Electrical Engineering

Quals Questions

2002

Stephen Boyd

Optimal design of a three-point constellation.

You are given the matrix $A \in \mathbf{R}^{m \times n}$.

How would you find three points $u,\ v,\ w \in \mathbf{R}^n$, with

$$\|u\|_2 \leq 1, \quad \|v\|_2 \leq 1, \quad \|w\|_2 \leq 1,$$

that maximize the minimum distance between Au, Av and Aw?

Explain completely.

Ask me about anything that is not clear.

Cioffi - Quals 2002 Solutions

$x = \begin{cases} \pm 1 \\ \pm 3 \end{cases}$ [illegible]

a). $E[x] = \frac{1}{4}(1-1+3-3) = 0$

b). $E[x^2] = \mathrm{var}(x) = \frac{1}{2}(1+9) = 5$

$$y_k = (x_k + x_{k-1} + x_{k-2} + x_{k-3})\,1/4$$

c). y_k approximates $E[x]$

d). $E[y_k] = \frac{1}{4}\,4 \cdot 0 = 0$

e). $\mathrm{var}(y_k) = \frac{1}{16}\sum_{i=0}^{3} \mathrm{var}(x_{k-i}) = \frac{1}{16}\cdot 4\ 5 = 5/4$

f) y_k is $\frac{1}{4}$ (sum of 4 odd numbers) $= \frac{1}{4}$ (even number)

max is $12/4$ min is $-12/4$

$y_k \in \frac{1}{4}[-12\ -10\ -8\ -6\ -4\ -2\ 0\ 2\ 4\ 6\ 8\ 10\ 12]$

<u>13</u> values

g). $r_{x,k} = 5\,\delta_k$

h). $r_{y,k} = h_k * h_{-k} * r_{x,k}$

$$= \frac{5}{16}\left([1\,1\,1\,1] * [1\,1\,1\,1]\right) = \begin{array}{cccccccc} & -3 & -2 & -1 & 0 & +1 & +2 & +3 \\ \frac{5}{16} & [1 & 2 & 3 & 4 & 3 & 2 & 1] \end{array}$$

i). $$p_y = \underbrace{\underbrace{\underbrace{p_x}_{\frac{1}{4}(1\,1\,1\,1)} * p_x}_{} * \underbrace{p_x * p_x}_{}}_{}$$

$\frac{1}{16}[1\,2\,3\,4\,3\,2\,1]$

$$\frac{1}{256}\,[1\ 4\ 10\ 20\ 31\ 40\ 44\ 40\ 31\ 20\ 10\ 4\ 1]$$

2002 Quals — J. Cioffi

TOTAL - 10 pts

Random Variables, Processes and Linear Systems - 10 pts

A random variable x takes on the 4 values $\pm 1 \quad \pm 3$ with equal probability.

a). What is the mean value of x? (.5 pt)

b). What is the variance of x ? (1 pts)

Independent selections of this random variable at different discrete points in time, k, form the stationary random process x_k. Another random process is computed according to

$$y_k = \frac{x_k + x_{k-1} + x_{k-2} + x_{k-3}}{4} \quad .$$

c). What does the process y_k approximate? (.5 pt)

d). Find $E[y_k]$. (1 pt)

e). Find the variance of y_k (1 pt)

f). How many distinct values are there for the random process y_k ? (1 pt)

g). What is the autocorrelation function of x_k, $r_{x,k} = E[x_n \cdot x_{n-k}]$? (1 pt)

h). What is the autocorrelation function of y_k ? (2 pts)

i). What is the probability distribution of y_k ? (2 pts)

X-Sender: billd@171.64.66.149
Date: Tue, 22 Jan 2002 10:55:40 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: Bill Dally <billd@csl.stanford.edu>
Subject: Re: Quals Meeting Today

Quals Question:

1. What oblivious routing function gives the highest worst-case throughput for a 7-node ring topology. All links have 1 unit of bandwidth in each direction.

- students invariably suggest a shortest path function which is incorrect

1a - What is the worst-case traffic pattern for this routing function - all send 3 nodes clockwise

1b - Have them compute the throughput of their routing function - 1/3

1c - Can they come up with a routing function that does better than shortest path

this sorts the students out - the better ones see that some traffic needs to go the long way around the ring.

the better students come up with a splitting algorihtm

1d - what is the throughput of this algorithm

2 - Same question for 7 x 7 torus topology

Only two students got this far.

At 09:26 AM 1/22/2002 -0800, you wrote:

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X-Authentication-Warning: theforce.Stanford.EDU: localhost.Stanford.EDU [127.0.0.1] didn't use HELO
protocol
To: Diane Shankle <shankle@ee.Stanford.EDU>
Subject: My qual question
Date: Tue, 22 Jan 2002 11:24:27 -0800
From: David Dill <dill@theforce.Stanford.EDU>


1. Suppose you are given a sparse directed graph represented so that
   we can get a list of successors for each vertex.  What is the fastest
   way you can think of to check whether the reflexive transitive closure
   would be a partial order?

  If students didn't know what a partial order was: "Suppose you have
  a directed graph which has one vertex from which all others can
  be reached.  What is a good algorithm for searching for a cycle.?"

2.  NP completeness

    What is a basic strategy for proving a problem NP-complete?

    Suppose you have an encryption scheme with the following properties:
      * Public key -- encryption is fast, inverse encryption is hard
                      but decrypting is fast if you have the private key.
      * Encryption function is injective

    How hard can it be to invert the encryption function (assume messages
            are are about the same size as encrypted messages)?

3. Suppose you have a tree which is represented in the computer so
   that the parent of each node can be accessed in constant time.
   What is a good algorithm for checking whether one node is an ancestor of another?

   Now suppose tree doesn't change, and we're going to make a lot of queries,
   so it would be worth doing some pre-processing if we could speed up the queries.
   What kind of pre-processing would be useful?
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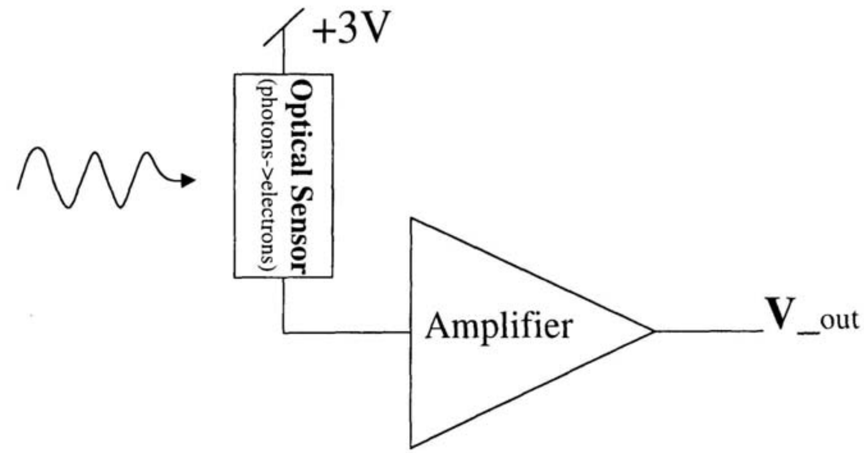


Consider the block diagram of a desired system to convert an incoming flux of photons into an output voltage signal.

- Explain what kind of an optical sensor you will use (and related details of how it works, circuit model for it etc.) and
- What kind of a circuit will you use to appropriately detect and amplify the electronic signal

Qualification exam, January 14-18, 2002, by

Consider two dielectric regions, each filled with a dielectric material with a dielectric constant ε_1 and ε_2, respectively. The interface is located at $z = 0$, as shown in Figure 1. An electromagnetic wave, with a frequency ω, is incident from the ε_1 region at $z < 0$ onto the interface at an incidence angle θ_1. Under most circumstances, part of the power is transmitted and part of the power is reflected.

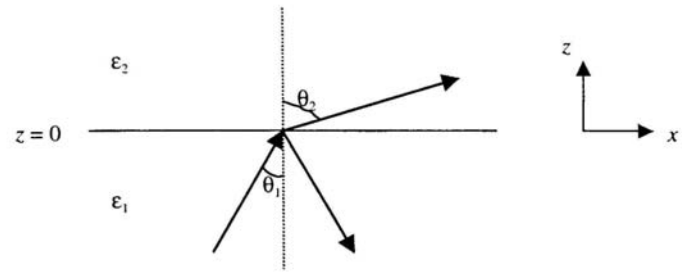


Figure 1

1. What is the relation between θ_1 and θ_2?

2. Under what conditions will there be no transmission? (i.e. what is the condition under which total internal reflection occurs?)

3. When total internal reflection does occur, how does the electromagnetic fields vary quantitatively as a function of z in the region $z > 0$?

Instead, consider the situation as shown in Figure 2. Let's assume that the incident angle and the materials are chosen such that the total internal reflection condition is satisfied at the interface $z = 0$.

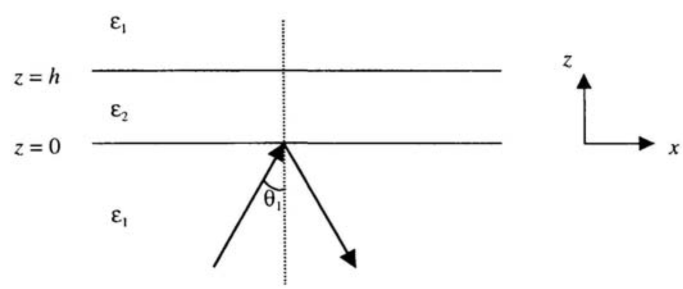


Figure 2

4. Do you expect any transmitted wave in the region $z > h$? If you do, why? What direction do you think the transmitted wave will be propagating at? How will the transmission coefficient relate to the thickness h of the ε_2 region?

OFFICE MEMORANDUM ◊ STAR LABORATORY

January 30. 2002

To: Diane Shankle

From: Tony Fraser-Smith

Subject: Ph.D. Quals Question, 2002

Penetration of Low-Frequency Electromagnetic Fields

The student is presented with three thin metal plates and asked about their electric and magnetic properties. Two of the plates are aluminum, with one about four times the thickness of the other. The other plate is steel; it is about the same thickness as the thinner aluminum plate. The student is very briefly asked about electrical conductivity σ, electrical permittivity ϵ ($= \epsilon_o$ in free space), and magnetic permeability μ ($= \mu_0$ in free space). Inevitably they can write the formula for skin depth $\delta = \sqrt{2/\omega\mu\sigma}$, i.e., the distance over which an electomagnetic wave of angular frequency ω will propagate through a conductor before its amplitude declines to $1/e$ of its initial value.

A strong horseshoe magnet is produced; the steel keeper is removed, and placed on the desk. The student is asked to check the metal plates to see if they are magnetic (the steel plate is) and then asked if the magnet will attract the keeper through the three different plates. In the subsequent discussion involving skin depth (δ) the student will need to recognize that ω is very small whereas σ is quite large for metals and μ is also quite large for the steel plate. Having discussed the situation and hopefully with the student having demonstrated some ability to think as an engineer, we carry out a test and find that the keeper is attracted in all cases, but the strength of attraction decreases as follows: (1) thin aluminum plate, strongest; (2) thick aluminum plate; (3) steel plate (weakest). This is exactly what would be expected if the skin depth was a major factor in the attractive force.

The student is now asked if there is any other reason for the strength of attraction between the magnet and its keeper being weaker when the thicker aluminum sheet is used. At this stage the student should realize that the distance between the magnet and the keeper is greater for the thicker sheet and that as a result the attraction could be weaker for that reason alone. A this time he/she might question the skin depth argument, but the situation can be rescued by the student comparing the attractive forces when the thin aluminum and the steel plates are used. Since they are the same thickness the different compositions as clearly an important factor.

Finally, the magnet is moved around on one side of the thin aluminum plate and it is seen how the keeper tracks the motion on the other side of the plate. The point here is that there has to be a component of attractive force parallel to the surface of the plate for the keeper to move and this requires the magnet to be offset relative to the keeper.

X-Sender: hector@db.stanford.edu
Date: Fri, 15 Feb 2002 16:44:16 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: Hector Garcia-Molina <hector@cs.stanford.edu>
Subject: Re: Quals Question 2002

At 04:26 PM 2/15/2002 -0800, you wrote:
> Please turn in your Quals Question 2002. You can send email or a hard copy

Here is my question...
hector

Hector Garcia-Molina
EE Quals Question 2002

The Fibonacci sequence is 1, 1, 2, 3, 5, 8, 13, ...
Each term is computed as the sum of the previous two terms.

(1) Write a function F(I) that computes the Ith number in the sequence. For example, F(6) should return 8.
First write the function using recursion, and then using iteration.

(2) Estimate the amount of main memory used by each of the two function implementations. For your estimate, just consider the number of variables allocated (1 unit for each), either on the stack, or in global space.

(3) Estimate the computation costs of each implementation. Count only the number of additions performed by each.

-------- Original Message --------
Subject: Re: [Fwd: Quals Question 2002]
Date: Mon, 25 Feb 2002 11:15:35 -0800
From: James Gibbons <gibbons@cis.stanford.edu>
To: Mary Cloutier <cloutier@cis.stanford.edu>
References: <3C73E3B3.934413AE@cis.stanford.edu>

Diane:

Here is the order of the questions I used for the quals.

QUESTION 1:

1. Explain how you would build a switching system using parts you can purchase at a hardware store that would allow you to change the state of a stairway light (from off to on or on to off) from either of two switches (one at the bottom of the stairwell and one at the top).

2. Can you construct a truth table for this operation?

3. How would you implement the truth table in CMOS?

4. How would you arrange the circuit so that it could be driven by capacitively coupled touch panels rather than switches that you mught buy at a hardware store?

QUESTION 2:

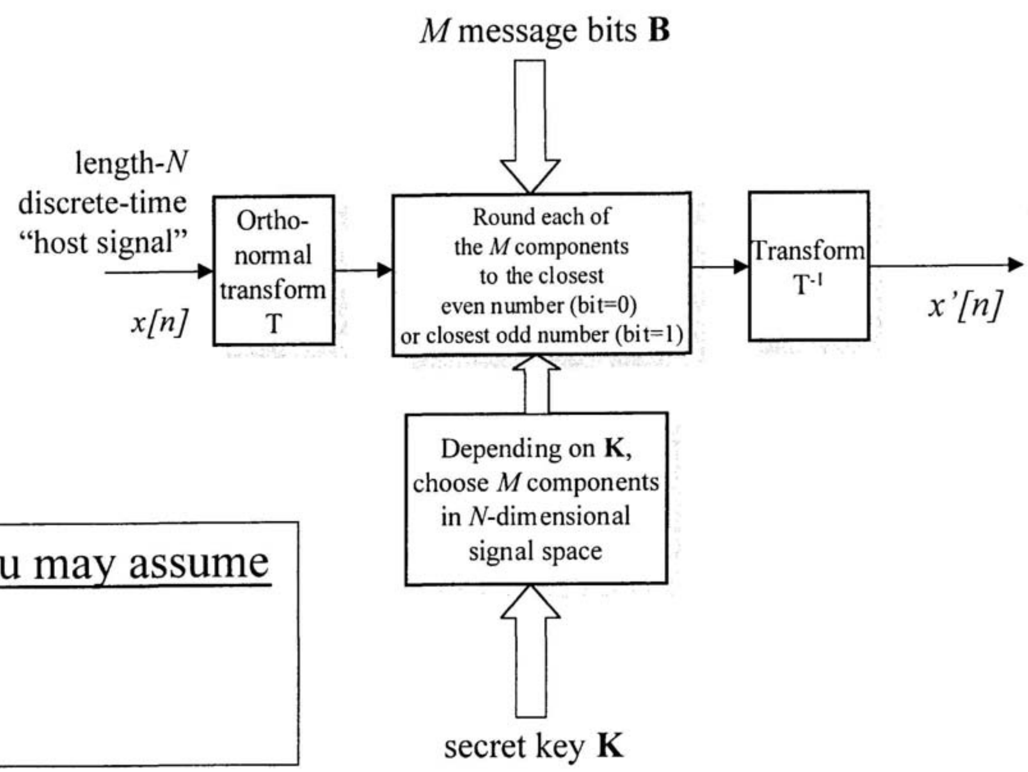
If the candidate did not make much progress on this line of questions (most did), I then asked them the following series of questions:

1. How does a photovoltaic cell work?

2. What are the factors that determine the ultimate limit of efficiency for such a cell.

3. How practical such cells might be for generating power at prices that would be competitive with the power grid.

Information hiding scheme

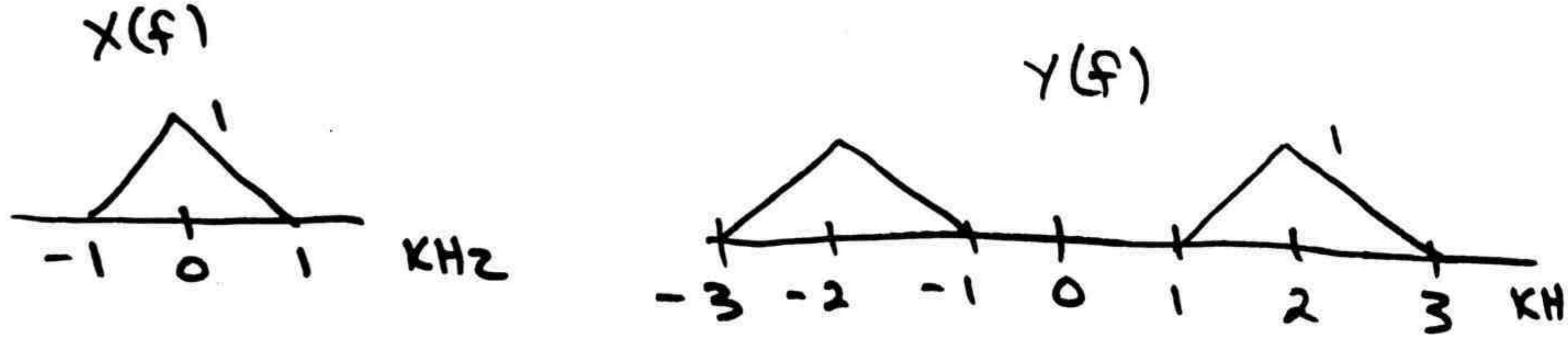


You may assume

- Calculate the mean-squared error between $x[n]$ and $x'[n]$
- Sketch the blockdiagram of a circuit that extracts the hidden message bits **B** from $x'[n]$, using key **K** (but not original signal $x[n]$)
- Devise one or more "attacks" that, without knowing key **K**, render the message bits unreadable, without substantially impairing the host signal $x[n]$

Andrea

1



(a) $x(t) \longrightarrow \boxed{?} \longrightarrow y(t)$

How is $y(t)$ generated from $x(t)$?

(b) What minimum sampling rate can be used such that $y(t)$ can be recreated from its samples

2) Let X and Y be independent random variables uniformly distributed between 0 and 5

What is $p(X+Y<3)$?

3

$$x(t) \longrightarrow \boxed{h(t) = a\delta(t - t_0)} \longrightarrow y(t)$$

$x(t)$ is a stationary Gaussian process with mean zero and PSD $\text{sinc}(f)$

a) $\{x(1), x(1.2), x(2.5)\}$

b) $y(t)$

c) $\int_0^T y(t)\,dt$

4 Why does sending data at a high data rate require a lot of bandwidth?

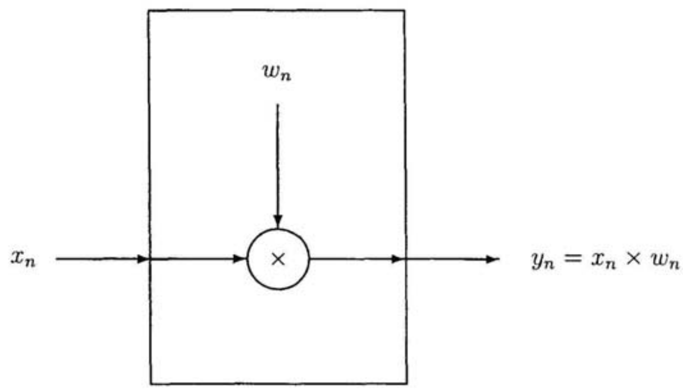
January 2002: R.M. Gray's Qualifier Question

My rough point guidelines are given below. The points were adapted to how well I thought the student performed in solving a problem as well as to whether or not they solved it. A problem might count more if the student showed particular creativity or adaptivity in solving it.

A system $\mathcal{S}$ has as input a real-valued discrete-time (DT) signal $\{x_n; n = \ldots, -1, 0, 1, 2, \ldots\}$ and output $\{y_n; n = \ldots, -1, 0, 1, 2, \ldots\}$ where

$$y_n = x_n \times w_n \text{ for all } n,$$

where $\{w_n; n = \ldots, -1, 0, 1, 2, \ldots\}$ is a fixed real-valued DT signal.



- Is the system $\mathcal{S}$
  - linear?
  - causal?
  - time invariant?

Solution

- The system is linear since if input $\{x_n^{(i)}\}$ results in output $\{y_n^{(i)} = x_n^{(i)} \times w_n\}$ for $i = 1, 2$, then the sequence $ax_n^{(1)} + bx_n^{(2)}$ will result in the corresponding linear combination $\{ay_n^{(1)} + by_n^{(2)}\}$
- The system is causal since if two input sequences agree up until a time n, then the corresponding output sequences must also agree up until that time. In other words, the current and past outputs cannot be changed by future inputs. Note the popular definition of haveing a delta response be 0 for negative delay is not appropriate here since this system is not time invariant and the output is not a convolution of the input with a delta response.

- For the system to be time-invariant we would need to know that if input $\{x_n\}$ produces $\{y_n\}$, then for any delay N $\{x_{n+N}\}$ will produce $\{y_{n+N}\}$. For this to hold we must have tha the system output when $\{x_{n+N}\}$ is the input, $x_{n+N}w_n$, must be the same as $\{y_{n+N} = x_{n+N}w_{n+N}\}$ for all n and N. This can only be true if $w_n = w_{n+N}$ for all N and n, i.e., if w_n is a constant. So the answer is no unless w_n is a constant.

I counted this part only one point and accepted very short answers. The question was to get things started with easy basics. If someone showed clear difficulty or got answers wrong, then I took more time to see if they understood these concepts.

Define a (discrete-time) Fourier transform pair for the signal $\{x_n\}$ as

$$X(f) = \sum_{n=-\infty}^{\infty} x_n e^{-j2\pi fn}$$

$$x_n = \int_0^1 X(f)e^{j2\pi fn}\,df$$

where $j = \sqrt{-1}$.
$Y(f)$ and $W(f)$ are defined similarly.
- Find an expression for $Y(f)$ in terms of $X(f)$ and $W(f)$.

Solution This is a standard manipulation and the question aimed at making sure people could correctly set up and prove this basic property of Fourier analysis. In particular, this and similar proofs involve a substitution and an interchange of integral and sum. There are many different ways to either derive the answer from scratch, or to prove the result if the answer is known. If a student chose a method that got messy fast, I looked for them to try variations or alternaive approaches that yielded more simple math.
A simple proof is the following:

$$\begin{aligned}
Y(f) &= \sum_{n=-\infty}^{\infty} x_n w_n e^{-j2\pi fn} \\
&= \sum_{n=-\infty}^{\infty} \left(\int_0^1 df' X(f')e^{j2\pi f'n}\,df'\right) w_n e^{-j2\pi fn} \\
&= \int_0^1 X(f') \left(\sum_{n=-\infty}^{\infty} w_n e^{-j2\pi(f-f')n}\right) df' \\
&= \int_0^1 X(f')W(f-f')\,df'
\end{aligned}$$

Many students tried substituting for both x_n and w_n, which led to a much messier solution involving an infinite sum of complex exponentials. This does not produce an elementary proof, but it will work if the sum is recognized as the generalized Fourier series of a Dirac delta sequence. For those who remembered the convolution formula correctly, one could replace either X or W or both by their representations as the Fourier transforms of x_n and y_n and it was easy to show the convolution yielded $Y(f)$.
A somewhat trickier approach tried by a few people, but no one got quite right, was to focus on focus on the fact that

$$y_n = x_n \times w_n = \int_0^1 Y(f)e^{j2\pi fn}\,df$$

and show that substituting X and Y into the product using the inverse Fourier transform produced a similar integral with the convolution of X and W replacing Y, which from the uniques of Fourier transforms proves the result. For example.

$$\begin{aligned} x_n y_n &= \left(\int_0^1 X(u)e^{j2\pi un}\,du\right) \times \left(\int_0^1 W(f')e^{j2\pi f'n}\,df'\right) \\ &= \int_0^1 X(u)e^{j2\pi un}\left(\int_0^1 W(f')e^{j2\pi f'n}\,df'\right)du \end{aligned}$$

In the inner integral make the substitution $f' = f - u$, $df' = df$:

$$\begin{aligned} x_n y_n &= \int_0^1 X(u)e^{j2\pi un}\left(\int_u^{1+u} W(f-u)e^{j2\pi(f-u)n}\,df\right)du \\ &= \int_0^1 X(u)e^{j2\pi un}\left(\int_0^1 W(f-u)e^{j2\pi(f-u)n}\,du\right)df \end{aligned}$$

since $W(f-u)e^{j2\pi(f-u)n}$ is a periodic function of f with period 1. Rearranging and interchanging the order of integration

$$x_n y_n = \int_0^1 \left(\int_0^1 W(f-u)X(u)\,du\right)e^{j2\pi fn}\,df$$

which implies the term inside the parentheses is $Y(f)$. Only one person came close to making this work.
I counted this part generally as about 3 points for fundamental Fourier knowledge.

- Suppose we know that the input sequence $\{x_n\}$ has the property that

$$e^u = \sum_{n=0}^{\infty} x_n u^n \text{ for all complex } u \quad (\star)$$

Find a simple description of the input sequence $\{x_n\}$ and its Fourier transform
$X(f) = \sum_{n=-\infty}^{\infty} x_n e^{-j2\pi fn}$

Solution This question was intended to require some thought and to not just check standard basic Fourier stuff. Alone it has nothing to do with Fourier analysis, but it tests basic calculus (which is an important part of systems theory and basic engineering) and provides a step needed for the final part of the question.
Students familiar with basic calculus should recognize this is a power series (or McLauren series or Taylor series) for the exponential and could just write down

$$x_n = \frac{1}{n!}\frac{d^n}{du^n}e^u \mid_{u=0} = \frac{1}{n!}.$$

Many students saw this quickly by recalling that the Taylor series for an exponential is

$$e^u = \sum_{n=0}^{\infty} \frac{u^n}{n!}$$

Those not seeing the connection had to think a bit and possibly get hints. The best start is to notice that if $u = 0$, the equation forces $x_0 = 1$ since all u^n are 0 in this case except for $n = 0$.

This suggests that all values of x_n can be found if only the factor for x_n is nonzero. This can be accomplished by differentiating, e.g., differentiating once yields

$$\frac{d}{du}e^u = e^u = \sum_{n=0}^{\infty} n x_n u^{n-1}$$

so that setting $u = 0$ forces $x_1 = 1$. Differentiating twice yields

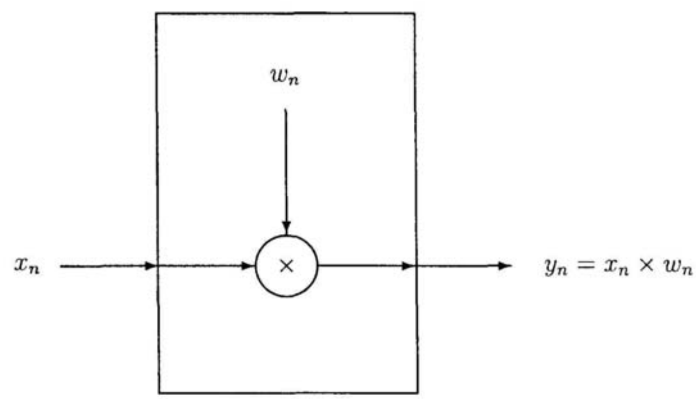
$$\frac{d^2}{du^2}e^u = e^u = \sum_{n=0}^{\infty} n(n-1)x_n u^{n-2}$$

forcing $2x_2 = 1$. Contininuing in this way results in $n!x_n = 1$ as before.
The condition says nothing about x_n for negative n, so it could be chosen arbitrarily. Choosing it to be all 0, the transform is simply ($\star$) with $u = e^{-j2\pi f}$ so that

$$X(f) = e^{e^{-j2\pi f}}$$

This fact could be recognized without first finding x_n if one observed that $x_n = 0$ for $n < 0$ was required.
I counted this part 3-4 points.

Final question:



Suppose that

$$x_n = \begin{cases} \frac{1}{n!} & n = 0, 1, 2, \ldots \\ 0 & \text{otherwise} \end{cases}$$

$$w_n = \begin{cases} 2^{-n} & n = 0, 1, 2, \ldots \\ 0 & \text{otherwise} \end{cases}$$

Find $W(f)$ and $Y(f)$.
Solution:

$W(f)$ follows from a geometric series.

$$W(f) = \sum_{n=0}^{\infty} 2^{-n} e^{-j2\pi fn} = \frac{1}{1 - e^{j2\pi f}/2}$$

$Y(f)$ could be found by convolving X and W, but here it is much easier to stay in the time domain:

$$y_n = \begin{cases} \frac{2^{-n}}{n!} & n = 0, 1, 2, \ldots \\ 0 & \text{otherwise} \end{cases}$$

so that

$$\begin{aligned} Y(f) &= \sum_{n=-\infty}^{\infty} y_n e^{-j2\pi fn} \\ &= \sum_{n=-\infty}^{\infty} \frac{2^{-n}}{n!} e^{-j2\pi fn} \\ &= \sum_{n=-\infty}^{\infty} \frac{(2e^{j2\pi f})^{-n}}{n!} \\ &= e^{e^{-j2\pi f}/2} \end{aligned}$$

using $(\star)$ with $u = (2e^{j2\pi f})^{-1}$.
I counted this part 2-3 points.

2002 PhD Quals Question
J. S. Harris

1. Can you draw an energy band diagram for a metal-n-type semiconductor junction?
   A. Explain what happens when you apply a bias voltage and draw the I-V Characteristic
   B. Explain what happens in the device and the resulting effect on the I-V characteristic if the metal layer is thin and optically transparent and I shine light on it with Energy, $E > E_g$ of the semiconductor.

2. What happens if I put a thin insulator (say 500Å) between the metal and semiconductor?
   A. What are the two main things that the insulating layer does?
   B. Since no current can flow, an I-V characteristic is not particularly insightful. What measurement might you make and what information can it provide?
   C. What is the difference between the low and high frequency C-V curves?
   D. How would the C-V Characteristic change if I shine light on the interface? (similar to M-S device, transparent metal and $E > E_g$)

3. If I use this basic structure to form an MOS transistor:
   A. Would shining light through the gate into the channel region have any effect on the gate C-V characteristic or the Drain I-V characteristics? Why or why not? Would it make any difference if the gate were biased above or below threnshold?

X-Sender: horowitz@vlsi.stanford.edu
Date: Wed, 13 Mar 2002 22:26:46 -0800
To: shankle@ee.Stanford.EDU
From: Mark Horowitz <horowitz@Stanford.EDU>
Subject: Quals question (finally)
Cc: penny Chumley <penny@csl.Stanford.EDU>

The quals question was based on building the interface to a dot-matrix display

1. Assume you have an array of 8x128 lights. If you had a wire to control each light how many wires would you need?

1024

2. Clearly that is too many wires. If we want to reduce the wires what can we do?

Need to use memory in the display, and use the fact that the eye is slow. So scan out image a column at a time. This will take 8 wires (one for each row) and either 7 wires for the columns (decode locally), or just two wires -- one for the clock and one for reseting the counter.

3. Now assume that I want to have an analog display rather than just turing the lights on or off. How could I do that?

Simplest solution is to drive an analog voltage on each row line. The light on that row's amplitude will be set but that control.

4. Assume that the actual picture elements are intrinsically digital. Can we still create an analog display?

Sure, use pulse width modulation. Pulse the row so the pulse width is proportional to the desired intensity

5. How does the clock on my desk work?

G. Kovacs' Quals Question – 2001/2002

The student was asked to sketch the design, on paper, of an analog circuit (amplifier, oscillator, filter, etc.) that used three NPN bipolar transistors and any passive components the student wished. The student was told that it was up to them to choose the circuit parameters. It was requested that the circuit not be overly complex, and that the exact component values were not wanted – rather, the student was asked to discuss the design process. Students who were not familiar with BJT's were told that they should design with MOS transistors and the examiner would help them "translate" their design. The student was told that giving rough design equations would be helpful, but that it was key to explain the desired properties and operation of the circuit.

EE Quals questions, ~~systems area~~; Winter 2002. S. Lall 1

We will be investigating stability of the linear system

$$x(t+1) = A\,x(t)$$

where at each time t, $x(t)$ is an n-dimensional vector

$$x(t) = \begin{pmatrix} x_1(t) \\ \vdots \\ x_n(t) \end{pmatrix}$$

and A is an $n \times n$ matrix.

Questions

1) What does stability mean?
   Hint: first consider the case when $x(t)$ and A are scalars.
2) How would you test the system for stability?
3) Is your condition necessary and sufficient?
4) Give me an example 2×2 matrix A for which the above system is stable.
5) Suppose $A = \begin{bmatrix} a & b \\ 0 & d \end{bmatrix}$. What values of a, b, d result in a stable system?

Question

5) Now consider the generalization

$$x(t+1) = A(t)\,x(t)$$

where
$$A(t) = \begin{cases} F_1 & \text{if } t \text{ is odd} \\ F_2 & \text{if } t \text{ is even} \end{cases}$$

and F_1 and F_2 are both stable, in the sense of question 3) above.

Is it stable?

7) If yes, prove it, if no, give a counterexample.

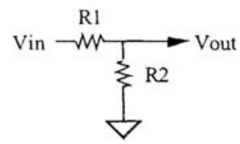
8) Can you think of a condition on F_1 and F_1 ~~test~~ for stability of the system?

9) Is your condition necessary and sufficient?

NAME: Thomas Lee Ph.D. Qualifying Exam, 2002

Problem 1: Consider a simple resistive voltage divider:

FIGURE 1. Resistive divider



What is the relationship between the output and input voltages?

Ans:

$$V_{out} = V_{in}\frac{R_2}{R_1 + R_2} \quad \text{(EQ 1)}$$

Can resistor R1 have a negative value? If yes, explain what happens if R1's magnitude exceeds that of R2. If no, explain why not.

Ans: Resistors can certainly have negative values. If R1's magnitude exceeds that of R2, however, the net resistance seen by the driving source is negative. Depending on the assumed nature of the driving impedance, this negative resistance can result in instability.

If this possibility is ignored, the voltage divider equation predicts that the network produces an inversion in signal polarities.

Can a negative resistor be physically realized as an element with a total of no more than two terminals? Explain.[Clarification: There is a completely sealed black box, out of which only two wires emerge. Can there appear between those two terminals a negative resistance that functions forever?]

Ans: A negative resistor can supply energy. As such, there needs to be an energy source somewhere. Since it's not allowed to be in the box, then more than two terminals are needed.

X-Sender: lusignan@ee.stanford.edu
Date: Tue, 22 Jan 2002 20:24:58 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: Bruce Lusignan <lusignan@ee.Stanford.EDU>
Subject: Re: Quals Meeting Today

Bruce Lusignan's Quals question was:

A) We have a synchronous communications satellite serving the United States with a 10 Watt transmitter and a transmit antenna beam that is 4 degrees by 8 degrees wide. We have a second satellite serving India. Because India is smaller it can be served with a transmit antenna beam that is 4 degrees by 4 degrees. What power is needed to serve India? The receive systems are the same for the two countries.

Ans: The link equation shows that the power required is inversely proportional to the transmit gain. The gain is inversely proportional to the beam width squared, or the product of the major and minor axes if the beam is elliptical. The Indian satellite therefore has twice the gain and requires half the transmit power. 1/2 of 10 watts is **5 watts.**

B) In the US satellite system the 10 watts transmitter power was based on using a 10 ft diameter parabolic antenna for the ground receiver. If we switch the receive antenna to 5 feet diameter, how much power is needed from the satellite?

Ans: The link equation shows that the received power is proportional to the area of the receive antenna. since the diameter is reduced by 2, the area is reduced by 4. The power therefore has to be increased by 4 to **40 watts.**

C) The US satellite with 10 watts ignored the effect of rain loss. In clear skies there was no absorption and the effective noise temperature of the ground receiver was 100 degrees Kelvin. How much power must be used if there is 3 dB of loss caused by the rain?

Ans: First the loss factor La is 3 dB, which equals 10**3/10 = 2, The power is reduced by 1/2 coming through the rain. Second the rain makes the sky noisy. An effective temperature of 290 (1-1/La) = 290 x 1/2 = 145. This ads to the effective noise temperature Teff = 100 + 145 = 245 degrees Kelvin. The power has to be increased by an added factor of 245/100 = 2.45. Total increase is 2 x 2.45 = 4.9. The power must be **49 watts.**

D) We now switch from analog TV to digital. With 10 watts we needed and analog bandwidth of 20 MHz, and a carrier to noise ratio of 10 dB. With digital we need a 5 Mbit/sec data rate and Eb/No of 7 dB. How much power do we need for digital TV?

Ans The relation is the "C/KT" facto, which is proportional to transmit power required, equals B x (C/N) = R x Eb/No . With analog 20 x 10*6 x 10 = 200 x 10*6. With digital 5 x 10*6 x 5 = 25 x 10*6. (7 dB = 10 * 7/10 = 5 approximately). 200 /25 = 8.
10 watts/ 8 = **1.25 watts.** (I looked for some proficiency in logs, dBs and simple math approximations.)

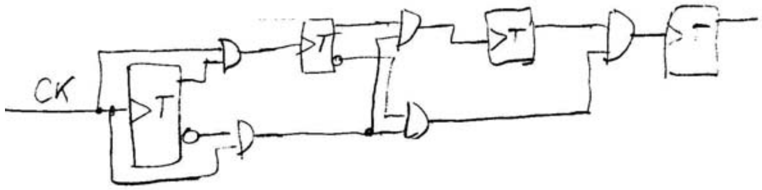
At 09:26 AM 1/22/02 -0800, you wrote:
>Reminder:
>Quals Meeting
>January 22, 2002
>Packard 101
>4:30 p.m.
>
>Please bring a copy of your Quals Question or you can email the question.
>Thanks,
>Diane Shankle
>Packard 165
>MC:9505
>(650) 723-3194
>Fax:(650) 723-1882
>

Professor Bruce B. Lusignan
Stanford University, Dept. of Electrical Engineering
Packard Electrical Engineering Bldg.
350 Serra Mall, #237
Stanford, CA 94305-9510

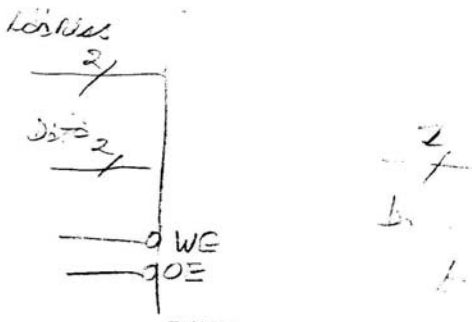
[illegible] Quiz 1-02

Question 1 Design a combinational circuit to multiply two binary numbers X and Y that are each 2 bits to get a binary product, $Z = X \cdot Y$.

Question 2
What does the following circuit do?



Question 3 An ASIC chip is to be designed using a library that does not include RAM elements. Please design a circuit that functions as an SRAM using only elements such as flip-flops, MUXes, gates, ...



Quals questions given by Professor Nick McKeown, January 2002.

Question #1:
I want to represent a counter with 8 different values using 3 incandescent light bulbs. The lifetime of the light bulbs is determined by how often they are switched on and off. The more often they are switched on or off, the more likely they are to fail.

a. If I want to maximize the time between replacing a light bulb, what code should I use to represent the 8 different values of the counter?

Question #2:
It's common for a room light to be switched on and off by two or more different light switches.

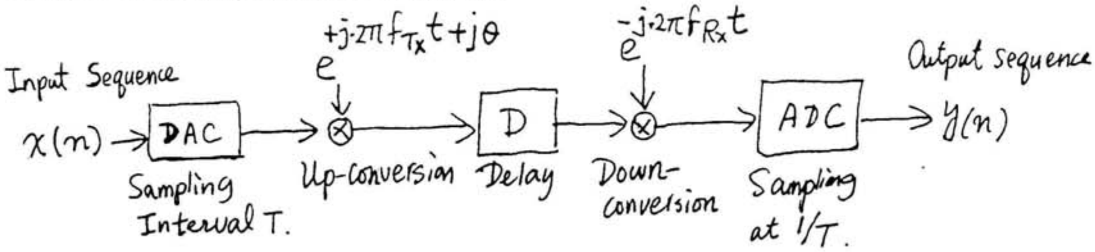
a. If two different switches control a light, how are the switches wired up?
b. What if there are three switches?

Question #3:
Network switches and routers process packets to decide where to send them, and to modify packet headers. Sometimes, a conventional CPU (such as a MIPs or Intel processor) is used to process the stream of packets.

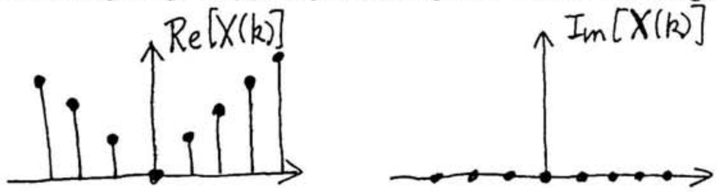
a. What do you think are the pros and cons of using a general-purpose processor for this application?
b. If you were to design a "packet-processor", how would it differ from a conventional processor?

Prof TERESA MENG

Consider the following communication system:



Assume that the input signal x(n) is a periodic sequence with the following DFT .



(1) Draw the DFT of the output sequence y(n) if $f_{Tx} = f_{Rx}$, $\theta = 0$, and $D = ½ * T$.

(2) Draw the DFT of the output sequence y(n) if $f_{Tx} = f_{Rx}$, $\theta = \pi/4$, and $D = 0$.

(3) Draw the DFT of the output sequence y(n) if $f_{Tx} = f_{Rx} + \frac{1}{2T}$, $\theta = 0$, and $D = 0$.

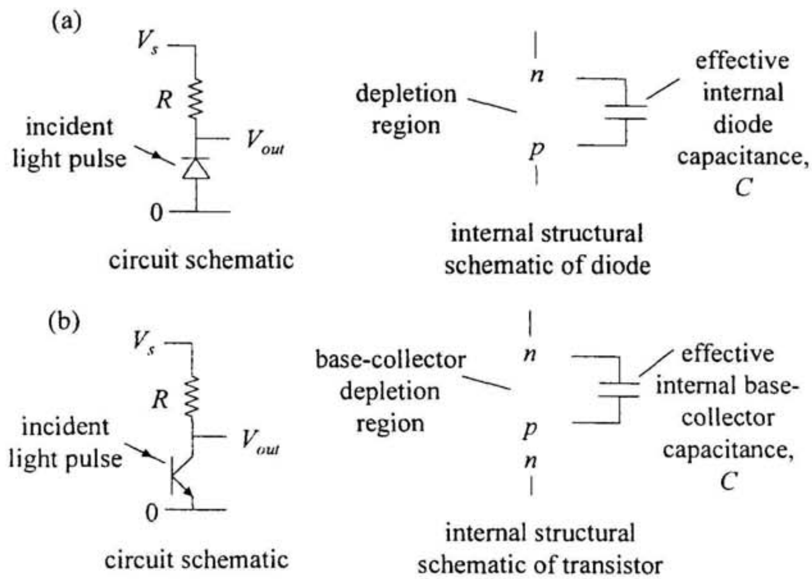
(4) Assume that there is no carrier frequency offset ($f_{Tx} = f_{Rx}$) and no carrier phase offset ($\theta = 0$), and that the DFT of the input sequence has only real components. If there is an unknown delay D, can you suggest a method in either the time or the frequency domain to detect it?

(5) Assume that there is no carrier phase offset ($\theta = 0$) and no delay ($D = 0$), and that the input sequence is real. If there is an unknown carrier frequency mismatch between f_{Tx} and f_{Rx}, can you suggest a method in either the time or the frequency domain to detect it?

EE Ph. D. Qualifying Exam 2002 - Prof. D. Miller

In one experiment, we shine a short pulse of light into the depletion region of a reverse-biased photodiode, as shown in the figure (a). In another experiment we shine a similar short pulse of light into the depletion region of a reverse-biased base-collector junction of a bipolar phototransistor as shown in the figure (b). In both cases, by the design of the structures, we ensure that the light is essentially completely absorbed in the depletion region.

In both cases, the diode structures have an internal capacitance, C, which we take to be similar in the two cases, and which is indicated by the capacitors in the figure. In both cases, the devices are biased through a similar resistor of value R to the positive supply, V_s. V_s is much larger than any of the forward voltages of any diode in the circuits, and it may be assumed that the photodiode and base-collector diode region remain reverse-biased, at least to some degree, at all times in the experiments. The base terminal of the transistor is not connected to any external circuit. Leakage currents may be neglected, though the transistor current gain is assumed to be finite. The emitter-base capacitance may be assumed to be much larger than C.



QUESTIONS

For each of the two experiments, sketch the behavior of the output voltage, V_{out}, as a function of time. Specifically, we want to know

What are the forms of the time behavior in the two experiments?

How does the maximum change in output voltage V_{out} compare between the two experiments?

How does the time dependence of the output voltage V_{out} compare between the two experiments?

(Note: if you finish these questions, supplementary questions will be asked)

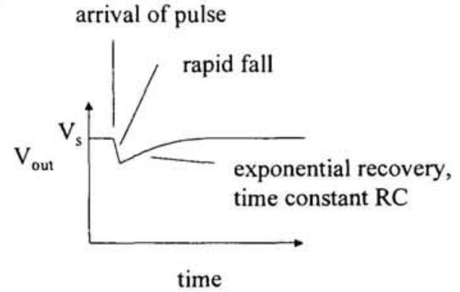
Supplementary question

For the case of the transistor, what do you think would be the simplest set of circuit elements (e.g., any combination of resistors, capacitors, transistors, power supplies) that could be substituted for the resistor R, and would lead to an output voltage response (not necessarily in the voltage V_{out}) that was both rapid and large?

Solution

In both cases, sending the short pulse of light into the diode generates electron-hole pairs. The pairs move rapidly within the diode (this actually takes only 10 – 100 ps), with the electrons moving to the n-region (collector of the transistor), and the holes moving to the p-region (base of the transistor). This movement rapidly discharges the diode capacitance by essentially the same amount in both the diode and transistor cases from its initial value of V_s (the transistor has no time to turn on anyway on such time scales, so it does not influence this initial drop).

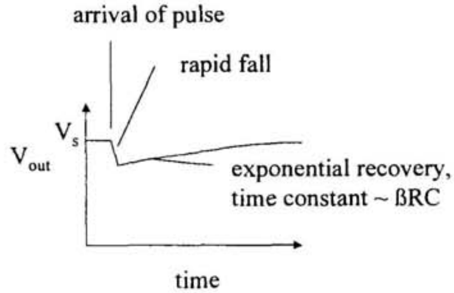
a)



In the case of the diode, the voltage charges back up through the resistor with a time constant RC.

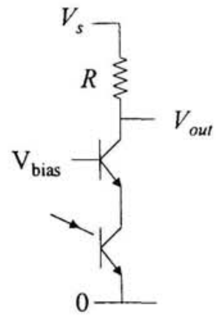
The case of the recovery of the voltage in the transistor is more subtle. As mentioned above, an effect of the pulse is to put holes on the transistor base. The only way the holes can get off of the base to lead to the eventual recovery of the whole system is for them to recombine with electrons inside the device because there is no external electrical connection to the base. In the normal operation of a transistor, such recombination takes place through a small fraction $\sim 1/\beta$ of the electron current into the emitter recombining with holes (where β is the transistor current gain). Therefore for each hole to be removed from the base, $\sim \beta$ electrons pass out of the collector. One might think that the collector current would simply get larger to allow this to happen rapidly, but it cannot. Any larger collector current would reduce the collector voltage because of the voltage drop over R, which would in turn reduce the base voltage by an equal amount (there is no way to take the charge off the base-collector capacitance other than removing holes, so lowering the collector voltage lowers the base voltage by an equal amount), turning off the transistor. Hence, the recovery of the phototransistor circuit is approximately β times slower (i.e., a time constant $\sim \beta RC$) because $\sim \beta$ times as much charge must be passed through the resistor at similar voltages. An answer that realized that the transistor case would be slower was sufficient for this part of the question.

b)



The above explanation of the transistor case is a mechanistic one based on thinking about where the electrons and holes go. An alternative and equivalent circuit-based explanation is to understand this longer time as being related to the Miller capacitance phenomenon well known in common emitter (and common source and vacuum triode) amplifier circuits in which the effective value of the base-collector capacitance is magnified by a factor of the transistor voltage gain. A circuit analysis of the phototransistor circuit can prove the above result, though it requires a charge-based rather than current-based approach.

The standard answer to the supplementary problem is to construct a cascode circuit, replacing the resistor with a common base amplifier. In "electron" terms, the existence of the common base amplifier allows large collector currents to flow without requiring the collector voltage to drop by large amounts. In circuit terms, such a circuit reduces the voltage gain in the first stage to -1, and transfers all the voltage gain to the common base amplifier circuit, eliminating the voltage gain that leads to the effective magnification of the base-collector capacitance.

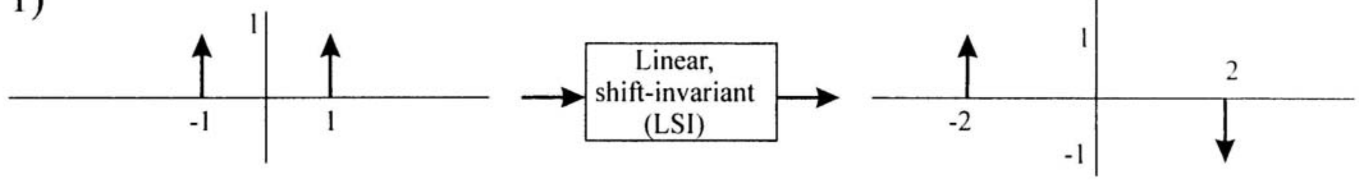


The average student on this problem got to about the point of understanding that the initial voltage drop in the transistor case was about the same as in the diode case. A few students managed to get all the way through to realizing that the transistor recovery would be slower. Many students thought (incorrectly) that the initial voltage fall after the short optical pulse would have an RC time constant (it has a time duration essentially independent of R, depending instead on how fast charge can move inside a reverse biased diode, a time of ~ 10 – 100 ps). One student got all the way through the supplementary problem with a creative answer. This problem is relatively straightforward if the student has an intuitive picture of how devices like transistors and photodiodes work, and looks at it in terms of where the electrons and holes are going. It is relatively hard if the student starts with analytic models, though one student got most of the way through with very adept circuit analysis.

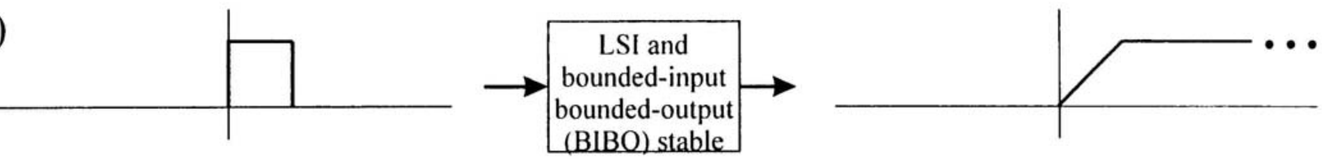
Quals Question : Nishimura ' 2002

For each system below, indicate if it is possible for the output function to occur, given the input function. (You may do them in any order you wish.)

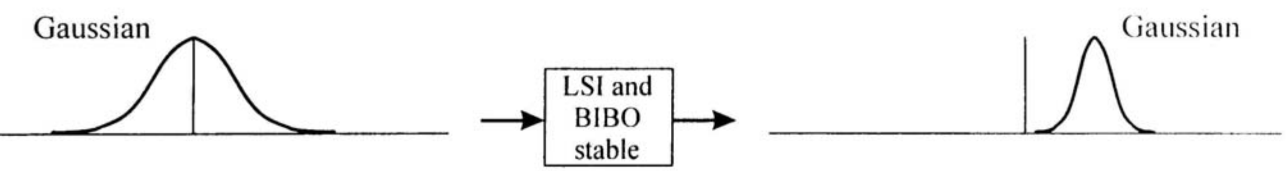
1)



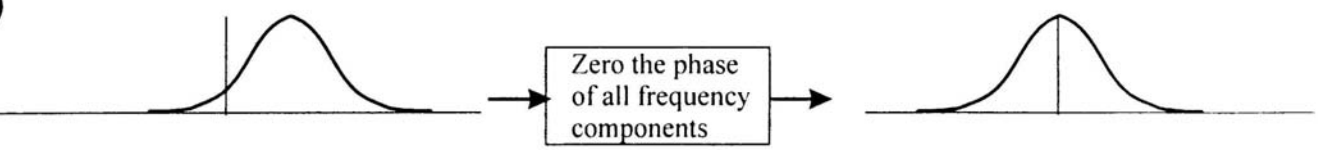
2)



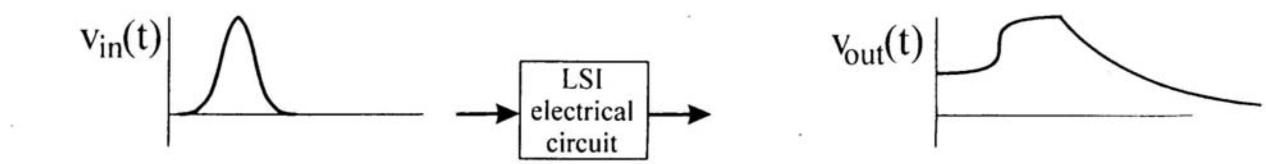
3)



4)



5)



Paulraj : Quals Questions 2002

Let x(t), [0,T] be a pulse of with some arbitrary waveform

1. In radar / sonar detection, what is a matched filter and why is it used

2. If x(t) is emitted and is returned after reflection by a target with some delay and Doppler shift, what does x(t) become

3. In presence of unknown delay and Doppler, what should receiver do to provide a matched filter for the arriving echo

4. Define the ambiguity function of a pulsed signal x(t)

5. Suppose

$$\Psi(\omega,\tau) = \frac{1}{T}\int_{-T}^{T} x(t)x(t-\tau)\exp^{j\omega t}\, dt$$

6. Let x(t) be a linear FM pulse whose frequency sweeps from $-\omega_0$ to $+\omega_0$ during the pulse period.

   Sketch $|\Psi(\omega,\tau)|$ of x(t) on ω,τ plane

7. Can you state some properties of $\Psi(.,.)$

8. What is the significance of $\Psi(.,.)$ with respect to radar receiver performance

From: "Fabian Pease" <pease@cis.stanford.edu>
To: "Diane Shankle" <shankle@ee.Stanford.EDU>
Cc: <group@jumpjibe.stanford.edu>
Subject: Quals question
Date: Tue, 22 Jan 2002 10:33:02 -0800
X-Priority: 3

Pease's Question(s):

1. Draw the circuit diagram of a CMOS inverter. Add a capacitive load C.
2. If you apply a square-wave clock to the input how much energy is dissipated in 1 clock cycle (and where and when) ?
3. Do you think chips will draw more or less power in the future ? Explain.
4. Why have batteries progressed so little in the last 100 years?

fp

Date: Sun, 27 Jan 2002 10:05:30 -0800
From: "Pianetta, Piero" <pianetta@SLAC.Stanford.EDU>
Subject: quals question
To: Diane Shankle <shankle@ee.Stanford.EDU>

*What determines the mobility of a semiconductor and explain the temperature dependence?

*Given three different materials-Si, Ge and GaAs-can you explain why their mobilities are different?

*What determines the mobility of polycrystalline silicon versus single crystal Si?

*Given a semiconductor with three contiguous regions of different mobility, how would you calculate the overall mobility of the material?

****

Piero Pianetta

Stanford Synchrotron Radiation Laboratory

2575 Sand Hill Road

Menlo Park, CA 94025

Phone: 650-926-3484

Fax: 650-926-4100

Balaji Prabhakar

Stanford University
Department of Electrical Engineering

EE Quals. Jan 14-18, 2002.

The problem is about crowd control – say at a web site. Requests arrive at a web site of limited capacity. The site employs the following admission control procedure.

Each arriving request is held for H units of time before possible admission. If, when a request is held, another arrives, then the first request is dropped (not admitted) and the second will be held for H units of time. And, so on...

1. Characterize the requests which are admitted.

2. If requests arrive according to a Poisson process of rate λ, what is the rate of admitted requests?

Hint: The probability that there are k requests in an interval of length T equals $e^{-\lambda T}\frac{(\lambda T)^k}{k!}$, for $k = 0, 1, 2, \ldots$.

3. Now suppose that the holding time H is random, distributed as an exponential random variable of mean 1. Also suppose that H is independent of the arrivals. What is the rate of admitted requests?

X-Sender: quate@quate.pobox.Stanford.EDU (Unverified)
Date: Tue, 22 Jan 2002 10:47:10 -0800
To: Diane Shankle <shankle@ee.Stanford.EDU>
From: cal quate <quate@Stanford.EDU>
Subject: Re: Quals Meeting Today

Quals Questions from quate

1 - Draw the magnetic field lines (both H & B) for a simple solenoid of finite length; a) filled with air, b) filled with a paramagnetic material with a relatively permeability of 3.

2- In a coaxial line the space between inner and outer conductor is filled with a lossy dielectric with a conductivity, sigma. Please calculate the resistance from the inner conductor to the outer conductor.

2 - What is the physical origin of resistance of a copper wire at room temperature?

cal quate
==============================
At 09:26 AM 1/22/2002 -0800, you wrote:

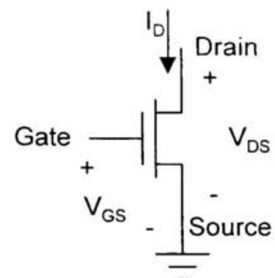
> Please bring a copy of your Quals Question or you can email the question.
> Thanks,
> Diane Shankle
> Packard 165
> MC:9505
> (650) 723-3194
> Fax:(650) 723-1882

- A conventional n-channel silicon-based MOSFET is shown to the right.

  a) Sketch typical I_D v.s. V_{DS} characteristics for a variety of V_{GS} values. Assume a threshold voltage (V_T) of 1 volt.

  b) Indicate triode, saturation and cutoff regions.

  c) Write an expression for I_D that is valid for the triode region. The expression should be in terms of: V_{DS} , V_{GS} , V_T , channel width (W), channel length (L), mobility (μ_n) and oxide capacitance (C_{ox}).



n-channel MOSFET

2) In saturation, $I_{DSAT} = (W\ /\ 2L)\ \mu_n\ C_{ox}\ (V_{GS} - V_T)^2$

   a) Write an expression for the small-signal model MOSFET transconductance (g_m).

   b) Does the transconductance depend linearly on I_D?

- Your boss now asks you to design a couple of simple digital circuits.

  a) First design a digital inverter. She says that the input voltage should be applied to V_{GS} , that the output voltage should be taken from V_{DS} , and that a resistor should be placed between the drain and the power supply (V_{DD}). You are concerned about having sufficient noise margins. Do you recommend a large or a small resistor value? Why?

  b) Your boss now mentions that the inverter will be driving a capacitive load (C_L) which should be placed between the drain and ground. You are now concerned about the transient response (switching speed). Should you replace the resistor with an active load (e.g., a p-channel MOSFET)? Why?

  c) You now remember that static power consumption is extremely important for all digital circuits that your company designs. What inverter could you recommend to minimize static power burn? Why?

  d) The second digital circuit that you are asked to design is a 2-input NOR circuit. The circuit should use no more than two (2) n-channel MOSFETs and two (2) p-channel MOSFETs and should not have a DC path between power supply and ground. Clearly indicate where the inputs ("A" and "B") are applied and where the output ("C") appears.

2002 Quals Questions with short answers

Olav Solgaard

1 *Sketch the complete electrostatic field of a parallel-plate capacitor with a voltage, V.*

The sketch should include the almost uniform field between the capacitor plates, and the weaker fringing fields.

2 *What is the definition of the electric field?*

The electric field equals the force field on a test charge divided by the magnitude of the test charge.

3 *Derive an approximate expression for the electric field between the capacitor plates. State your assumptions.*

$E = \frac{V}{d}$ where V is the applied voltage, and d is the separation of the plates. Assumptions: The field is uniform between the plates and zero elsewhere.

4 *Given the same assumptions, derive an expression for the magnitude of the capacitance?*

$C = \frac{\varepsilon ?\varepsilon_0 ?A}{d}$ where $\varepsilon \cdot \varepsilon_0$ is the dielectric constant of the medium between the plates, A is the area of the plates, and d is again the separation of the plates.

5 *How can you tell by inspection that the approximate field you found is not an exact solution to Maxwell's equations?*

Faraday's law for electrostatic fields $\left(\leftarrow \vec{E} = 0 \right)$ is not fulfilled at the transition between the uniform field between the plates and the zero field outside the plates.

6 *What is the force on each capacitor plate? Again use the approximate expression for the electric field.*

$$F = \frac{\varepsilon ?\varepsilon_0 ?A ?V^2}{2d^2}$$

7 *What is the energy storage in the capacitor?*

$$W = \frac{1}{2} ?C ?V^2$$

8 *Suppose you start with a parallel-plate capacitor with no stored charge, and proceed to charge the capacitor until the voltage is V. How much energy is dissipated during the charging process?*

$$W = \frac{1}{2} ?C ?V^2$$

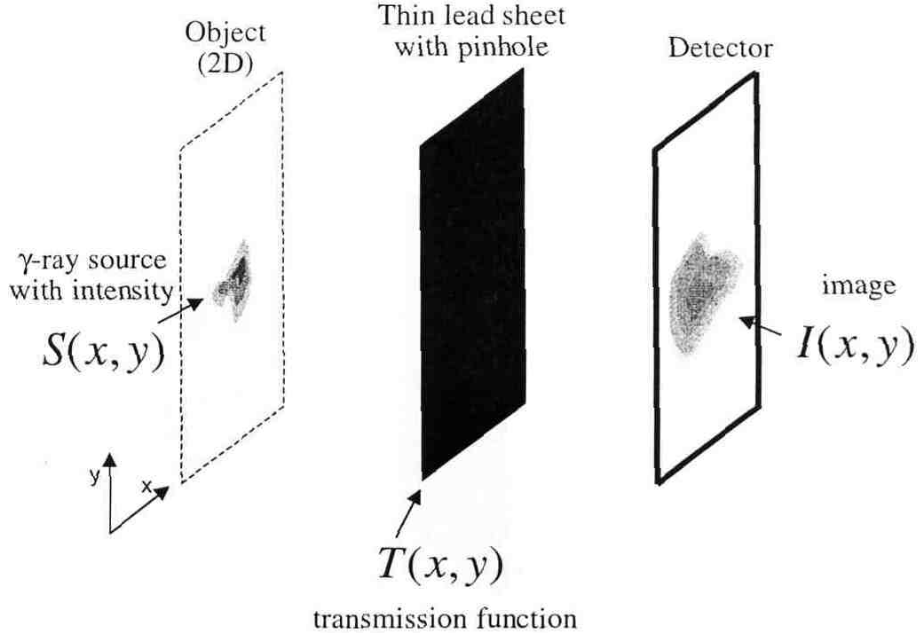
9 *What does this imply for the noise processes that are important in capacitive sensors?*

Thermal noise will limit the accuracy of capacitive sensors even though ideal capacitors are non-dissipative.

10 *Is the total charge on a capacitor a multiple of the electron charge?*

The charge is quantized, but the distribution of the charge, and therefore the energy storage, is not.

Nuclear Medicine Pinhole Camera

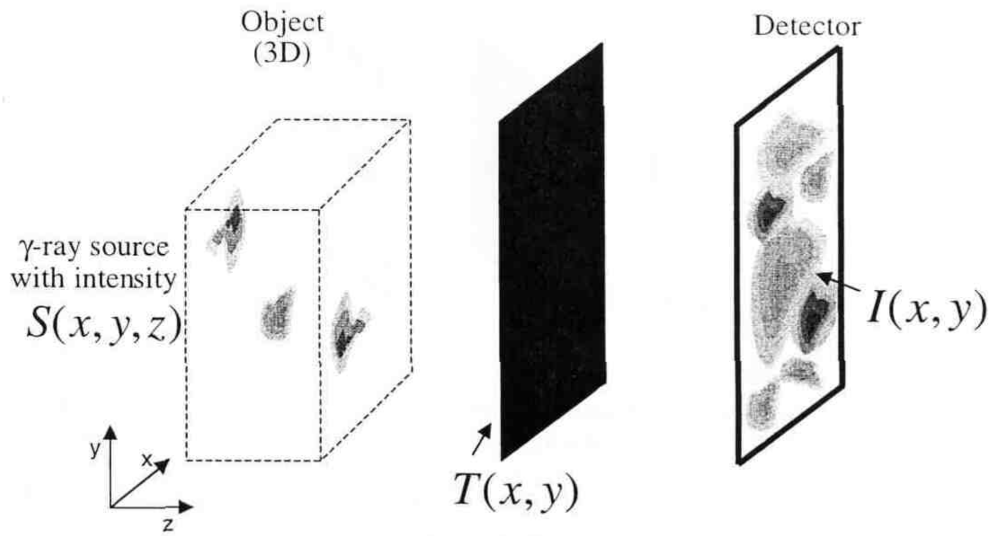


1. a) Can this camera be analyzed as a linear space-invariant system?

   b) What is the impulse response?

   c) How does one choose the size of the pinhole?

Consider the general 3D case:

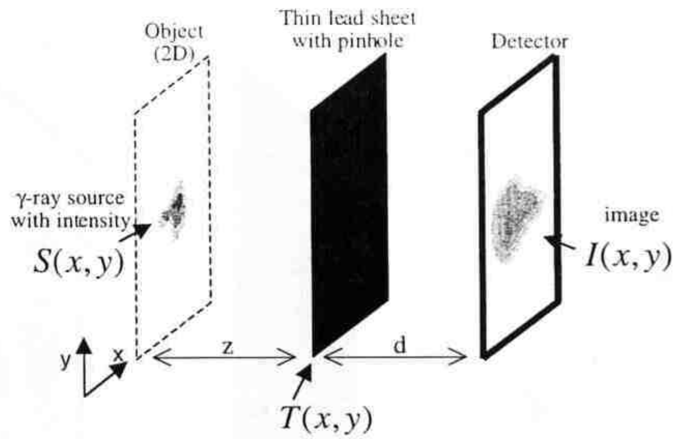


with image reconstruction algorithm:

$$\tilde{I}(x,y) = I(x,y) * T(-x,-y)$$

3. How does the camera behave for a 3-dimensional γ-ray source? Is there a revised reconstruction algorithm that yields resolution in z?

Solutions:



1. a) Yes, the pinhole camera can be interpreted as a linear and space-invariant system. Strictly speaking, the output image is an inverted, filtered, magnified version of the source distribution (easily derived from ray tracing and the definition of the convolution integral). Ignoring obliquity (which causes a decrease in image intensity as one moves away from the center of the detector),

$$I(x,y) \propto S(x/M, y/M) * T(x/m, y/m)$$

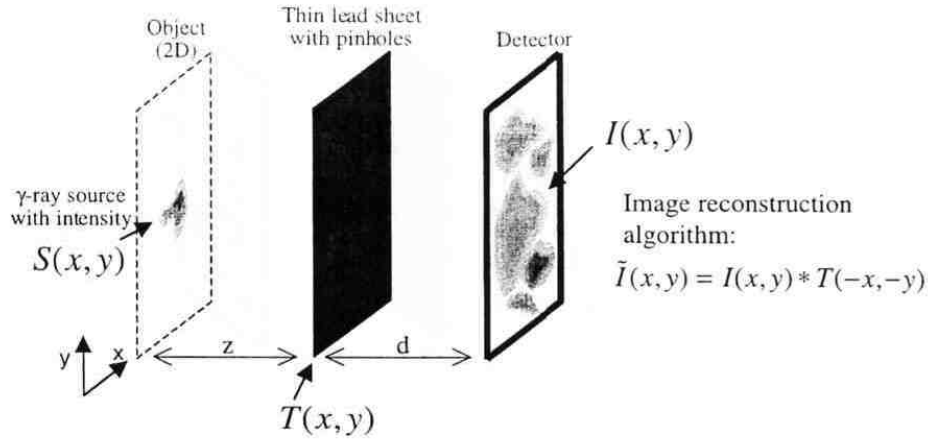
where * denotes 2-d convolution and M and m are magnification factors given by $-z/d$ and $(z+d)/z$ respectively. The key point of this question was the derivation of the convolution relationship. Thus is was ok if students ignored the image flipping and magnification terms (most people did).

b) The impulse response is simply $T(x/m, y/m)$.

c) The size of the pinhole is a tradeoff between resolution and SNR. The smaller the hole, the sharper the image, but the fewer photons get through to the detector (typical nuclear medicine pinhole systems have a capture efficiency of around 10^{-4}).

Bonus: SNR is proportional to the square root of the number of photons used to make the image (Poisson statistics).

Solutions (cont.):



2. a) Yes, this system is known as coded aperture imaging. Using the results from 1.a),

$$\tilde{I}(x,y) \propto S(x/M, y/M) * T(x/m, y/m) * T(-x/m, -y/m)$$

$$\tilde{I}(x,y) \propto S(x/M, y/M) * \left(T(x/m, y/m) \star T(x/m, y/m)\right)$$

where $\star$ denotes the 2d cross-correlation. Thus, the system impulse response is given by the autocorrelation of *T(x/m,y/m).* As with Question 1, it was ok for students to ignore the magnification factors.

b) We want to choose *T(x,y)* such that its autocorrelation is a good approximation to a delta function, i.e. having a single sharp peak at the origin. Choosing the pinholes, for example, randomly distributed or samples of a 2d chirp function ($\sin(\omega x^2)\ \sin(\omega y^2)$) will work well. Periodic pinhole patterns are a poor choice. Also, spreading the pinholes out far enough to prevent image overlap is generally impractical given the finite size of the detector.

Bonus: even the best choice of T(x,y) will result in an autocorrelation function that consists of a sharp central peak of height n (given n pinholes) in addition to a broad baseline of height 1 (corresponding to the overlap of a single pinhole in the autocorrelation). This results in the desired image plus a broad low spatial frequency component. This broad component will also degrade the noise performance. Thus, in practice, coded aperture systems only have better SNR than a pinhole camera for a small or sparse source distribution. However, one advantage, as will be shown in Question 3, is that z resolution is achievable independent of the size and shape of the source.

Solutions (cont.):

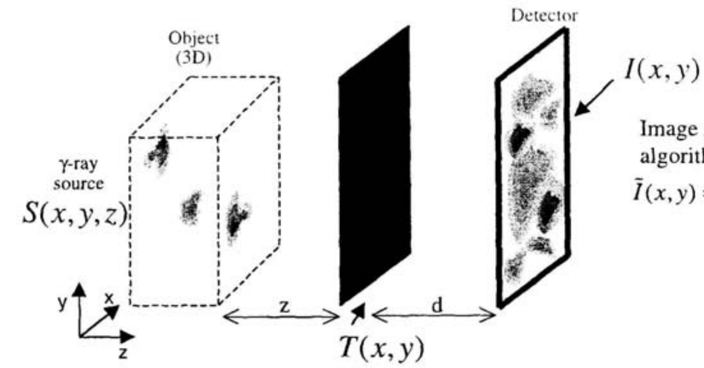


Image reconstruction algorithm:

$\tilde{I}(x,y) = I(x,y) * T(-x,-y)$

3. For the 3d case, we *cannot* ignore the magnification factors. Use the revised reconstruction algorithm:

$$\tilde{I}(x,y,z_0) \propto I(x,y) * T\big(-x/m(z_0),-y/m(z_0)\big)$$

where

$$I(x,y) = \int_z S\big(x/M(z),y/M(z)\big) * T\big(x/m(z),y/m(z)\big)dz.$$

Resolution in z can be evaluated by plotting

$$T\big(x/m(z),y/m(z)\big) * T\big(-x/m(z_0),-y/m(z_0)\big)$$

versus z.

Images from this system will be a sum of the desired plane, z_0, in focus with all other planes blurred out.

X-Authentication-Warning: sidon.stanford.edu: tobagi owned process doing -bs
Date: Mon, 25 Feb 2002 19:28:40 -0800 (PST)
From: "Fouad A. Tobagi" <tobagi@Stanford.EDU>
To: Diane Shankle <shankle@ee.Stanford.EDU>
Subject: Re: Quals Question 2002

Consider a video clip of 30 frames per second and of a given duration, say T seconds. It is compressed using variable bit rate encoding; that is, the number of bits per frame varies from frame to frame. We let bi denote the number of bits in frame i.

The compressed vide file is stored on a video server from which it is streamed at constant bit rate c to a client. The client has a buffer of size B. Determine the possible values of c and B that allows a smooth playback of the video at the client.

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Fouad Tobagi

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Professor of Electrical Engineering — Fax: (650) 725-6221
and by courtesy, Computer Science — Email: Tobagi@stanford.edu

Gates Building, Room 3A-339, Stanford University, Stanford, CA 94305-9030
==

To: shankle@ee.Stanford.EDU
Subject: EE quals question/solution
Date: Fri, 25 Jan 2002 17:54:18 -0800
From: Jennifer Widom <widom@DB.Stanford.EDU>

Diane,

I suddenly realized I forgot to turn in my EE Quals question/ solution. I can't remember if you're the one to take it, but if not please forward accordingly.

Thanks,
Jennifer Widom

===========================
2002 EE Quals, Prof. Jennifer Widom

Problem
=======

Consider directed graphs with a single source (root) node R from which there is at least one path to every other node N in the graph.

We can represent such graphs in a couple of ways:

(1) As a data structure of nodes and edges. Each node N has i >= 0 out-neighbors (children) accessed as N.1, N.2, ..., N.i
[show example]

(2) As a K x K square matrix M for a graph with K nodes. M[x,y] = 1 if there is an edge from node x to node y; M[x,y] = 0 otherwise
[show same example]

Write a program that determines whether such a graph contains a cycle. The program should return YES if there is one or more cycles, NO otherwise.

* You may use whichever of the two graph representations you prefer.

* You may use any pseudocode notation you like, including function definitions and calls if it helps you.

* Your solution will be graded on simplicity as well as correctness.

Solution
========
Here are two possible solutions but there are many correct variants.

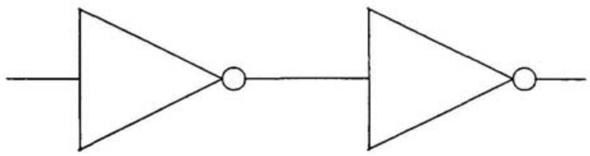
Using representation (1):

```
  main program: cycle(R, {R})

  function cycle(N, seen-set)
    if N has no children return(NO)
    else if any of N.1, N.2, ..., N.i are in seen-set return(YES)
    else if cycle(N.i, seen-set U {N.i})=YES for any i
      then return(YES) else return(NO)
```

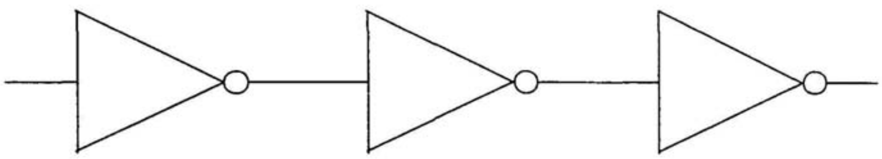
Using representation (2):

```
  repeat until M is unchanged:
    M <- M + M*M
  if M[x,x] > 0 for any x then return(YES) else return(NO)
```



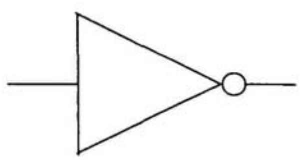
For the circuit shown above, if the output is connected to the input, how will the circuit behave ?

Ans. Latch



For the circuit shown above, if the output is connected to the input, how will the circuit behave ?

Ans. Ring Oscillator



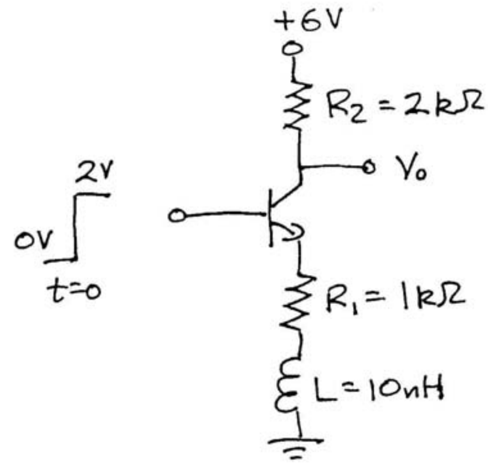
For the single stage inverter shown above, will it oscillate if the output is connected to the input ?

Ans. No. A typical inverter (e.g., CMOS inverter) has one dominant pole. The maximum phase shift around the loop will be 270° (180 ° for the inversion and 90 ° from the pole) < 360 °, not sufficient to cause instability. The three-stage circuit has three poles and hence there will be sufficient phase shift around the loop to cause instability.

Bonus : What components (except transistor) will you add to make the single stage inverter oscillate ?

BRUCE WOOLEY

EE QUALS 01-02



What is the response at V_o for $t > 0$?

EE Qualifying Examination January 2002

Yoshihisa Yamamoto

1. A linear amplifier amplifies the signal and noise powers by the same gain factor. In addition, amplifier internal noise is added to the output signal and so the signal-to-noise ratio is always decreased by linear amplification. Then, why is a linear amplifier used in various communication systems and computing systems.

2. A nonlinear regenerator reproduces a clean signal pulse from a distorted signal pulse. The signal-to-noise ratio of the output pulse is larger than that of the input pulse, which is the ultimate origin for the superiority of a digital system over an analog system. What is the cost we have to pay in improving the signal-to-noise ratio in a nonlinear regenerator?

3. Discuss the above two points in the context of optical communication systems with fiber amplifier repeaters.