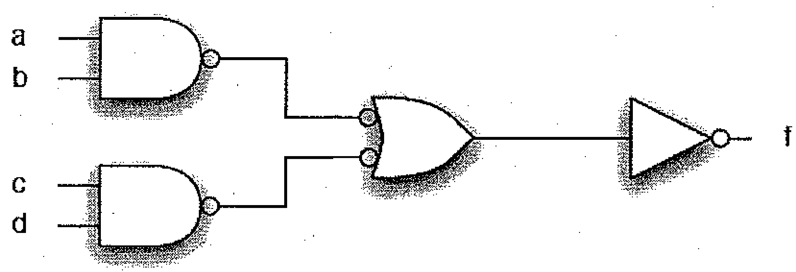
Problem 4 (Core: CompE-ECE3060) Code Number:_________

(a) (2 pt) Implement *f* using only one CMOS complex gate and some number of inverters. Give a transistor schematic and a gate schematic for this implementation. Your solution should minimize the total number of inverters used.





(b) (2 pt) Consider an alternate implementation using only two-input NAND gates and inverters. Draw a gate level schematic for this circuit.



(c) (4 pts) Derive delay expressions for both implementations above, assuming that i) they are sized to minimize delay, ii) all gates are designed to equalize rise and fall time in the worst case, and iii) the input load (on every input) is C_{inv}. Delay should be expressed in units of $\tau = RC$ where τ is the delay of a minimum size inverter driving a minimum size inverter. Also, the effect of parasitic delay should be accounted for.

The delay in part (a) is given by $D_a = (2C_L + 4)\tau$, since the logical effort of the AOI22 is 2, the electrical effort in the given circuit is C_L, and the parasitic delay is 4.

The delay in part (b) is given by $D_b = (3\hat{f} + 5)\tau$ where $\hat{f} = F^{\frac{1}{3}} = (GBH)^{\frac{1}{3}} = \left(\frac{4}{3}\right)^{\frac{2}{3}} C_L^{\frac{1}{3}}$

Note that the parasitic delay for this implementation is 5τ, and the path logical effort is $G = \left(\frac{4}{3}\right)^2$

(d) (2 pts) Qualitatively describe the conditions at which you would prefer one implementation over the other (with regard to delay).

Implementation (a) will be faster when the ideal number of stages is closer to one, i.e. approximately when $D_a \le (5.8 + 4)\tau$. As we increase C_L, at some point implementation (b) will become faster. The crossover point could be identified by setting $D_a = D_b$ and solving for C_L. For all larger values of C_L, implementation (b) will be faster.

Problem 5 (Core: E&M-ECE3025) Code Number:__________

Problem 1 : Answer

a) $\Gamma_L = \dfrac{100-j100-50}{100-j100+50} = \dfrac{50-j100}{150-j100} = \boxed{0.62\angle -29.74^\circ}$

b) $Z_{in} = Z_0\left[\dfrac{Z_L + jZ_0 \tan\beta l}{Z_0 + jZ_L \tan\beta l}\right]$

$f = 10^9\ Hz$

$\lambda = \dfrac{3\times 10^{10}}{10^9} = 30\ cms$ (air filled)

$\beta l = \dfrac{2\pi}{30}\times 5 = \dfrac{\pi}{3}$

$Z_{in} = 50\left[\dfrac{(100-j100)+(j50\times 1.732)}{50+j(100-j100)(1.732)}\right]$

$= 50\left[\dfrac{100-j13.4}{223.2+j173.2}\right] = \boxed{17.86\angle 314.56^\circ\ (\Omega)}$

c) $V_{in} = \dfrac{5\angle 0^\circ \times 17.86\angle 314.56^\circ}{17.86\angle 314.56^\circ + 50} = \boxed{1.4\angle -33.93^\circ\ (V)}$

$I_{in} = \dfrac{1.4\angle -33.93^\circ}{17.86\angle 314.56^\circ} = \boxed{0.078\angle -348.49^\circ\ (A)}$

d) $V_{in} = V^+\left[e^{j\beta l} + \Gamma_L e^{-j\beta l}\right]$

$1.4\angle -33.93^\circ = V^+\left[e^{j\pi/3} + 0.62\angle -29.74^\circ\ e^{-j\pi/3}\right]$

$= V^+\left[0.503 + j0.247\right]$

$\therefore V^+ = 2.5\angle -59.6^\circ$

$V_L = V^+(1+\Gamma_L) = 2.5\angle -59.6^\circ\,(1 + 0.62\angle -29.74^\circ)$

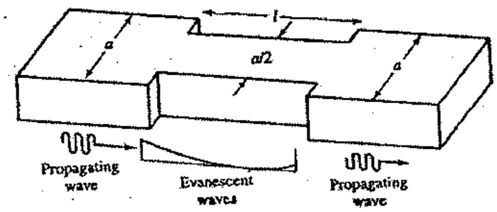
$\therefore \boxed{V_L = 3.92\angle 289.1^\circ\ (V)}$

e) $P_{avg} = \frac{1}{2} Re\left[V I^*\right] = \frac{1}{2} Re\left[2.5\angle -59.6^\circ \times 0.028\angle 25.9^\circ\right]$

$\therefore \boxed{P_{avg} = 29.1\ (mW)}$

Problem 6 (Core: E&M-ECE3065) Code Number: ____________

An attenuator can be made using a section of a rectangular waveguide operating below cut-off, as shown below. If a=2.286 cm and the operating frequency is 12 GHz, determine the required length of the below cut-off section of the rectangular waveguide to achieve an attenuation of 80 dB between the input and output guides. Ignore the effect of reflections at the step discontinuities.



In the section of guide with width $a/2$ the TE_{10} mode is below cut-off with an attenuation constant α:

$$\alpha = \sqrt{\left(\frac{\pi}{a/2}\right)^2 - k_o^2} \Rightarrow \quad k_o = \frac{2\pi f_o}{c} = \frac{2\pi \times 12 \times 10^9}{3 \times 10^8} = 251.3 \ m^{-1}$$

$$\alpha = \sqrt{\left(\frac{2\pi}{.02286}\right)^2 - (251.3)^2}$$

$$\alpha = 111.3 \ Np/M$$

To obtain 80 dB attenuation (ignoring reflections)

$$-80 dB = 20 \log e^{-\alpha l} \Rightarrow$$

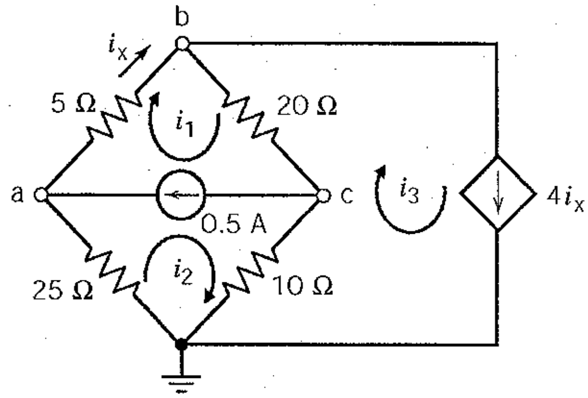
$$10^{-4} = e^{-\alpha l} \Rightarrow l = \frac{4 \ln 10}{\alpha} = \frac{9.21}{111.3} =$$

$$= 8.27 cm.$$

Problem 7 (Core: EDA-ECE2040) **Code Number:__________**

Solution:

Label the mesh currents.



Express i_x in terms of the mesh currents:

$$i_x = i_1$$

Express $4i_x$ in terms of the mesh currents:

$$4i_x = i_3$$

Express the current source current in terms of the mesh currents to get:

$$0.5 = i_1 - i_2 \quad \Rightarrow \quad i_2 = i_x - 0.5$$

Apply KVL to supermesh corresponding to the current source to get

$$5i_1 + 20(i_1 - i_3) + 10(i_2 - i_3) + 25i_2 = 0$$

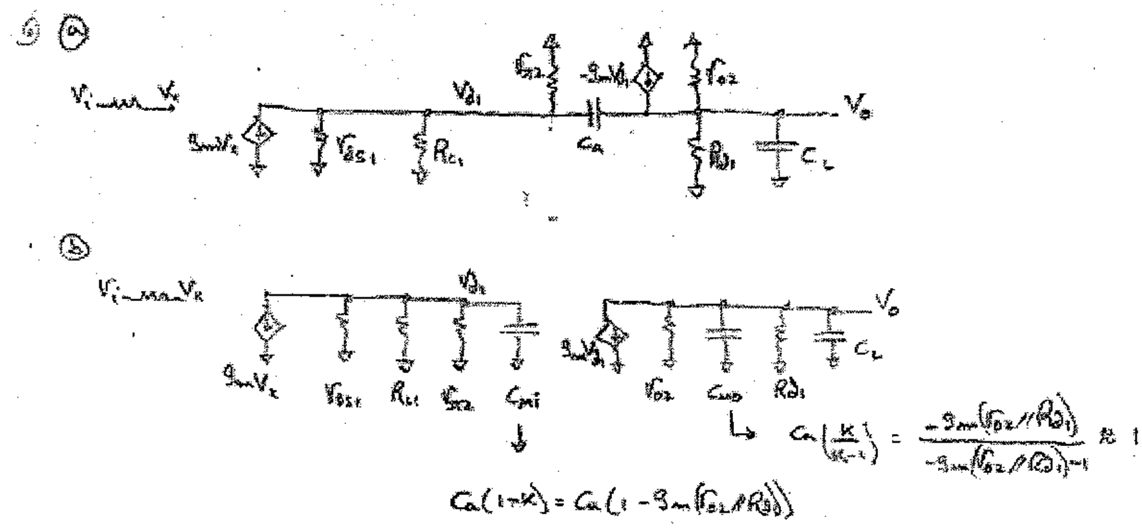
Substituting gives

$$5i_x + 20(-3i_x) + 10(i_x - 0.5 - 4i_x) + 25(i_x - 0.5) = 0 \quad \Rightarrow \quad i_x = -\frac{35}{120} = -0.29167$$

So the mesh currents are

$$i_1 = i_x = -0.29167 \quad \text{A}$$
$$i_2 = i_x - 0.5 = -0.79167 \quad \text{A}$$
$$i_3 = 4i_x = -1.1667 \quad \text{A}$$

Problem 8 (Core: EDA-ECE3050) **Code Number:**__________



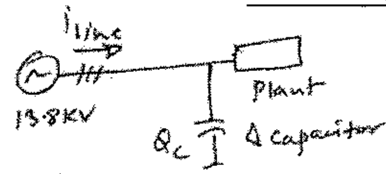
c) Give an expression for the dominant pole frequency of the circuit (i.e., the lowest high-frequency pole) using r_π, r_{o1}, g_{m1}, g_{m2}, r_{ds2}, R_{c1}, R_{d1}, C_c, and/or C_L. **(2.5 pts)**

C_{Miller}: $p_1 \approx 1/2\pi\{1+g_{m2}(R_{d1}||r_{o2})C_a\}(r_{\pi2}||R_{c1}||r_{ds1})$

d) Give an expression for the 2[nd] dominant pole (i.e., the next higher frequency pole) using r_π, r_{o1}, g_{m1}, g_{m2}, r_{ds2}, R_{c1}, R_{d1}, C_c, and/or C_L. **(2.5 pts)**

C_a shorts: $p_2 \approx g_{m2}/2\pi C_L$

Problem 9 (Core: Power-ECE3070) **Code Number:__________**



a) At 10MW @ 0.8PF

$P = 10\text{ MW} \quad Q = 7.5\text{ MVAR}$

$S = 12.5\text{ MVA}$

At 30MW @ 0.92PF

$P = 30\text{MW} \quad Q = 12.76\text{MVAR}$

$S = 32.6\text{ MVA}$

$$I_{Line} = \frac{32{,}600}{\sqrt{3} \times 13.8\text{KV}} = 1{,}364 \text{ Amps}.$$

b) Cap VARs for minimum VAR loading

$$Q_c = \frac{7.5 + 12.76}{2} = 10.1 \text{ MVAR}$$

$$\text{Each } \Delta \text{ Capacitor} = \frac{10.1}{3} = 3.66 \text{ MVAR} = \frac{V_{\ell-\ell}^2}{X_c} = \frac{[13.8 \times 10^3]^2}{X_c}$$

$$\therefore X_c = 56.5\,\Omega = \frac{1}{\omega C} = \frac{1}{377.C}$$

$$C = 46.9\,\mu F.$$

$$PF @ 10MW = \frac{10}{\sqrt{10^2 + 2.6^2}} = 0.968 \text{ leading}$$

$$PF @ 30MW = \frac{30}{\sqrt{30^2 + 2.6^2}} = 0.996 \text{ lagging}$$

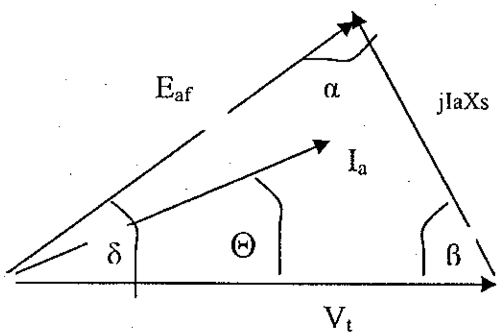
Problem 10 (Core: Power-ECE3070) **Code Number:__________**

(a) The output frequency of a synchronous generator is proportional to the speed. For a 4 pole 60 Hz machine, the speed is 1800 rpm.

(b) The mechanical speed is 188.5 rad/sec. With no losses the input shaft power is (2.5 x 0.8)MW

Thus torque is (2500 x 0.80)/188.5 = 10.6 k Nm

(c) Phasor diagram:



Per phase terminal voltage V_t = (6.6 x 1000)/1.732 = 3811 volts
Armature current per phase I_a = (2500 x 1000)/ (1.736 x 6600) = 218.7 A

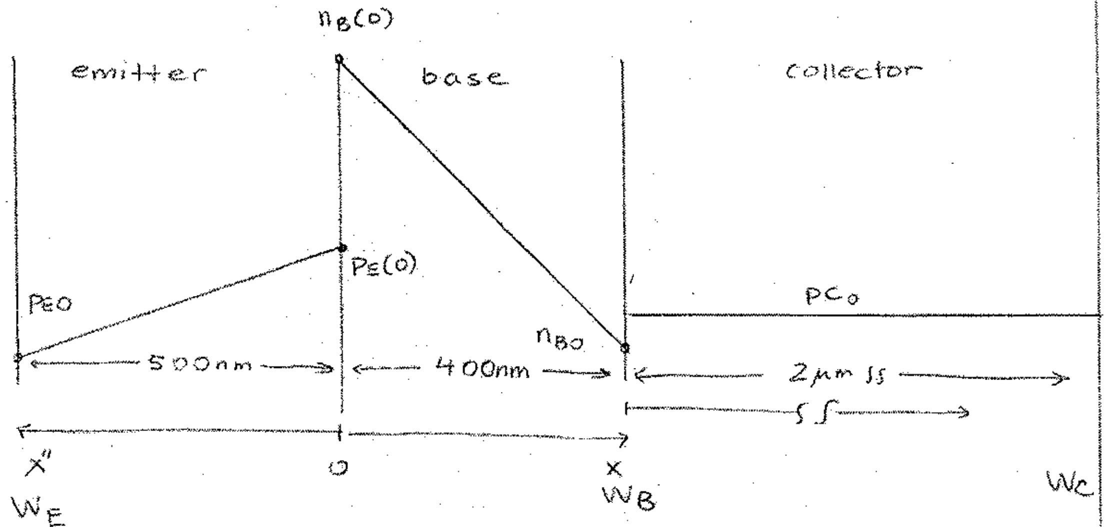
$\widetilde{E}_{af} = 3811 + j218.7(0.8 + j0.6)(10)$ = 3050 at an angle of 35 degrees. So the magnitude of this voltage is 3050 volts.

(d) The new value of the terminal voltage will be 3050 volts per phase.

A

Problem 11 (Core: Microsystems-ECE3040) Code Number:__________

The approximate sketch is



Collector: Since $V_{BC} = 0$, the excess carrier conc. is the same as equilibrium

using the law of the junction:

$$n_C(0) \cdot p_C(0) = n_i^2 e^{V_{BC}/V_T} = n_i^2$$

$$p_C(0) = p_{C0}$$

Base

Similarly for the electron concentration at the edge of the base on the collector side:

$$n_B(W) \cdot p_B(W) = n_i^2 e^{V_{BC}/V_T} = n_i^2$$

$$n_B(W) \cdot N_B = n_i^2 \quad \therefore \quad n_B(W) = n_{B0}$$

For the emitter base junction, the exact values can be determined, once we know V_{BE}

$$I_C = I_S e^{V_{BE}/V_T} \qquad e^{V_{BE}/V_T} = \frac{I_C}{I_S}$$

$$V_{BE} = V_T \ln\left[\frac{I_C}{I_S}\right]$$

Calculating $I_S = \dfrac{q D_B n_{BO} A_E}{W_B}$

$$I_S = \frac{(1.6\times10^{-19}C)\left(16\frac{cm^2}{s}\right)(50\mu m^2)\left(\frac{1cm}{10^{-4}\mu m}\right)^2 \overbrace{(6.7\times10^{2}cm^{-3})}^{n_{BO}}}{400\times10^{-7}cm}$$

$\boxed{I_S = 2.1\times10^{-17}A}$

$$V_{BE} = 0.026 \ln\left[\frac{25\mu A}{2.1\times10^{-17}A}\right]V = 0.026 \ln\left[1.2\times10^{12}\right] V$$

$\underline{V_{BE} = 0.722\ V}$

Back to calculating $n_B(0)$, $p_E(0)$ . . .

$$n_B(0) = n_{BO}\, e^{V_{BE}/V_T} = n_{BO}\cdot\frac{I_C}{I_S} = 6.7\times10^{2}cm^{-3}(1.2\times10^{12})$$

$\underline{n_B(0) = 8.0\times10^{14}\ cm^{-3}}$

$$p_E(0) = p_{EO}\cdot\frac{I_C}{I_S} = 11\ cm^{-3}\cdot 1.2\times10^{12} = \underline{13.2\times10^{12}cm^{-3}}$$

note exact values for $n_B(0)$, $p_E(0)$, $n_D(W_B)$ we not req'd but useful.

Problem 12 (Core: Microsystems-ECE3080) Code Number:__________

(a) $n = N_D = \underline{5 \cdot 10^{16}\ cm^{-3}}$ (majority carrier conc.)

$p = n_i^2/n = \underline{2 \cdot 10^{3}\ cm^{-3}}$ (minority carrier conc.)

(b) $n = N_c e^{-(E_c - E_F)/kT}$

$\Rightarrow E_c - E_F = -kT \ln\left[\frac{n}{N_c}\right] = kT \ln\left[\frac{N_c}{n}\right] = \underline{0.165\ eV}$

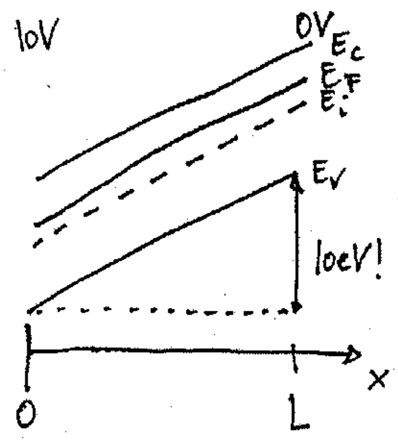
(c) $\rho = \frac{1}{q(\mu_n n + \mu_p p)} \approx \frac{1}{q\mu_n n} = \underline{0.127\ \Omega cm}$

(d) $R = \frac{V}{I} = 500\ \Omega$ required sample resistance $A = 85\ \mu m^2$

$R = \rho \frac{L}{A} \Rightarrow L = \frac{RA}{\rho} = \frac{V}{I}\frac{A}{\rho} = \underline{33.5\ \mu m}$

(e) $\rho = \frac{1}{q[\mu_n(n+\Delta n) + \mu_p(p+\Delta p)]} = \underline{0.034\ \Omega cm}$

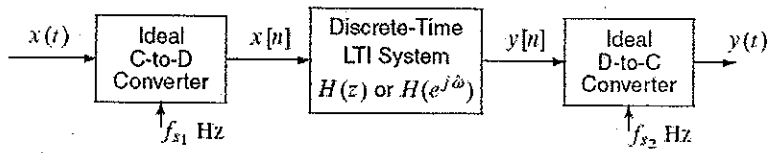
(f)



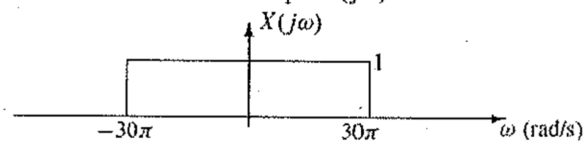
- Band bending because of applied potential
- E-field is constant throughout the device, i.e. E-diagram must be linear function of x!

Problem 13 (Core: DSP-ECE2025) Code Number:_________

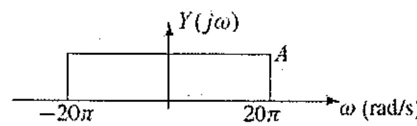
Consider the following system for continuous-time and discrete-time processing of a signal:



In all parts, the Fourier transform of the input $X(j\omega)$ is



(a) If $H(e^{j\hat{\omega}}) = 1$, and the Fourier transform of the output is $Y(j\omega)$:



If $f_{s_1} = 48$ Hz, then determine the rate of the D-to-C converter, f_{s_2}, as well as A. Note: $f_{s_2} \neq f_{s_1}$.

ANSWER: Since $f_{s_1} = 48$ Hz is greater than twice the highest frequency in $X(j\omega)$, i.e., greater than $2 \times 15 = 30$ Hz, there is no aliasing. Thus, it is sufficient to figure out how the input frequency component at $\omega = 30\pi$ rad/s becomes the output frequency component at $\omega = 20\pi$ rad/s.
The C-to-D converter takes $\omega = 30\pi$ rad/s and gives a frequency of $\hat{\omega} = 30\pi/f_{s_1} = 30\pi/48 = 5\pi/8$ for $x[n]$, which is also the same for $y[n]$ because $H(e^{j\hat{\omega}}) = 1$.
Thus, the output frequency will be $\omega = \hat{\omega} f_{s_2} \Rightarrow 20\pi = (5\pi/8) f_{s_2}$, and solving for the rate of the D-to-C converter, we obtain $\boxed{f_{s_2} = 32\,\text{Hz}}$.

There are several ways to find A, so here is one: we observe that in the time-domain $x(0)$ must equal $y(0)$. Since the $t = 0$ value in the time-domain is the area of the Fourier transform divided by 2π, we get

$$x(0) = y(0) \quad \Rightarrow \quad \frac{1}{2\pi}\int X(j\omega)d\omega = \frac{1}{2\pi}\int Y(j\omega)d\omega$$

$$\Rightarrow \quad 1(30\pi - (-30\pi)) = A(20\pi - (-20\pi)) \quad \Rightarrow 60\pi = A(40\pi) \quad \Rightarrow \boxed{A = 1.5}$$

(b) In this part, assume that $H(z) = z^{-3}$ and $f_{s_1} = f_{s_2} = 33$ Hz; then determine a closed-form mathematical expression for the output time signal, $y(t)$.

ANSWER: Since $f_{s_1} = 33$ Hz is greater than twice the highest frequency in $X(j\omega)$, i.e., greater than 30 Hz, there is no aliasing. If $H(e^{j\hat{\omega}})$ were equal to one, then $y(t) = x(t)$, BUT when $H(z) = z^{-3}$, the discrete-time LTI system is a delay of 3 samples, i.e., $y[n] = x[n-3]$. At a sampling rate of 33 Hz, 3 samples corresponds to a time delay of $3/33 = 1/11$ secs. Thus we conclude that $y(t) = x(t - 1/11)$. Taking the inverse Fourier transform of $X(j\omega)$, which is a rectangle, gives a "sinc function."

$$x(t) = \frac{\sin(30\pi t)}{\pi t} \quad \Rightarrow \quad \boxed{y(t) = x(t - 1/11) = \frac{\sin(30\pi(t - 1/11))}{\pi(t - 1/11)}}$$

(c) In this part, assume that the discrete-time LTI system, $H(e^{j\hat{\omega}})$, is a highpass filter (given below), and that the output Fourier transform $Y(j\omega)$ is given by the formula below.

$$H(e^{j\hat{\omega}}) = \begin{cases} 0 & |\hat{\omega}| \le \hat{\omega}_c \\ 1 & \hat{\omega}_c < |\hat{\omega}| \le \pi \end{cases} \qquad Y(j\omega) = \begin{cases} 0 & |\omega| \le 20\pi \\ 1 & 20\pi < |\omega| \le 30\pi \\ 0 & |\omega| > 30\pi \end{cases}$$

If $f_{s_1} = f_{s_2} = 50$ Hz, then determine the cutoff frequency of the highpass filter, $\hat{\omega}_c$.

ANSWER: Since $f_{s_1} = 50$ Hz is greater than twice the highest frequency in $X(j\omega)$, i.e., greater than 30 Hz, there is no aliasing. Therefore, we can use the concept of an *effective filter* to state that the *effective frequency response* from $x(t)$ to $y(t)$ is:

$$H_{\text{eff}}(j\omega) = H(e^{j\hat{\omega}})\Big|_{\hat{\omega}=\omega/f_{s_1}} = H(e^{j\omega/50})$$

In other words, the relationship between the effective analog cutoff frequency and the cutoff frequency of the digital filter is

$$\hat{\omega} = \omega/50 \qquad \Rightarrow \qquad \omega = 50\hat{\omega}$$

which means that the *effective filter* has a passband from $\omega = 50\hat{\omega}_c$ to $\omega = 50(\pi)$, because the highpass digital filter has its passband from $\hat{\omega} = \hat{\omega}_c$ to $\hat{\omega} = \pi$.

In order to obtain the desired output Fourier transform which extends from 20π to 30π rad/s, we need the lower cutoff frequency of the passband to be at $\omega = 20\pi$, so we solve

$$50\hat{\omega}_c = 20\pi \qquad \Rightarrow \qquad \boxed{\hat{\omega}_c = 0.4\pi}$$

The upper frequency of 30π in $X(j\omega)$ lies in the passband of the *effective filter*, so the upper frequency in $Y(j\omega)$ will also be 30π rad/s.

Problem 14 (Core: DSP-ECE3075)   Code Number:__________

a) $f_X(x) = \exp\{-\lambda|x|\} \quad -\infty < x < \infty$

$\int_{-\infty}^{\infty} \exp\{-\lambda|x|\}dx = 2\int_0^{\infty} \exp\{-\lambda x\}dx = \frac{2}{\lambda} = 1$ therefore, $\lambda = 2$

b) $Y = |X| \qquad y = \begin{cases} x, & x \geq 0 \\ -x, & x < 0 \end{cases} \qquad \frac{dx}{dy} = \begin{cases} 1, & x \geq 0 \\ -1, & x < 0 \end{cases} \qquad f_X(x) = \exp\{-2|x|\} \quad -\infty < x < \infty$

$f_Y(y) = (|1|)\exp\{-2y\} + (|-1|)\exp\{-2y\} = 2\exp\{-2y\} \quad 0 \leq y < \infty$

c) $W = \frac{1}{3}(X + 2|X|) \qquad w = \begin{cases} x, & x \geq 0 \\ -\frac{x}{3}, & x < 0 \end{cases}, \quad w \geq 0$

$f_W(w) = u(w)\left[3\exp\{-6w\} + \exp\{-2w\}\right]$ where $u(\bullet)$ is the step function.

d) $Z = \frac{X}{2Y} = \begin{cases} \frac{1}{2}, & x \geq 0 \\ -\frac{1}{2}, & x < 0 \end{cases}$

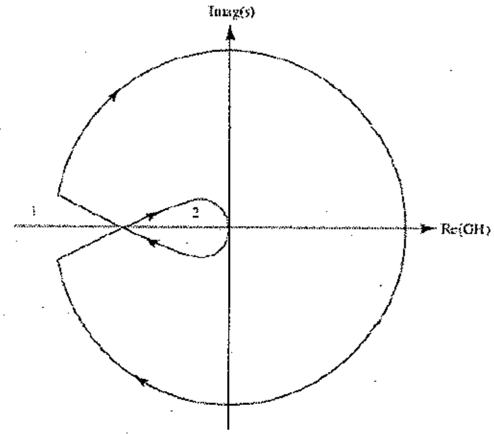
Since $f_X(x) = \exp\{-2|x|\}, -\infty < x < \infty$, is symmetric with respect to 0, $\Pr\left\{Z = \frac{1}{2}\right\} = \Pr\left\{Z = -\frac{1}{2}\right\} = 0.5$

$\overline{Z} = E\{Z\} = -\frac{1}{2}\times 0.5 + \frac{1}{2}\times 0.5 = 0 \qquad \overline{Z^2} = E\{Z^2\} = \left(\frac{1}{2}\right)^2 \times 0.5 + \left(-\frac{1}{2}\right)^2 \times 0.5 = \frac{1}{4}$

Problem 15 (Core: S&C-ECE3085) Code Number:__________

Solution

a. Define $GH_1(s) = GH(s)|_{K=1}$. The Nyquist plots of $GH(s)$ and $GH_1(s)$ are of the form



b. Since

$$
\begin{aligned}
GH_1(j\omega) &= -\frac{j\omega+1}{\omega^2(j\omega+10)(j\omega+20)} \\
&= -\frac{j\omega+1}{\omega^2(200-\omega^2)+j30\omega^3} \\
&= -\frac{j\omega+1}{\omega^2(200-\omega^2)+j30\omega^3}\cdot\frac{\omega^2(200-\omega^2)-j30\omega^3}{\omega^2(200-\omega^2)-j30\omega^3} \\
&= -\frac{\omega^2(200-\omega^2)+30\omega^4+j(\omega^3(200-\omega^2)-30\omega^3)}{\omega^4(200-\omega^2)^2+900\omega^6}
\end{aligned}
$$

the Nyquist plot of $GH_1(s)$ crosses the negative real axis for $s = j\sqrt{170}$, and

$$
\begin{aligned}
GH_1(j\sqrt{170}) &= -\frac{170(30)+30(170^2)}{170^2(30^2)+900(170^3)} \\
&= -\frac{1+170}{170(30)+30(170^2)} \\
&= -\frac{1+170}{170(30)(1+170)} \\
&= -\frac{1}{5100}
\end{aligned}
$$

The closed-loop system will be stable if and only if the Nyquist plot of $GH_1(s)$ does not pass through or encircle the point $s = -\frac{1}{K}$, which will be the case provided that $s = -\frac{1}{K}$ lies between $-\infty$ and $-\frac{1}{5100}$, i.e. if $0 < K < 5100$. If instead $s = -\frac{1}{K}$ lies between $-\frac{1}{5100}$ and 0, i.e. if $5100 < K < \infty$, then the Nyquist plot will encircle the critical point twice, so there will be 2 RHP poles and 2 LHP poles. If instead $s = -\frac{1}{K}$ lies between 0 and $+\infty$, i.e. if $-\infty < K < 0$, then the Nyquist plot will encircle the critical point once, so there will be 1 RHP pole and 3 LHP poles.

An alternative to the above solution is to apply a Routh test to the characteristic polynomial

$$s^4 + 30s^3 + 200s^2 + Ks + K$$

which also shows that the closed-loop system is stable if and only if $0 < K < 5100$.

Problem 16 (Core: S&C-ECE3085) Code Number:__________

Step 1:
Open loop Transfer Function:

$$KGH = \frac{K(s+1)}{s^2(s+3.6)}$$

zeros: $s = -1$
poles: $s_1 = -3.6$ $\qquad s_2 = 0$ (double pole)

Step 2:
Asymptote Centroid

$$\sigma_n = \frac{\sum poles - \sum zeros}{n-m} = \frac{(0-3.6)-(-1)}{3-1} = \frac{-2.6}{2} = -1.3$$

⇨ asymptotes intersect at =1.3

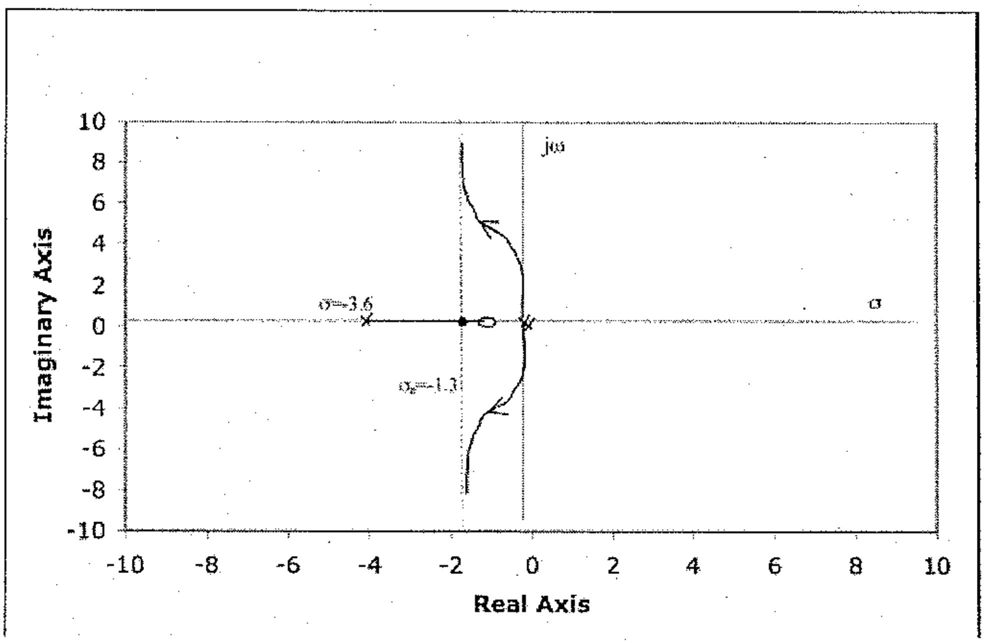
Step 3:
Asymptote Angles

$$\sigma_a = \frac{(2q+1)180^\circ}{n-m}$$

$$a_0 = \frac{180}{2} = 90^\circ$$

$$a_1 = \frac{540}{2} = 270^\circ$$

=> Root locus has only one breakout pt at s=0 and no jw axis crossings



Using Magnitude Criterion

$$K = \frac{1}{|GH|}\Big|_{s_1=-0.75+j1.7} = \left|\frac{s_1^2(s_1+3.6)}{(s_1+1)}\right| = \frac{|-0.75+j1.7|^2|2.85+j1.7|}{|.25+j1.7|}$$

$$= \frac{|1.858|^2|3.318|}{1.718} = \frac{11.454}{1.718} = 6.7$$

K = 6.7

Problem 17 (Specialized: Comp Science-CS3210) Code Number:______

Mutual Exclusion. Assume that we have an multi–threaded application using the Unix/Linux *pthreads* library, and that we have created 10 threads using `pthread_create`. Further assume that the system we are running on has *only one CPU*. Part of our application requires exclusive access to a particular section of code called the *Critical Section*. Exclusive access means that only one thread can execute in the critical section at any point in time. We decide to implement this exclusive access using a global *Locked* flag, as follows:

```
1  // Define the global locked flag.  0 means unlocked, 1 means locked
2  int Locked = 0;
3
4  // The code for the critical section.
5  void CriticalSection()
6  {
7    // Wait until any other thread having the lock releases it
8    while(Locked == 1)
9      {
10       // Do nothing here, or perhaps sleep for a small amount of time
11     }
12
13   // No other thread in the critical section, set the locked flag
14   Locked = 1;
15
16   // Now the code for the actual critical section
17   // Ommited here
18
19   // We are now done with the critical section, unlock the flag
20   Locked = 0;
21 }
```

Program q1-code.cc

At first glance, the locking mechanism above seems reasonable and correct, but when we test our program it fails mysteriously and at apparently random times.

1. Explain clearly what the problem is in the code snippet above. Refer to the lines of code by line number as needed.

   Answer: The code will fail if a thread gets interrupted after the while loop at line 8 exits, but before the lock flag is set at line 14. A second thread could enter the critical section, see the lock flag is not set, and assume it can enter the critical section. While the second thread is in the critical section, that thread is pre–empted and the first thread re-schedules. It has already checked the locked flag, and thus enters the critical section.

2. Explain what could be done to fix the problem, without a complete re-write of the code. If you need to make assumptions about the CPU instruction set, state what assumption you made.

   Answer: The best solution is to use an atomic *test–and–set* instruction (if such is supported by the CPU architecture). This will return the value of a flag, and set the flag in a single, non–interruptable operation. The while loop at line 8 becomes:

   ```
   while(TestAndSet(&Locked) == 1)   // do nothing
   ```

   And line 14 is no longer needed. If an atomic test–and–set instruction is not available on the CPU, then a more complex solution, such as Lamport's Bakery algorithm, is needed.

Problem 18 (Specialized: Software Sys- ECE3035) Code Number: _____

Prelim Problem Solutions **Computing Mechanisms**

Below is a snapshot of heap storage in byte addressed memory. Values that are pointers are denoted with a "$". The heap pointer is $6128. The heap has been allocated contiguously beginning at $6000, with no gaps between objects. Objects are word-aligned. Each allocated object has a header word containing the size of the object in bytes. The address of the object points to the word just after the header (i.e., the first word allocated for the object's data). For example, the object pointed to by $6004 has size 16 bytes, so the next allocated object in the heap is of size 8 bytes and is pointed to by $6024.

addr	value	addr	value	addr	value	addr	value	addr	value
6000	16	6032	12	6064	8	6096	12	6128	0
6004	33	6036	28	6068	4	6100	$6092	6132	0
6008	$6100	6040	24	6072	$6092	6104	16	6136	0
6012	16	6044	$6092	6076	8	6108	0	6140	0
6016	$6080	6048	12	6080	$6004	6112	12	6144	0
6020	8	6052	$6080	6084	0	6116	0	6148	0
6024	25	6056	16	6088	4	6120	0	6152	0
6028	52	6060	$6004	6092	$6004	6124	0	6156	0

Part A Suppose the stack holds a local variable whose value is the memory address $6036. No registers or static variables currently hold heap memory addresses. List the addresses of all objects in the heap that are *not* garbage.

Addresses of Non-Garbage Objects: **$6036, $6092, $6004, $6100, $6080**

Non-garbage objects are color-coded above (green indicates an object header containing size info in bytes and blue indicates the space allocated for the object's data). The address of each object is the address of the first word of data allocated for the object. Non-garbage objects are those reachable from the root address $6036.

Part B Create a free list by scanning the memory for garbage, starting at address $6000 and pushing each reclaimed object on the front of the free list. List the addresses of the objects (in order) on the free list at the end of the scan.

Free List: **$6116, $6068, $6052, $6024**

Part C Based on the free list created in part B, If an object of size 7 bytes is allocated, what address will be returned as a pointer to the newly allocated object using a first-fit allocation strategy?

Address: **$6116**

Part D Based on the free list created in part B, if an object of size 13 bytes is allocated, what address will be returned using a *best-fit* allocation strategy?

Address: **$6132**

None of the objects on the free list (i.e., those that are garbage) are large enough to hold 13 bytes. So the new object is allocated at the top of the heap. Its size info goes in $6128 and $6132 is the pointer returned to the newly allocated object.

Part E If the local variable whose value is the address $6036 is popped from the stack, which addresses listed in Part A will be reclaimed by each of the following strategies? If none, write "none."

Reference Counting:	**$6036 (the rest are involved in cyclic references)**
Mark and Sweep:	**$6036, $6092, $6004, $6100, $6080**

Problem 19 (Specialized: Telecom-ECE3076) Code Number:__________

Solution:

In order for CSMA/CD to always provide better performance than CSMA. it needs to always detect a collision before the corresponding transmission is complete.

The time taken for transmitting a frame T_t is given by:

$$T_t = \frac{P}{B} \text{ seconds}$$

The time taken for the worst case delay before a station will realize that there is a collision is two times the propagation delay between the stations separated by the largest distance, which is:

$$T_p = \frac{2*D}{c} \text{ seconds}$$

Thus, for CSMA/CD to always provide better performance than CSMA,

$T_t > T_p$

O

$$\frac{P}{B} > \frac{2*D}{c}$$

Problem 20 (Specialized: Optics-ECE4500) Code Number:_________

Fraunhofer Diffraction by Two Slits

With $d = 4a$, the transmittance is

$$f_2(x) = \begin{cases} 1 & \frac{3a}{2} < |x| < \frac{5a}{2} \\ 0 & \text{elsewhere} \end{cases}$$

Solution No. 1

$f_2 \; = \; f_a(x) \; + \; f_b(x)$ for the two slits taken individually

The Fourier transforms are obtained using $t \rightarrow x$, $\omega \rightarrow k_x$, and $\tau \rightarrow a$ and where $k_x = \frac{2\pi}{\lambda} sin\theta$.

$$F_2(jk_x a) \; = \; F_a(jk_x a) \; + \; F_b(jk_x a)$$

Using

$$F(jk_x a) \; = \; F[f(x-d)] \; = \; F(jk_x a)\, e^{-jk_x d}$$

then

$$F_a(jk_x a) \; = \; F[f_1(x-2a)] \; = \; F_1(jk_x a)\, e^{-j2k_x a}$$

$$F_b(jk_x a) \; = \; F[f_1(x+2a)] \; = \; F_1(jk_x a)\, e^{+j2k_x a}$$

therefore

$$F_2(jk_x a) \; = \; F_1(jk_x a)\, [e^{+j2k_x a} + e^{-j2k_x a}]$$

$$F_2(jk_x a) \; = \; 2\, F_1(jk_x a)\, cos\, (2k_x a)$$

$$F_2 \; = \; 2a\, \frac{sin\left(\frac{k_x a}{2}\right)}{\left(\frac{k_x a}{2}\right)}\, cos(2k_x a)$$

Solution No. 2

Alternatively

$f_2(x) = f_{21} - f_{22}$

with

$$f_{21} = 1 \qquad \frac{-5a}{2} < x < \frac{5a}{2}$$

$$f_{22} = 1 \qquad \frac{-3a}{2} < x < \frac{3a}{2}$$

The Fourier transforms are obtained using $t \rightarrow x$, $\omega \rightarrow k_x$, and $\tau \rightarrow a$ and where $k_x = \frac{2\pi}{\lambda} sin\theta$.

$$F_{21}(jk_x a) = 5a \frac{sin(\frac{5k_x a}{2})}{\frac{5k_x a}{2}}$$

$$F_{22}(jk_x a) = 3a \frac{sin(\frac{3k_x a}{2})}{\frac{3k_x a}{2}}$$

$$F(jk_x a) = F_{21} - F_{22}$$

$$F_{21} - F_{22} = 5a \frac{sin(\frac{5k_x a}{2})}{\frac{5k_x a}{2}} - 3a \frac{sin(\frac{3k_x a}{2})}{\frac{3k_x a}{2}}$$

$$F_{21} - F_{22} = a \frac{1}{\frac{k_x a}{2}} [sin(\frac{5k_x a}{2}) - sin(\frac{3k_x a}{2})]$$

$$sin(\frac{5k_x a}{2}) = sin(\frac{4k_x a}{2} + \frac{k_x a}{2}) \equiv sin(\frac{4k_x a}{2})cos(\frac{k_x a}{2}) + cos(\frac{4k_x a}{2})sin(\frac{k_x a}{2})$$

$$sin(\frac{3k_x a}{2}) = sin(\frac{4k_x a}{2} - \frac{k_x a}{2}) \equiv sin(\frac{4k_x a}{2})cos(\frac{k_x a}{2}) - cos(\frac{4k_x a}{2})sin(\frac{k_x a}{2})$$

and so

$$F_{21} - F_{22} = 2a \frac{sin(\frac{k_x a}{2})}{(\frac{k_x a}{2})} cos(2k_x a)$$

or

$$F_2 = 2a \frac{sin(\frac{k_x a}{2})}{(\frac{k_x a}{2})} cos(2k_x a)$$

where $k_x = \frac{2\pi}{\lambda} sin\theta$.

Problem 21 (Specialized: Optics-ECE4501) Code Number:__________

A fiber communication link consists of three sections of single-mode fiber that are joined end-to-end to form a net transmission span of 14 km. The three segments are:

1. a 2.0-km length having dispersion $D_1 = +16.0$ ps/nm-km, and loss coefficient, $\alpha_1 = 0.20$ dB/km;
2. a 6.0-km length having dispersion $D_2 = -5.0$ ps/nm-km, and loss coefficient, $\alpha_2 = 0.15$ dB/km;
3. a 6.0-km length having dispersion $D_3 = +4.1$ ps/nm-km, and loss coefficient, $\alpha_3 = 0.15$ dB/km.

a. Assuming that all spilce losses are negligible, what is the net power loss of the link in decibels?

$$\text{Loss (dB)} = 0.20 \times 2.0 + 0.15 \times 6.0 + 0.15 \times 6.0 = \underline{2.2 \text{ dB}}$$

b. Suppose the fibers are all to be replaced by a single fiber, having the same overall length. What value of D should the new fiber have, such that the link dispersion penalty is unchanged?

$$D_{net} = \frac{D_1 L_1 + D_2 L_2 + D_3 L_3}{L_1 + L_2 + L_3} = \frac{16.0(2.0) + (-5.0)(6.0) + 4.1(6.0)}{14}$$

$$= \underline{1.9 \text{ ps/nm-km}}$$

c. In the original three-fiber link, an optical source provides an input average power of 10 dBm. What is the link output power in mW?

$$P_{out}(\text{dBm}) = 10 \text{ dBm} - 2.2 \text{ dB} = 7.8 \text{ dBm}$$

$$P_{OUT}(\text{mW}) = 10^{0.78} = \underline{6.0 \text{ mW}}$$

d. The input power in mW is now doubled, so that the span distance can be increased. Assuming that additional lengths of the original three fibers are available, make a choice as to the best fiber to use *to add to the existing link*, and find the new span length (such that the output power is the same as in part c). The new link should be as long as possible. What other factor(s) (if any) influenced your decision?

Doubled input power means that an additional 3dB of loss is allowed. Lowest loss fibers are 2 and 3 (both with 0.15 dB/km). Choose fiber 2 because it has negative dispersion, and will thus minimize the net dispersion penalty: have $3 \text{ dB} = 0.15(\Delta L) \Rightarrow \Delta L = 20.0 \text{ km}$ | Overall length: $L' = L + \Delta L = \underline{34 \text{ km}}$

Problem 22 (Specialized: Microsystems-ECE4752) Code Number:_____

Pre deposition

$$D_1 = D_0 e^{-E_a/kT} = 10.5\, e^{-3.6/(8.62\times10^{-5})(1173)}$$

$$D_1 = 3.62\times10^{-15}\ cm^2/sec$$

$$Dose = Q_T = 2 C_S \sqrt{\frac{D_1 t_1}{\pi}} = 2\,(1.1\times10^{20})\sqrt{\frac{(3.62\times10^{-15})\,900}{\pi}}$$

$$\boxed{Q_T = 2.24\times10^{14}\ cm^{-2}}$$

Drive-in

$$X_j = \sqrt{4 D_2 t_2 \ln\left(\frac{Q_T}{C_B\sqrt{\pi D_2 t_2}}\right)}$$

where $D_2 = 10.5\, e^{-3.6/(8.62\times10^{-5})(1373)}$

$$D_2 = 6.47\times10^{-13}\ cm^2/sec$$

$$X_j = \sqrt{4(6.47\times10^{-13})(18000)\ln\frac{2.24\times10^{14}}{(3\times10^{16})\sqrt{\pi(6.47\times10^{-13})(18000)}}}$$

$$\boxed{X_j = 4\times10^{-4}\ cm = 4\ \mu m}$$

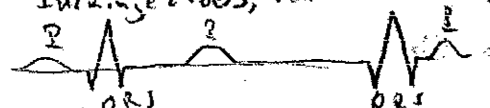
Problem 23 (Specialized: Bio Eng-ECE4784) Code Number:__________

a) (3 pts) Describe the proper sequence of firing events for the heart.

b)(4pts)Sketch the anticipated ECG/EKG in the case where there is an AV block, i.e. the atrial muscle and the AV node lose their electrical connection.

c)(4pts)Explain, using mathematical arguments why the p wave in the case where the SA node fails to fire will be inverted. Be sure to use the curve for the membrane voltage of the atrial muscle as a function of distance in your answer.

4. a) S/A node, atrial muscles, A/V node, bundle of His, bundle branches, Purkinje Fibers, ventricle muscles

b) 

time between P & QRS is uncorrelated

c) atrial contraction will travel in the opposite direction to the normal so the potential will be determined by a ∇V_m that points in the opposite direction

Problem 24 (Specialized: Bio Eng-ECE4781) Code Number:__________

(3 pts) What type of electrode is used to measure the resting membrane potential of a neuron? Draw a sketch and label the parts of the electrode.

(2 pts) What are the most important specs of the Differential Amplifier used to measure the membrane potential of a neuron?

Noise is not such a big problem, so common-mode rejection is not very important. It's the current levels that could be a cell-killer. The input impedance needs to be huge.

(3 pts) The Nernst Equation estimates the membrane potential across a neuron as a function of ion concentrations. What happens to the membrane potential if the absolute temperature increases by 10%?

According to Nernst, Em is proportional to the absolute temperature, so Em will also increase by 10%.

(2 pts) What physiological mechanism could cause this change?

The Na/K pump could transport sodium and potassium ions across the membrane at a faster rate, but Nernst probably conceived that the permeability of the membrane would change at higher temperatures.

Problem 25 (Specialized: Bio Eng-ECE4782) Code Number:__________

Pretend that you have accepted a senior research position at a company called Mortem Ltd, and they have asked you to design a noninvasive, portable diagnostic device that will prove that a comatose patient on a ventilator is legally dead. The selling price of this medical device needs to be less than $10,000.

(3 pts) What biological signal(s) should you measure and explain your reasoning?

Testing for a pupil response using a flash of light won't work because the patient could be blind. The same thinking applies to an auditory response. EKG will verify that the heart is still able to contract, but some patients on ventilators will continue to have heart contractions when the brain is dead. Most medical doctors accept brain activity as the primary indicator of life.

Unfortunately, surface EEG is only able to detect neural activity a few mm below the skull. How can we be sure that there is no neural activity deep inside the brain? Should we drill several small holes through the skull and probe deep into the brain? This would be dangerous and definitely invasive. Magneto Encephalogram would be a good option, since deeper neural activity can be detected, but this would require expensive gradiometers. I like the idea of using a combination of surface EEG electrodes measuring evoked responses on the scalp plus an ERG electrode on each eye measuring the same nonvisual evoked responses. The electrode on each eye would have a relatively low resistance pathway to the brain stem.

(4 pts) What types of stimuli should you use? Remember that the RMS Signal-to-Noise levels could be really small.

The best approach would be to use a battery of several different stimuli, e.g. an auditory stimulus (e.g. a loud click of sound), a tactile stimulus (e.g. a solenoid striking the skin), and an electrical stimulus next to a peripheral nerve. The important detail will be to use a large number of large-amplitude, pulse stimuli, so that it will be possible to perform ensemble averaging. Since the patient is immobile, the pulse stimuli could be repeated for several hours. White-noise stimuli could be used for auditory and electrical stimuli. If none of the stimuli produce a measurable response, the last type of stimulus could test the pain pathway, e.g. a device that mechanically pinches the skin at some location a large number of times. In this situation, you could forget about pressure ulcers.

(3 pts) If one of the above stimuli is Gaussian and White-noise modulated, write an equation that shows how to calculate the first-order Wiener Kernel.