

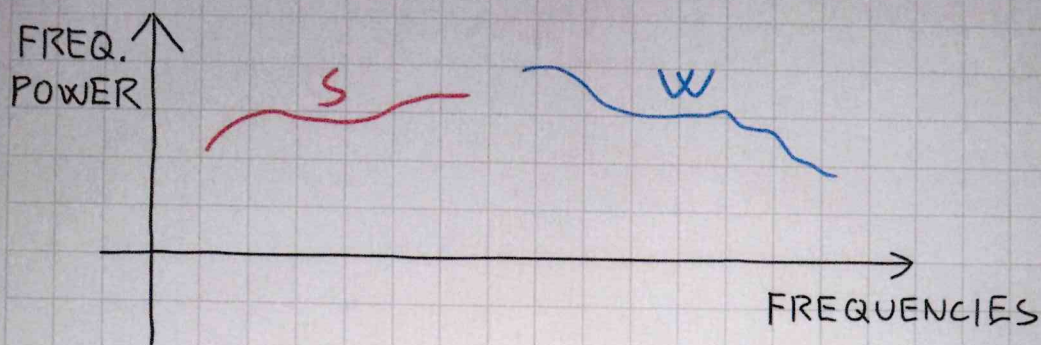
ESTIMATION AND FILTERING: A BRIEF INTRODUCTION

SIGNAL: A FUNCTION THAT DEPENDS ON TIME t , $s(t)$. IT CAN CARRY CRUCIAL INFORMATION FOR MODELING AND CONTROL PURPOSES.

PROBLEM: THEY CAN BE NON DIRECTLY MEASURABLE AND/OR CORRUPTED BY NOISE. CONSIDER THIS LAST CASE.

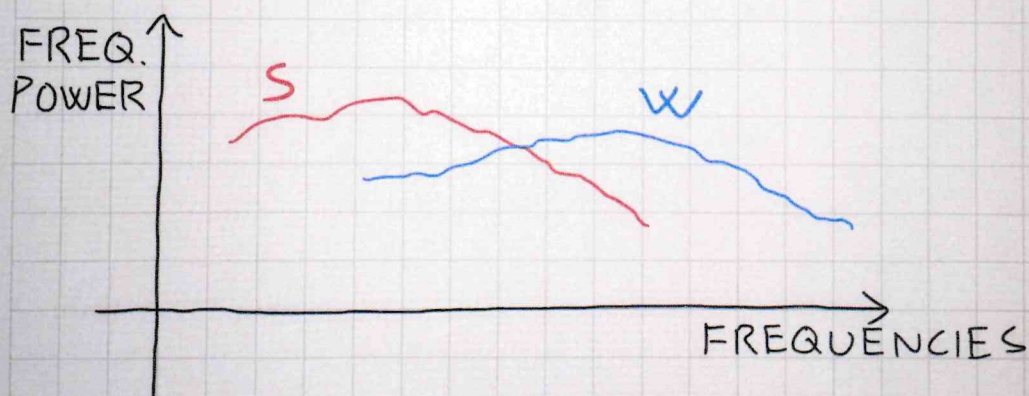
$$\begin{pmatrix} \text{MEASUREMENTS} \\ y \end{pmatrix} = \begin{pmatrix} \text{SIGNAL} \\ s \end{pmatrix} + \begin{pmatrix} \text{NOISE} \\ w \end{pmatrix}$$

ONE SIMPLE SCENARIO IS WHEN s AND w ARE KNOWN TO "LIVE" OVER DIFFERENT FREQUENCIES



A SIMPLE LOW-PASS FILTER CAN BE USED TO RECOVER S FROM y

BUT A MUCH MORE REALISTIC SCENARIO INVOLVES OVERLAP



HOW TO DIVIDE S FROM w ?

WIENER (1940)

- FROM DETERMINISTIC TO STATISTICAL SIGNAL PROCESSING

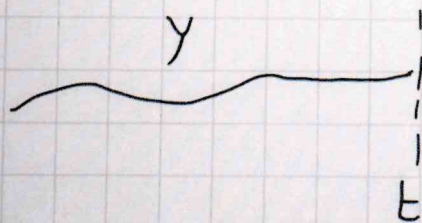
WIENER (1940)

FROM DETERMINISTIC TO STATISTICAL
SIGNAL PROCESSING

S, W SEEN AS STOCHASTIC

PROCESSES \rightarrow COLLECTION OF
RANDOM VARIABLES
INDEXED BY TIME

- STATISTICAL PROPERTIES OF S, W
ARE EMBEDDED IN THEIR PROBABILITY
DENSITY FUNCTIONS
- RECOVERING S FROM W BECOMES
A STATISTICAL ESTIMATION PROBLEM
- WIENER THEORY \rightarrow ALLOWS ON-LINE
STATISTICAL
SIGNAL
PROCESSING



ESTIMATE $S(t)$ USING ONLY THE

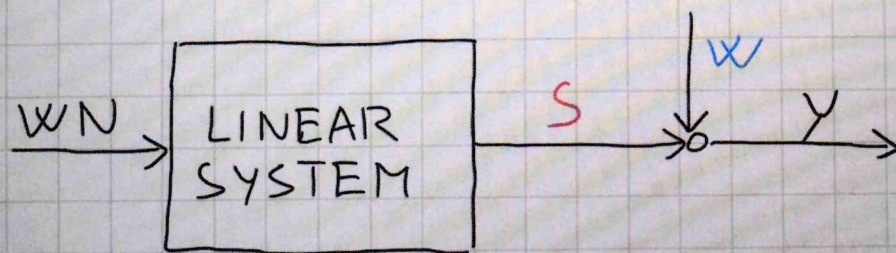
$\{y(\tau)\}_{\tau \leq t}$ (CAUSAL ESTIMATOR)

AND UPDATE EFFICIENTLY THE

ESTIMATE $s(t)$ USING ONLY THE
 $\{y(\tau)\}_{\tau \leq t}$ (CAUSAL ESTIMATOR)
AND UPDATE EFFICIENTLY THE
ESTIMATE AS t GOES ON AND
NEW DATA ARRIVE. THIS IS
THE FILTERING PROBLEM

ONE FEATURE AND
ONE KEY LIMITATION
OF WIENER FILTER

- SIGNALS SEEN AS OUTPUTS OF
LINEAR SYSTEMS DRIVEN BY
WHITE NOISES. INPUT-OUTPUT
VIEW



- ALL THE SIGNALS NEED TO BE

- ALL THE SIGNALS NEED TO BE ASSUMED STATIONARY STOCHASTIC PROCESSES, IMPLYING E.G.

$$E s(t) = \mu \quad \forall t$$

$$\text{VAR } s(t) = \sigma_s^2 \quad \forall t$$

$$\text{VAR } w(t) = \sigma^2 \quad \forall t$$

FROM WIENER TO

KALMAN (1959):

FROM INPUT-OUTPUT TO

STATE-SPACE MODELS

WE STILL USE LINEAR SYSTEMS BUT
IN STATE-SPACE FORM

$$x(t+1) = A x(t) + B v(t) + B_w u(t)$$

$$y(t) = C x(t) + w(t)$$

v, w ARE NOISES (STOCHASTIC PROCESSES)

u CAN BE A DETERMINISTIC INPUT

x, y ARE THUS STOCHASTIC PROCESSES

WE CAN MEASURE THE OUTPUT y BUT NOT THE STATE x WHICH CONTAINS ALL THE INFORMATION ON THE SYSTEM. THE SIGNAL OF INTEREST s CAN BE SEEN AS "CONTAINED" IN x (ONE OR MORE COMPONENTS OF x , OR LINEAR COMBINATIONS OF THE STATES).

EXAMPLE

WE WANT TO MODEL A VEHICLE GOING IN A STRAIGHT LINE.

$$x = \begin{bmatrix} x_1 \\ x_2 \end{bmatrix} = \begin{bmatrix} \text{POSITION} \\ \text{VELOCITY} \end{bmatrix}$$

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u = COMMANDED ACCELERATION

y = MEASURED POSITION

SAY WE CAN CHANGE THE ACCEL.
AND MEASURE THE POSITION EVERY
 T SECONDS

ELEMENTARY PHYSICS

LAWS SAY

$$x_2(t+1) = x_2(t) + T u(t) \quad \text{VELOCITY EVOLUTION}$$

BUT THIS IS NOT REALISTIC!

VELOCITY WILL BE PERTURBED

BY NOISE DUE TO WIND, POT-HOLES,
OTHER UNFORTUNATE REALITIES.

THE $x_2(t)$ IS BETTER DESCRIBED

AS A RANDOM VARIABLE SUBJECT

TO TRANSITION NOISE v :

$$x_2(t+1) = x_2(t) + T u(t) + v(t)$$

TO TRANSITION NOISE v :

$$x_2(t+1) = x_2(t) + T u(t) + v_2(t)$$

SIMILARLY, FOR THE POSITION

$$x_1(t+1) = x_1(t) + T x_2(t) + \frac{1}{2} T^2 u(t) + v_1(t)$$

SINCE THE MEASURED OUTPUT IS THE POSITION CORRUPTED BY SOME MEASUREMENT NOISE $w(t)$,

ONE HAS

$$x(t+1) = \begin{bmatrix} 1 & T \\ 0 & 1 \end{bmatrix} x(t) + \begin{bmatrix} T^2 \\ T \\ 0 \end{bmatrix} u(t) + v(t)$$

$$y(t) = [1 \ 0] x(t) + w(t)$$

IF WE WANT TO CONTROL

THE VEHICLE WITH SOME

SORT OF FEEDBACK, WE

NEED ESTIMATES OF x_1 AND x_2 ,

A STATE ESTIMATOR!

THIS IS WHERE THE KALMAN

$$x(t+1) = \begin{bmatrix} 1 & T \\ 0 & 1 \end{bmatrix} x(t) + \begin{bmatrix} T^2 \\ T \end{bmatrix} u(t) + v(t)$$

$$y(t) = [1 \ 0] x(t) + w(t)$$

IF WE WANT TO CONTROL THE VEHICLE WITH SOME SORT OF FEEDBACK, WE NEED ESTIMATES OF x_1 AND x_2 , A STATE ESTIMATOR!

THIS IS WHERE THE KALMAN FILTER COMES IN: DEEP CONSEQUENCES DUE TO THE STATE-SPACE VIEWPOINT.

LET US HAVE A LOOK AT THE INTRO OF THE ORIGINAL KALMAN PAPER

A New Approach to Linear Filtering and Prediction Problems¹

R. E. KALMAN

Research Institute for Advanced Study,²
Baltimore, Md.

The classical filtering and prediction problem is re-examined using the Bode-Shannon representation of random processes and the "state transition" method of analysis of dynamic systems. New results are:

(1) *The formulation and methods of solution of the problem apply without modification to stationary and nonstationary statistics and to growing-memory and infinite-memory filters.*

(2) *A nonlinear difference (or differential) equation is derived for the covariance matrix of the optimal estimation error. From the solution of this equation the coefficients of the difference (or differential) equation of the optimal linear filter are obtained without further calculations.*

(3) *The filtering problem is shown to be the dual of the noise-free regulator problem. The new method developed here is applied to two well-known problems, confirming and extending earlier results.*

The discussion is largely self-contained and proceeds from first principles; basic concepts of the theory of random processes are reviewed in the Appendix.

Introduction

AN IMPORTANT class of theoretical and practical problems in communication and control is of a statistical nature. Such problems are: (i) Prediction of random signals; (ii) separation of random signals from random noise; (iii) detection of signals of known form (pulses, sinusoids) in the presence of random noise.

In his pioneering work, Wiener [1]³ showed that problems (i) and (ii) lead to the so-called Wiener-Hopf integral equation; he also gave a method (spectral factorization) for the solution of this integral equation in the practically important special case of stationary statistics and rational spectra.

Many extensions and generalizations followed Wiener's basic work. Zadeh and Ragazzini solved the finite-memory case [2]. Concurrently and independently of Bode and Shannon [3], they also gave a simplified method [2] of solution. Booton discussed the nonstationary Wiener-Hopf equation [4]. These results are now in standard texts [5-6]. A somewhat different approach along these main lines has been given recently by Darlington [7]. For extensions to sampled signals, see, e.g., Franklin [8], Lees [9]. Another approach based on the eigenfunctions of the Wiener-Hopf equation (which applies also to nonstationary problems whereas the preceding methods in general don't), has been pioneered by Davis [10] and applied by many others, e.g., Shinbrot [11], Blum [12], Pugachev [13], Solodovnikov [14].

In all these works, the objective is to obtain the specification of a linear dynamic system (Wiener filter) which accomplishes the prediction, separation, or detection of a random signal.⁴

Present methods for solving the Wiener problem are subject to a number of limitations which seriously curtail their practical usefulness:

(1) The optimal filter is specified by its impulse response. It is not a simple task to synthesize the filter from such data.

(2) Numerical determination of the optimal impulse response is often quite involved and poorly suited to machine computation. The situation gets rapidly worse with increasing complexity of the problem.

(3) Important generalizations (e.g., growing-memory filters, nonstationary prediction) require new derivations, frequently of considerable difficulty to the nonspecialist.

(4) The mathematics of the derivations are not transparent. Fundamental assumptions and their consequences tend to be obscured.

This paper introduces a new look at this whole assemblage of problems, sidestepping the difficulties just mentioned. The following are the highlights of the paper:

(5) *Optimal Estimates and Orthogonal Projections.* The Wiener problem is approached from the point of view of conditional distributions and expectations. In this way, basic facts of the Wiener theory are quickly obtained; the scope of the results and the fundamental assumptions appear clearly. It is seen that all statistical calculations and results are based on first and second order averages; no other statistical data are needed. Thus difficulty (4) is eliminated. This method is well known in probability theory (see pp. 75-78 and 148-155 of Doob [15] and pp. 455-464 of Loève [16]) but has not yet been used extensively in engineering.

(6) *Models for Random Processes.* Following, in particular, Bode and Shannon [3], arbitrary random signals are represented (up to second order average statistical properties) as the output of a linear dynamic system excited by independent or uncorrelated random signals ("white noise"). This is a standard trick in the engineering applications of the Wiener theory [2-7]. The approach taken here differs from the conventional one only in the way in which linear dynamic systems are described. We shall emphasize the concepts of *state* and *state transition*; in other words, linear systems will be specified by systems of first-order difference (or differential) equations. This point of view is

¹ This research was supported in part by the U. S. Air Force Office of Scientific Research under Contract AF 49 (638)-382.

² 7212 Bellona Ave.

³ Numbers in brackets designate References at end of paper.

⁴ Of course, in general these tasks may be done better by nonlinear filters. At present, however, little or nothing is known about how to obtain (both theoretically and practically) these nonlinear filters.

Contributed by the Instruments and Regulators Division and presented at the Instruments and Regulators Conference, March 29-April 2, 1959, of THE AMERICAN SOCIETY OF MECHANICAL ENGINEERS.

NOTE: Statements and opinions advanced in papers are to be understood as individual expressions of their authors and not those of the Society. Manuscript received at ASME Headquarters, February 24, 1959. Paper No. 59-1RD-11.

AND NONLINEAR
SYSTEMS?

LAST PART OF THE
COURSE

BASED ON STOCHASTIC
SIMULATION TECHNIQUES
RECENTLY DEVELOPED

COURSE OUTLINE

- Overview of probability theory
- Static Bayesian estimation
- On-line estimation: Wiener filter
- On-line estimation: Kalman filter
- Static Bayesian estimation using rejection sampling and MCMC
- Nonlinear filtering and prediction: particle filters

COURSE OUTLINE

- Overview of probability theory
- Static Bayesian estimation
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- Static Bayesian estimation using rejection sampling and MCMC
- Nonlinear filtering and prediction: particle filters

EXAMINATION (two hours)

- Exercise on Static Bayesian Estimation (SBE) and on Kalman Filter (KF): 4 parts, **20 points** overall, minimum **11 points** are required
- Theoretical questions on SBE and KF (convergence theorem), MCMC and particle filters: 2 parts, **11 points** overall, minimum **7 points** are required
- It is mandatory for any student to register in the examination list within the deadline (typically one week before the examination date), no exception to this rule



Appelli d'esame » Visualizza appelli

Visualizza appelli

Appelli di: ESTIMATION AND FILTERING [INQ0091318]

[visualizza dettagli >>](#)

CONTROL SYSTEMS ENGINEERING [IN2546] (LM)...

Elenco appelli d'esame

Nuova prova finale

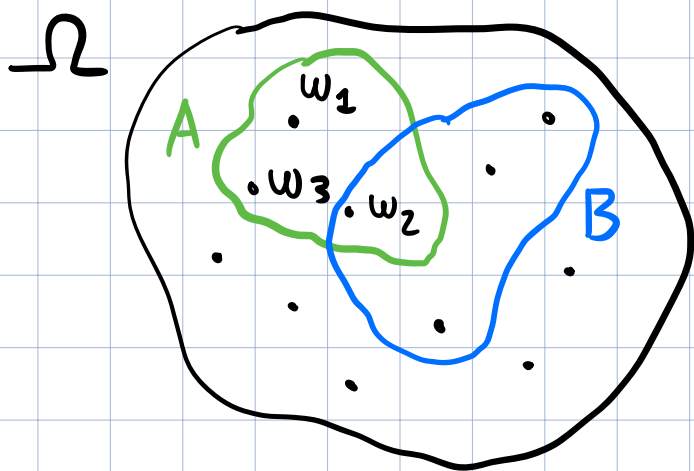
Nuova prova parziale

Visualizza

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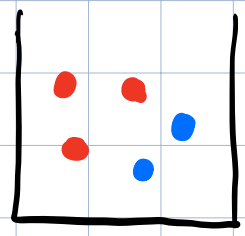
Descrizione	Data, ora e aula	Numero iscritti	Esiti	Verbali caricati	Azioni
Estimation and Filtering	06/09/2024 13:00				
Estimation and Filtering	10/07/2024 09:00				
Estimation and Filtering	17/06/2024 09:00				

SAMPLE SPACE (SET),
EVENTS (SUBSETS),
PROBABILITY (SUBS \rightarrow $[0,1]$)



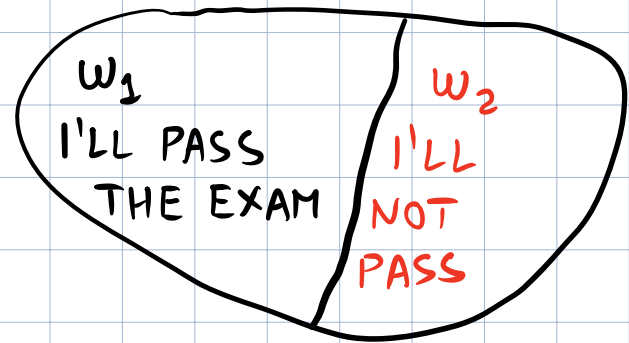
EXAMPLES

URN



$P(\bullet) = 3/5$
(OBJECTIVE)

STUDENT



$P(w_1) = \dots$
(SUBJECTIVE)

AXIOMS AND PROPERTIES

$$P(\emptyset) = 0, \quad P(\Omega) = 1,$$

$$P\left(\bigcup_i A_i\right) = \sum_i P(A_i), \quad A_i \cap A_j = \emptyset \quad (i \neq j)$$

$$P(A) = 1 - P(\bar{A}), \quad \bar{A} = \Omega \setminus A$$

$$P(A \cup B) = P(A) + P(B) - P(A \cap B)$$

$$\begin{array}{l} A, B \\ \text{INDEPENDENT} \end{array} \Leftrightarrow P(A \cap B) = P(A)P(B)$$

$$\left(\begin{array}{l} A, B, C \text{ IND.} \Leftrightarrow P(A \cap B) = P(A)P(B), \\ P(A \cap C) = P(A)P(C), \\ P(B \cap C) = P(B)P(C), \\ P(A \cap B \cap C) = P(A)P(B)P(C) \end{array} \right)$$

$$A, B, C, D \text{ IND} \Leftrightarrow \dots \quad \left. \vphantom{A, B, C, D \text{ IND}} \right)$$

BAYES RULE

$$P(A|B) = \frac{P(A \cap B)}{P(B)}$$

• $|B \Rightarrow B$ IS
THE "NEW Ω "

$$P(B|A) = \frac{P(A \cap B)}{P(A)}$$

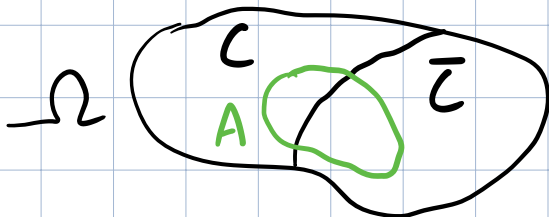


$$P(B|A) = \frac{P(A|B)P(B)}{P(A)}$$

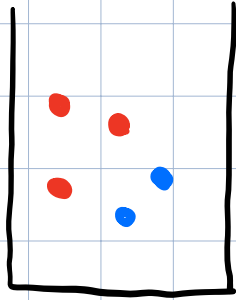
BAYES
RULE

$$P(A) = P(A|C)P(C) + P(A|\bar{C})P(\bar{C})$$

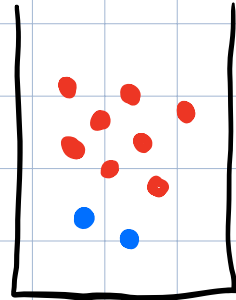
TOTAL
PROBABILITY



EXERCISE



URN A



URN B

AT EXPERIMENT 1 I RANDOMLY
DRAW A OR B, $P(A) = 0.5$

AT EXPERIMENT 2 AND 3 I RANDOMLY
DRAW A BALL FROM THE URN SELECTED
AT EXPERIMENT 1. THEN I REINSERT
THE BALL.

THE "DATA"

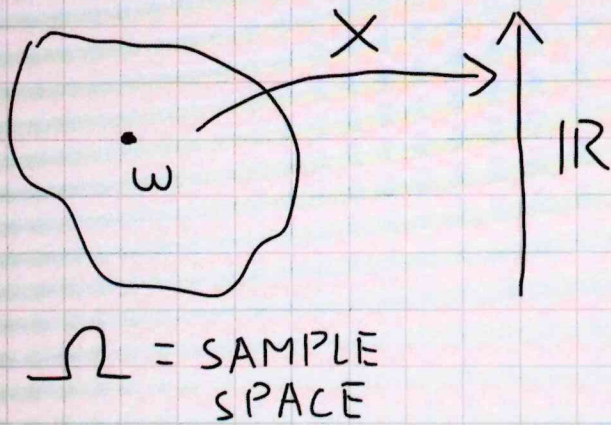
EXPERIMENT 2 : IT IS KNOWN THAT
A BLUE BALL
WAS SELECTED

THE QUESTION

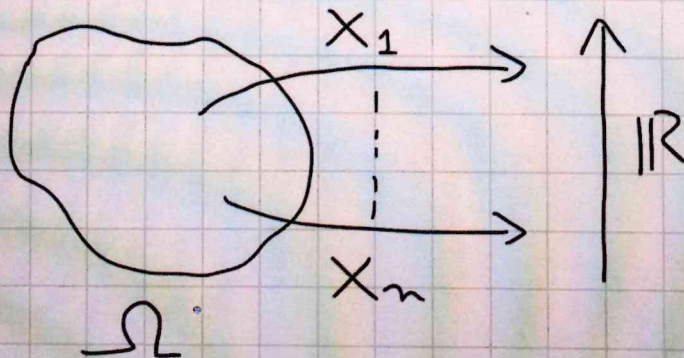
$P(\bullet)$ AT EXPERIMENT 3 = ?

PROBABILITY THEORY

RANDOM VARIABLE



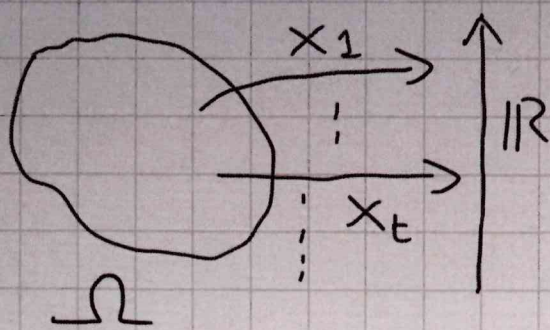
RANDOM VECTOR



$$X = [X_1 \dots X_n]^T$$

COLUMN VECTOR

STOCHASTIC PROCESS



TIME t NOW
INDEXES THE R.V.

$$x_t = s(t), t \in \mathbb{Z}$$

PROBABILITY

$X(\omega)$ IN PRACTICE IS NEVER GIVEN.

WE NEED A TOOL TO COMPUTE PRO-
BABILITY OF EVENTS.

LET EACH R.V. BE CONTINUOUS (IT
CAN ASSUME AN INFINITE NUMBER
OF VALUES) AND X EQUIPPED
WITH A PROBABILITY DENSITY

$$p_x(x) = p_x(x_1, \dots, x_n)$$

THEN WE CAN COMPUTE

- PROBABILITY OF EVENTS

$$P(x \in A) = \int_A p_x(x) dx$$

- EXPECTATION

$$E[x] = \int x \cdot p_x(x) dx$$

- EXPECTATION

$$E[x_i] = \int x_i p_x(x) dx$$

$$E[X] := \begin{bmatrix} E[x_1] \\ \vdots \\ E[x_n] \end{bmatrix}$$

- VARIANCE

$$\text{VAR}(x_i) = E[(x_i - E x_i)^2]$$

$$= E x_i^2 - (E x_i)^2$$

FUNDAMENTAL EXAMPLE:

GAUSSIAN

SCALAR

$$x \sim N(\mu, \sigma^2)$$

$$p(x) = \frac{1}{\sqrt{2\pi}\sigma} e^{-\frac{1}{2}\left(\frac{x-\mu}{\sigma}\right)^2}$$

VECTOR

$$x \sim N(\mu, \Sigma), \quad \Sigma = \Sigma^T \geq 0, \quad \Sigma \in \mathbb{R}^{n \times n}$$

VECTOR

$$x \sim N(\mu, \Sigma), \quad \Sigma = \Sigma^T \geq 0, \quad \Sigma \in \mathbb{R}^{n \times n}$$

$\Sigma =$ COVARIANCE, FOR SIMPLICITY
FULL RANK BELOW

$$p(x) = \frac{1}{\sqrt{|2\pi\Sigma|}} e^{-\frac{1}{2}(x-\mu)^T \Sigma^{-1} (x-\mu)}$$

$$|\Sigma| = \det \Sigma$$

$$E x = \mu = \begin{bmatrix} \mu_1 \\ \vdots \\ \mu_n \end{bmatrix}, \quad \text{VAR } x = \Sigma$$

JOINTLY GAUSSIAN

x, y ARE JOINTLY GAUSSIAN

IF $z = \begin{bmatrix} x \\ y \end{bmatrix}$ IS A GAUSSIAN VECTOR

$$z \sim N\left(\begin{bmatrix} \mu_x \\ \mu_y \end{bmatrix}, \Sigma\right), \quad E x = \mu_x, \quad E y = \mu_y$$

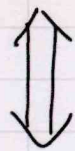
$$\Sigma = \begin{bmatrix} \Sigma_x & \Sigma_{xy} \\ \Sigma_{yx} & \Sigma_y \end{bmatrix}, \quad \Sigma_{xy} = \Sigma_{yx}^T$$

$$\Sigma = \begin{bmatrix} \Sigma_x & \Sigma_{xy} \\ \Sigma_{yx} & \Sigma_y \end{bmatrix}, \quad \Sigma_{xy} = \Sigma_{yx}^T$$

$$\text{VAR } X = \Sigma_x, \quad \text{VAR } Y = \Sigma_y$$

INDEPENDENCE

$$\{x_i\}_{i=1}^n \text{ INDEPENDENT}$$



$$p_x(x) = \prod_{i=1}^n p_{x_i}(x_i)$$

GAUSSIAN EXAMPLE

$$\Sigma = \begin{bmatrix} \sigma_1^2 & & 0 \\ & \ddots & \\ 0 & & \sigma_n^2 \end{bmatrix} \Rightarrow p(x) = \prod_{i=1}^n \frac{e^{-\frac{1}{2} \left(\frac{x-\mu_i}{\sigma_i}\right)^2}}{\sqrt{2\pi} \sigma_i}$$

INDEPENDENT \iff UNCORRELATED
GAUSSIAN
PECULIARITY

QUESTION

QUESTION:

X, Y $\stackrel{?}{\implies}$ $X+Y$ GAUSSIAN
GAUSSIAN

NO IN GENERAL

YES IF X, Y JOINTLY GAUSSIAN,
E.G. IF X, Y INDEPENDENT

CALCULATIONS OF EXPECTATIONS AND VARIANCES

- $E[aX_1 + bX_2], a, b \in \mathbb{R}$

||

$$aE X_1 + bE X_2$$

- $E[g(x)] = \int g(x) p(x) dx$

- $\text{VAR}[aX] = a^2 \text{VAR} X$

- $\text{VAR}[X_1 + X_2] = \text{VAR} X_1 + \text{VAR} X_2$

IF X_1, X_2 UNCORRELATED

NOTE ON UNCORRELATION

WE OFTEN WRITE $X_1 \perp X_2$
TO SAY " X_1 IS UNCORRELATED
FROM X_2 "

$$X_1 \perp X_2 \iff \text{COV}(X_1, X_2) = 0$$
$$\parallel$$
$$E(X_1 - \mu_1)(X_2 - \mu_2)$$

OFTEN WE ASSUME ZERO MEAN

$$X_1 \perp X_2 \iff E X_1 X_2 = 0$$

WITH

$$E X_1 = E X_2 = 0$$

LINEAR AND AFFINE MAPS

IF X IS A RANDOM VECTOR OF
DIMENSION n AND $A \in \mathbb{R}^{m \times n}$,

ONE HAS

- Ax IS A LINEAR TRANSFORMATION OF X
- $Ax + b$ IS AN AFFINE $\parallel \parallel \parallel$

$$\text{VAR}(Ax + b) = \text{VAR} Ax = A \Sigma_x A^T$$

• $Ax + b$ IS AN AFFINE || || ||

• $\text{VAR}(Ax + b) = \text{VAR} Ax = A \Sigma_x A^T$

WHERE $\text{VAR} x = \Sigma_x$

COVARIANCES FROM LINEAR TRANSFORMATION

$$x = \begin{bmatrix} x_1 \\ \vdots \\ x_n \end{bmatrix}, \quad y = \begin{bmatrix} y_1 \\ \vdots \\ y_r \end{bmatrix}$$

$$A \in \mathbb{R}^{a \times n}, \quad B \in \mathbb{R}^{b \times r}$$

$$\text{COV}(x, y) = \Sigma_{xy} = \underbrace{E[(x - \mu_x)(y - \mu_y)^T]}_{\in \mathbb{R}^{n \times r}}$$

Σ_{xy} IN GENERAL IS NOT SYMMETRIC,
HENCE IT IS NOT SEMIDEFINITE
POSITIVE

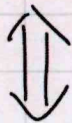
$$\begin{aligned} \text{COV}(Ax, By) &= E\{A(x - \mu_x)(y - \mu_y)^T B^T\} \\ &= A \Sigma_{xy} B^T \end{aligned}$$

(IF $x=y$ AND $A=B$ THE PREVIOUS FORMULA IS OBTAINED)

NOTE ON $\Sigma \geq 0$

- Σ IS S.D.P., I.E. $\Sigma \geq 0$, IF

$$\Sigma = \Sigma^T \text{ AND } v^T \Sigma v \geq 0 \quad \forall v \in \mathbb{R}^n$$



ALL THE EIGENVALUES OF $\Sigma = \Sigma^T$
ARE ≥ 0

- IF $\Sigma = \text{VAR} X$, THEN $\Sigma \geq 0$

- $\Sigma_1 \geq \Sigma_2$ MEANS $\Sigma_1 - \Sigma_2 \geq 0$

- $\Sigma_1 \geq \Sigma_2 \implies \Sigma_1(i,i) \geq \Sigma_2(i,i)$
 $\forall i$

BAYESIAN

ESTIMATION

BAYESIAN ESTIMATION

OFTEN A GOOD MODEL FOR

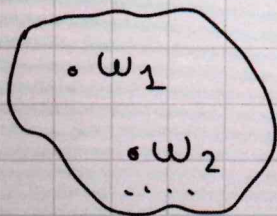
m MEASUREMENTS $\{y_k\}_{k=1}^m$ IS

$$y_k = h_k(x) + \varepsilon_k$$

WHERE

- h_k IS A KNOWN FUNCTION $\forall k$
- ε_k ARE MEASUREMENT ERRORS

THIS IS DESCRIBED USING A
PROBABILITY SPACE



$w_i =$ OUTCOME OF AN
EXPERIMENT

SO, ALL THE w_i ACCOUNT FOR
ALL THE POSSIBLE EXPERIMENTAL
CONDITIONS LEADING TO THE y_k

OFTEN, THE ε_k ARE INDEPENDENT
ZERO-MEAN GAUSSIAN

KEY POINTS: WHAT IS A PROBABILITY SPACE?

KEY POINT: WHAT IS x AND HOW CAN WE OBTAIN IT JUST ASSUMING TO KNOW h , TO KNOW THE PDFS OF $\{\epsilon_k\}$, AND TO OBSERVE $\{y_k\}$

FISHER APPROACH

x IS A DETERMINISTIC VECTOR

BAYES APPROACH

x IS ALSO A RANDOM VECTOR.

THIS VIEW IS USEFUL SINCE WE CAN INCLUDE "A PRIORI" INFORMATION, I.E. BEFORE SEEING $\{y_k\}$, IN THE

PDF $p_x(x)$ OF x

$$p_x(x) = p_x(x_1, \dots, x_n) \quad \text{PRIOR}$$

BAYESIAN ESTIMATION

PROBLEM

TO RECONSTRUCT THE RANDOM

VECTOR x FROM E.G. THE

MODEL

$$y_k = h_k(x) + \varepsilon_k$$

AND THE REALIZATIONS OF THE
RANDOM VARIABLES $\{y_k\}_{k=1}^m$

FOR A BAYESIAN THE SOLUTION
IS THE A POSTERIORI DENSITY
FUNCTION (POSTERIOR) THAT IS
DEFINED BY THE BAYES RULE

POSTERIOR

INGREDIENTS

- LIKELIHOOD, I.E. THE DENSITY
OF $y = [y_1 \dots y_m]$ GIVEN x
 $f(y|x)$ OR $p(y|x)$

EXAMPLE

$m=1$ AND $\varepsilon \sim N(0, \sigma^2)$ AND

ε INDEPENDENT OF x

$$f(y|x) = \frac{1}{\sqrt{2\pi} \sigma} e^{-\frac{1}{2} \left(\frac{y-h(x)}{\sigma} \right)^2}$$

$$f(y|x) = \frac{1}{\sqrt{2\pi}\sigma} e^{-\frac{1}{2}\left(\frac{y-h(x)}{\sigma}\right)^2}$$

- PRIOR DENSITY $p_x(x)$
- MEASUREMENTS y^2 (REALIZATIONS OF y)
- BAYES RULE

$$p(x|y=y^2) = \frac{p(y|x)p_x(x)}{p_y(y)}$$

OFTEN OMITTED

POSTERIOR

$$\propto p(y|x)p_x(x)$$

LIKELIHOOD PRIOR

NOTE THAT

$$p_y(y) = \int p_{xy}(x,y) dx$$

JOINT DENSITY

$$= \int p(y|x)p_x(x) dx$$

JOINT DENSITY

$p_y(y)$ BECOMES JUST A NUMBER
SETTING $y = y^2$, I.E. IT IS

$p_y(y)$ BECOMES JUST A NUMBER
SETTING $y = y^2$, I.E. IT IS
A SCALE FACTOR AFTER
OBSERVING THE MEASUREMENTS

EXAMPLE

$y = h(x, w)$, x INDEP. OF w ,
BOTH RANDOM
VARIABLES

CALCULATE $p(y|x)$ SAYING IF
IT IS NECESSARY TO KNOW
BOTH $p_w(w)$ AND $p_x(x)$

SOL.

KNOWLEDGE OF x DOES NOT AFFECT
THE DENSITY OF w . SO

$$y = h(x, w) = g_x(w)$$

WHERE $g_x(\cdot)$ IS KNOWN WITHOUT
USING p_x .

WE CAN NOW EXPLOIT THE
GENERAL FORMULA

$$y = g(w) \Rightarrow p_y(y) = \sum_{\substack{w \\ \text{s.t.}}} \frac{p_w(w)}{\left| \frac{dg(w)}{dw} \right|}$$

$$y = g(w) \Rightarrow p_y(y) = \sum_{\substack{w \\ \text{s.t.} \\ y = g(w)}} \frac{p_w(w)}{\left| \frac{\partial g(w)}{\partial w} \right|}$$

NOTE: RECALL THAT IN THE CASE OF RANDOM VECTORS THE FORMULA REMAINS CORRECT: IF $g: \mathbb{R}^n \rightarrow \mathbb{R}^m$

$$\frac{\partial g}{\partial w} \in \mathbb{R}^{m \times n}, \quad \frac{\partial g}{\partial w} \Big|_{ij} = \frac{\partial g_i}{\partial w_j},$$

$|\cdot| =$ ABSOLUTE VALUE OF THE DETERMINANT

IF $g: \mathbb{R}^n \rightarrow \mathbb{R}^m$ WITH $m < n$ ONE CAN ADD $n - m$ g_i LIKE $g_i(w) = w_i$ AND STILL USE THE FORMULA

EXAMPLE

LET $y_i \stackrel{||}{\sim} N(\mu, \sigma^2)$ BUT

ACTUALLY NOT GAUSSIAN SINCE

σ^2 IS A RANDOM VARIABLE

$\{y_i\}^n$ CANNOT BE INDEPENDENT

EXAMPLE

LET $y_i \stackrel{||}{\sim} N(\mu, \sigma^2)$ BUT

ACTUALLY NOT GAUSSIAN SINCE

σ^2 IS A RANDOM VARIABLE

$\{y_i\}_{i=1}^n$ CANNOT BE INDEPENDENT

BUT WE ASSUME $y_i | \sigma^2$ INDEP.

COMPUTING $p(y | \sigma^2)$, $y = [y_1 \dots y_n]^T$

$$p(y | \sigma^2) = \prod_{i=1}^n p(y_i | \sigma^2)$$

$$= \frac{1}{(2\pi)^{n/2} (\sigma^2)^{n/2}} \prod_{i=1}^n e^{-\frac{1}{2} \left(\frac{y_i - \mu}{\sigma} \right)^2}$$

$$= \frac{1}{(2\pi)^{n/2} (\sigma^2)^{n/2}} \cdot e^{-\frac{S}{2\sigma^2}},$$

$$S := \sum_{i=1}^n (y_i - \mu)^2$$

LET $\sigma^2 \sim \text{INVGAMMA}(\alpha, \beta)$, THEN

COMPUTE THE POSTERIOR OF σ^2

SINCE $\Theta = \sigma^2$, THE INVGAMMA HAS

USING $\Theta = \sigma^2$, THE INV GAMMA HAS
PDF

$$p_{\Theta}(\theta) \propto \theta^{-\alpha-1} e^{-\beta/\theta}$$



OUR INFO ON σ^2
CAN BE CHANGED
E.G. USING DIFFERENT
 α .

IF $\alpha \downarrow$ THE PRIOR
IS FLAT, LESS INFO

(β COULD ALSO
VARY TO
DESCRIBE OTHER
SHAPES)

NOW:

$$p(\sigma^2 | y) \propto p(y | \sigma^2) p_{\sigma^2}(\sigma^2)$$

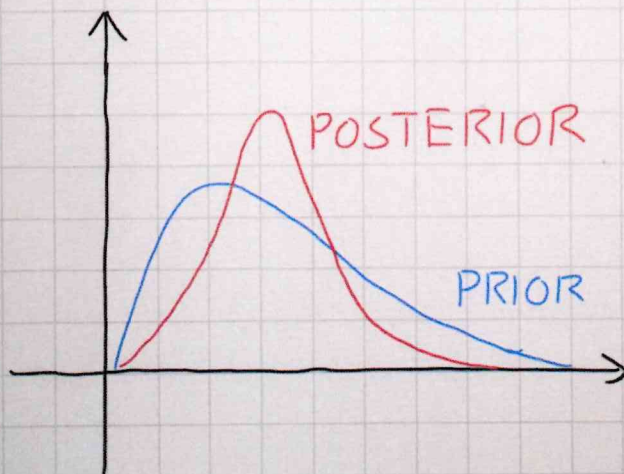
$$\propto \frac{e^{-\frac{S}{2\sigma^2}}}{(\sigma^2)^{n/2}} \cdot (\sigma^2)^{-\alpha-1} e^{-\beta/\sigma^2}$$

$$= (\sigma^2)^{\overset{\text{"NEW } \alpha}{-\alpha - \frac{n}{2}} - 1} e^{-\overset{\text{"NEW } \beta}{\left(\frac{S}{2} + \beta\right)}/\sigma^2}$$

$$\propto \text{INV-GAMMA} \left(\alpha + \frac{n}{2}, \beta + \frac{S}{2} \right)$$

NOTE THAT $\alpha \rightarrow \alpha + \frac{n}{2}$
 $\beta \rightarrow \beta + \frac{S}{2}$ } BOTH INCREASE

SO THE POSTERIOR TENDS TO BE MORE CONCENTRATED, DECAYING FASTER TO ZERO, WITH A DIFFERENT PEAK REGULATED ALSO BY β



IS THE POSTERIOR IN GENERAL ALWAYS LESS "UNCERTAIN" THAN THE PRIOR?

EXAMPLE

THE SCORE OF A STUDENT
(OUTCOME OF AN EXAMINATION)
IS A RANDOM VARIABLE V

PDF OF V IS DEFINED AS FOLLOWS:

EVENT A = "SCORE GIVEN BY"
PROFESSOR #1

EVENT B = "SCORE GIVEN BY"
PROF. #2

$$h(V|A) = \begin{array}{c} \text{DIRAC} \\ \text{DELTA} \end{array} \begin{array}{c} \bullet \\ | \\ \bullet \end{array} \begin{array}{c} \text{AREA}=1 \\ \text{30} \end{array}$$

$h(V|B) = \text{UNIFORM OVER } \{18, 19, \dots, 30\}$

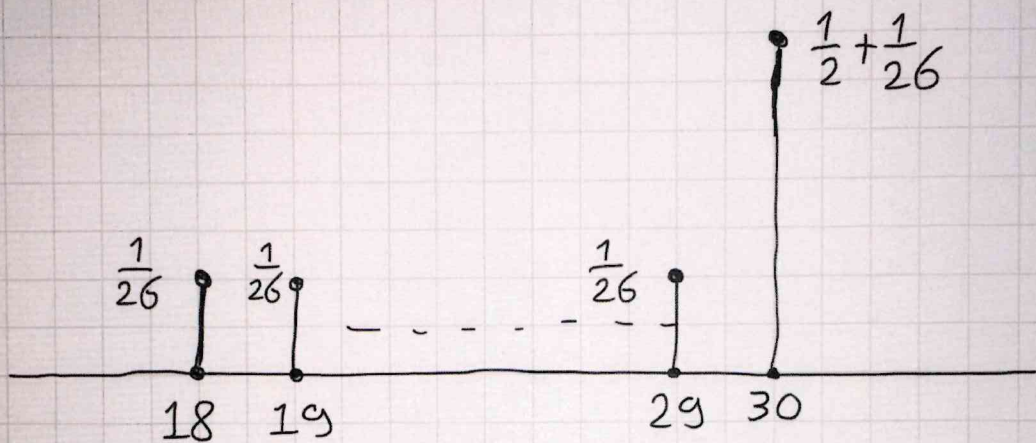
SO THAT $P(V=v|B) = \begin{cases} \frac{1}{13} & \text{IF } v \in \{18, \dots, 30\} \\ 0 & \text{OTHERWISE} \end{cases}$

#1 GIVES THE SCORE WITH PROB. $\frac{1}{2}$

#2 " " " " " " $\frac{1}{2}$

THIS DEFINES THE
PRIOR ON V

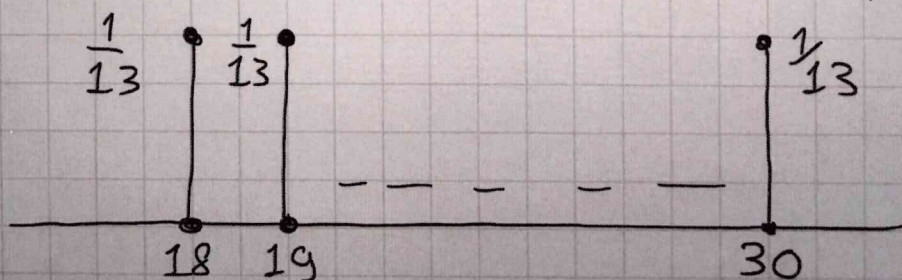
$$h_v(v) = \frac{1}{2} h(v|A) + \frac{1}{2} h(v|B)$$



IF WE CONDITION ON B ,
THE PRIOR BECOMES THE
POSTERIOR $h(v|B)$

WHICH IS THE UNIFORM
(MUCH MORE "UNCERTAIN")

DENSITY HERE DISPLAYED

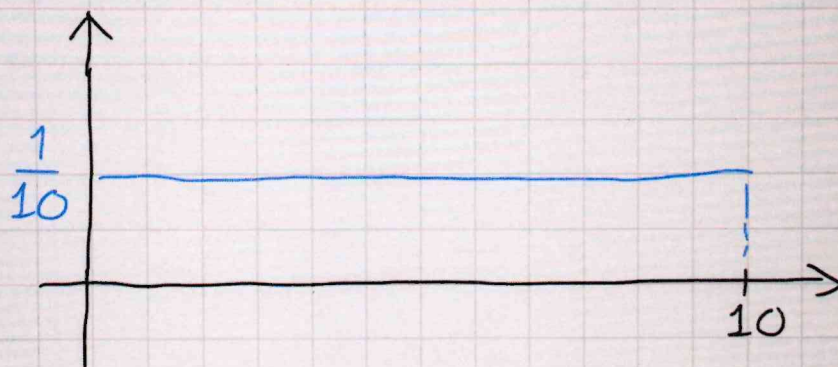


BAYESIAN ESTIMATION:

AN EXAMPLE

$$x \sim U(0, 10)$$

$$h_x(x)$$



$$y = x + \varepsilon, \quad \varepsilon \sim N(0, 1) \quad \text{INDEP. OF } x$$

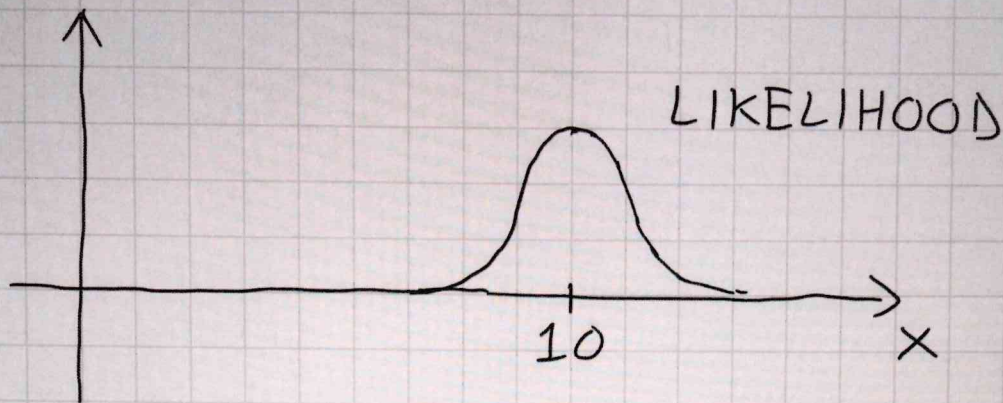
ASSUME TO OBSERVE $y^2 = 10$

$$h(y|x) = \frac{1}{\sqrt{2\pi}} e^{-\frac{1}{2}(x-10)^2}$$

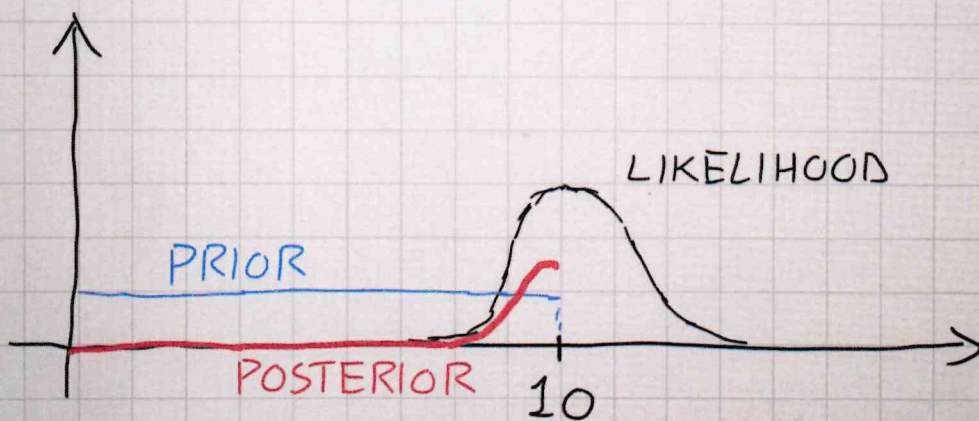
WE PLOT $h(y|x)$ WITH $y=10$

AS A FUNCTION OF x

AS A FUNCTION OF x



THE POSTERIOR IS OBTAINED
(APART FROM A NORMALIZATION
FACTOR) BY MULTIPLYING BY $\mu_x(x)$



SUCH POSTERIOR THUS IS:

$$\mu(x|y=10) = \begin{cases} k e^{-\frac{1}{2}(x-10)^2} & , 0 \leq x \leq 10 \end{cases}$$

SUCH POSTERIOR THUS IS:

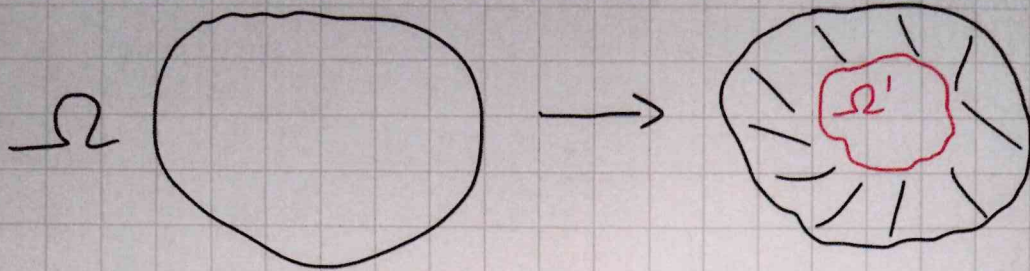
$$p(x|y=10) = \begin{cases} k e^{-\frac{1}{2}(x-10)^2} & , 0 \leq x \leq 10 \\ 0 & , \text{ELSEWHERE} \end{cases}$$

WITH

$$k = \left(\int_0^{10} e^{-\frac{1}{2}(x-10)^2} dx \right)^{-1}$$

NOTES

- THE POSTERIOR IS THE SOLUTION OF AN ESTIMATION PROBLEM IN A BAYESIAN SETTING
- THE EFFECT OF $|y=10$ HERE (AND IN GENERAL) IS TO REMOVE FROM THE SAMPLE SPACE Ω ALL THE ω NOT COMPATIBLE WITH THE OBSERVATIONS



PROBABILITY IS
NOW CONCENTRATED
ONLY ON Ω' AND
NORMALIZED TO ONE

THE FORMULA

$$p(x|y) = \frac{p(y|x)p_x(x)}{p(y)}$$

EXACTLY HAS THIS EFFECT

- OFTEN FROM $p(x|y)$ WE NEED TO OBTAIN POINT ESTIMATES AND BAYES INTERVALS, E.G.

$A =$ BAYES INTERVAL OF 99% LEVEL

IF

$$\int_A p(x|y=y^2) dx = 0.99,$$

SO THAT, AFTER SEEING THE

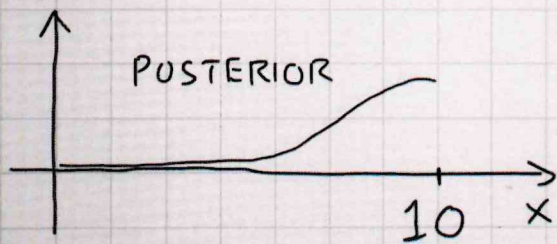
SO THAT, AFTER SEEING THE DATA, $x \in A$ WITH PROBABILITY 0.99,

$$P(x \in A | y = y^2) = 0.99$$

POINT ESTIMATES ARE DISCUSSED BELOW

MAP ESTIMATE

$$\hat{x}^{\text{MAP}} = \underset{x}{\text{ARGMAX}} p(x | y = y^2)$$



IN THE PREVIOUS EXAMPLE

$$\hat{x}^{\text{MAP}} = 10$$

POSTERIOR MEAN/BAYES ESTIMATE/

MINIMUM VARIANCE ESTIMATE

$$\hat{x} = E[x | y = y^2]$$

$$= \int x p(x | y = y^2) dx$$

$$\hat{x} = \int_0^{10} x k e^{-\frac{1}{2}(x-10)^2} dx \quad \text{IN THE PREVIOUS EXAMPLE}$$

NOTE THAT \hat{x} CAN BE DIFFICULT TO OBTAIN SINCE IT CAN REQUIRE DIFFICULT INTEGRALS, OFTEN ALSO IN HIGH-DIMENSION

\hat{x} HAS A FUNDAMENTAL PROPERTY.

TO GRASP IT WE HAVE TO MOVE

FROM ESTIMATE \rightarrow ESTIMATOR
(JUST A NUMBER) (RANDOM)

BAYES ESTIMATOR

$$\hat{x} = \int x \mu(x|y=y^2) dx \quad \text{HAS NOW}$$

TO BE THOUGHT NOT AS A

NUMBER BUT AS A FUNCTION OF y^2

$\hat{x}(y^2)$. BUT NOW WE REPLACE

y^2 WITH THE RANDOM VECTOR y

$$\hat{x}(y) = \int x \mu(x|y) dx \quad \text{IS A RANDOM VECTOR}$$

EXAMPLE

$$\hat{x}(y) = \int_0^{10} x K e^{-\frac{1}{2}(x-y)^2} dx$$

FUNCTION OF y , WITH y TO BE SEEN AS A RANDOM VARIABLE

$$y = " U(0, 10) + N(0, 1) "$$

SUM OF INDEPENDENT R.V. \Rightarrow PDF IS THE CONVOLUTION OF THE TWO PDFs

NOTATION

$$\hat{x}(y) = E[x|y] = \text{CONDITIONAL MEAN}$$

= ESTIMATOR OF x

= RANDOM VECTOR

MINIMUM VARIANCE

ESTIMATOR

STARTING POINT (SCALAR)

x, y RANDOM VARIABLES

WITH KNOWN μ_{xy}

- AIM: TO RECONSTRUCT x FROM y

AIM: TO RECONSTRUCT x FROM y
SO, WE LOOK FOR A $g(y)$ THAT IS
"CLOSE" TO x . MANY TYPES OF
DISTANCES, E.G. $E |x - g(y)|$, BUT
IT HAS ADVANTAGES TO USE

$$E (x - g(y))^2 \quad \text{MEAN SQUARED ERROR}$$

THE OPTIMAL g WILL BE THE
MINIMUM VARIANCE ERROR
ESTIMATOR

IT SEEMS REALLY HARD
TO MINIMIZE W.R.T. g ,
INSTEAD:

THEOREM: CONSIDER THE PROBLEM

$$\hat{g}_y = \underset{g}{\text{ARGMIN}} E (x - g(y))^2$$

THEN

$$\hat{g}_y(y) = E(x|y)$$

THEN

$$\hat{g}(y) = E[X|Y]$$

SO

- $\hat{g}(y)$ IS THE POSTERIOR MEAN

- POSTERIOR MEAN IS THE MINIMUM VAR. ESTIMATOR

- IT IS UNBIASED!

$$E(X - \hat{g}(y)) = E\left[E[(X - \hat{g}(y))|Y]\right]$$

$$= E\left[E[X|Y] - \hat{g}(y)\right]$$

$$= E\left[E[X|Y] - E[X|Y]\right]$$

$$= E[0] = 0$$

PROOF:

$$E(X - g(y))^2 = \int (X - g(y))^2 p_{xy}(x, y) dx dy$$

PROOF:

$$E (x-g(y))^2 = \int (x-g(y))^2 p_{xy}(x,y) dx dy$$

$$= \int \left[\int (x-g(y))^2 p(x|y) dx \right] p_y(y) dy$$

$$= E \left[E \left[(x-g(y))^2 | y \right] \right]$$

LET US STUDY THE
INNER E

$$E \left[(x-g(y))^2 | y \right]$$

$$= E \left[(x - E[x|y] + E[x|y] - g(y))^2 | y \right]$$

$$= E \left[(x - E[x|y])^2 | y \right] \quad (1)$$

$$+ E \left[(E[x|y] - g(y))^2 | y \right] \quad (2)$$

$$+ 2 E \left[\underbrace{(x - E[x|y])}_{\text{HAS ZERO MEAN WHEN CONDITIONED ON } y} \underbrace{(E[x|y] - g(y))}_{\text{DETERMINISTIC WHEN CONDITIONED ON } y} | y \right] \quad (3)$$

HAS ZERO
MEAN
WHEN
CONDITIONED
ON y

DETERMINISTIC
WHEN CONDITIONED
ON y

③ = 0, ① = DOES NOT DEPEND ON y

② = 0 IF $\forall y$ WE DEFINE g
S.T. $g(y) = E[x|y]$

VECTOR CASE

$$(x - g(y))^2 \rightarrow (x - g(y))^T Q (x - g(y))$$
$$Q = Q^T \geq 0$$

AND ONE HAS

$$\text{OPTIMAL } g = \begin{bmatrix} E[x_1|y] \\ \vdots \\ E[x_n|y] \end{bmatrix} = E[x|y]$$

($Q > 0 \Rightarrow$ SUCH SOLUTION IS UNIQUE)

GAUSSIAN ESTIMATION

LET X, Y BE JOINTLY GAUSSIAN

$$X \sim N(\mu_x, \Sigma_x)$$

$$Y \sim N(\mu_y, \Sigma_y), \quad \Sigma_y > 0$$

$$Z = \begin{bmatrix} X \\ Y \end{bmatrix} \sim N\left(\begin{bmatrix} \mu_x \\ \mu_y \end{bmatrix}, \Sigma_z\right)$$

$$\Sigma_z = \begin{bmatrix} \Sigma_x & \Sigma_{xy} \\ \Sigma_{yx} & \Sigma_y \end{bmatrix}$$

PROPOSITION:

$$E[X|Y] = \mu_x + \Sigma_{xy} \Sigma_y^{-1} (Y - \mu_y)$$

$$\text{VAR}[X|Y] = \text{VAR}(X - E[X|Y])$$

COVARIANCE
MATRIX

$$\left(\begin{array}{l} \text{SCALAR CASE} \\ = E(X - E[X|Y])^2 \end{array} \right)$$

$$= \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

$$\text{VAR}[X|Y] = \text{VAR}(X - E[X|Y])$$

COVARIANCE
MATRIX

$$\left(\begin{array}{c} \text{SCALAR CASE} \\ = E(X - E[X|Y])^2 \end{array} \right)$$

$$= \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

(DOES NOT DEPEND ON Y!)

PROOF:

$$\bar{z} = \begin{bmatrix} \bar{x} \\ \bar{y} \end{bmatrix} = \begin{bmatrix} x - \mu_x \\ y - \mu_y \end{bmatrix}$$

$$\hat{x} = \bar{x} + A\bar{y}$$

$$\hat{y} = \bar{y}$$

WHICH A MAKES $\hat{x} \perp \hat{y}$?

$$\text{COV}(\hat{x}, \hat{y}) = \text{COV}(\bar{x} + A\bar{y}, \bar{y})$$

$$= \Sigma_{xy} + A\Sigma_y$$

SO, IT IS = 0 IF

$$A = -\Sigma_{xy} \Sigma_y^{-1}$$

$$A = -\Sigma_{xy} \Sigma_y^{-1}$$

USING SUCH A:

$$E[\bar{x} | \bar{y}] = E[\hat{x} - A\bar{y} | \bar{y}] \quad (\bar{y} = \tilde{y})$$

$$= E[\hat{x} | \bar{y}] - E[A\bar{y} | \bar{y}]$$

$$= E[\hat{x}] - A\bar{y}$$

$$= -A\bar{y}$$

$$= \Sigma_{xy} \Sigma_y^{-1} \bar{y}$$

PLUGGING BACK THE MEANS

$$E[x | y] = \mu_x + \Sigma_{xy} \Sigma_y^{-1} (y - \mu_y)$$

DETERMINISTIC IF $|\bar{y}$

$$\text{VAR}[\bar{x} | \bar{y}] = \text{VAR}(\tilde{x} - \tilde{A}\bar{y} | \bar{y})$$

$$= \text{VAR}(\hat{x} | \bar{y})$$

$$= \text{VAR}(\tilde{x})$$

$$= \text{VAR}(\bar{x} + A\bar{y})$$

$$= \text{COV}(\bar{x} + A\bar{y}, \bar{x} + A\bar{y})$$

$$\dots \quad \dots \quad \dots \quad \dots \quad \dots$$

$$\begin{aligned}
&= \text{VAR}(\bar{x} + A\bar{y}) \\
&= \text{COV}(\bar{x} + A\bar{y}, \bar{x} + A\bar{y}) \\
&= \text{VAR}(\bar{x}) + \text{VAR}(A\bar{y}) + \text{COV}(\bar{x}, A\bar{y}) \\
&\quad + \text{COV}(A\bar{y}, \bar{x})
\end{aligned}$$

$$A^T = -\Sigma_y^{-1} \Sigma_{yx}$$

$$= \Sigma_x + A \Sigma_y A^T + \Sigma_{xy} A^T + A \Sigma_{yx}$$

$$\begin{aligned}
&= \Sigma_x + \Sigma_{xy} \Sigma_y^{-1} \Sigma_y \Sigma_y^{-1} \Sigma_{yx} - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx} \\
&\quad - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}
\end{aligned}$$

$$= \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx} \quad (*)$$


NOTES:

- $\tilde{x} = x - E[x|y] = \text{RECONSTRUCTION ERROR}$

- $\tilde{x} \perp y$

- $\hat{x} \perp y$

- $$\begin{aligned} \text{VAR}(\hat{x}) &= \text{VAR}(x - E[x|y]) \\ &= \text{VAR}(x - E[x|y] | y) \\ &\quad \downarrow \\ &\quad \text{SINCE } \hat{x} \perp y! \\ &= \text{VAR}(x | y) \end{aligned}$$

-  POINTS OUT THAT THE PRIOR VARIANCE (COVARIANCE) Σ_x IS REDUCED BY $\Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$


- SUCH TERM IS $\text{VAR} E(x|y)$, I.E.

$$\text{VAR}[E(x|y)] = \text{VAR}(\mu_x + \Sigma_{xy} \Sigma_y^{-1} (y - \mu_y))$$

$$= \text{VAR} \Sigma_{xy} \Sigma_y^{-1} (y - \mu_y)$$

$$= \Sigma_{xy} \Sigma_y^{-1} \Sigma_y \Sigma_y^{-1} \Sigma_{yx}$$

$$= \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

 MSE = MEAN SQUARED ERROR

- MSE = MEAN SQUARED ERROR

$$\begin{aligned}
 \text{MSE GAUSSIAN ESTIMATOR} &= \text{TRACE}[\text{VAR}(X - E[X|Y])] \\
 &= \text{TRACE}[\text{VAR}(X|Y)] \\
 &= \sum_i E(X_i - E[X_i|Y])^2
 \end{aligned}$$

POSTERIOR MEAN

AND VARIANCE:

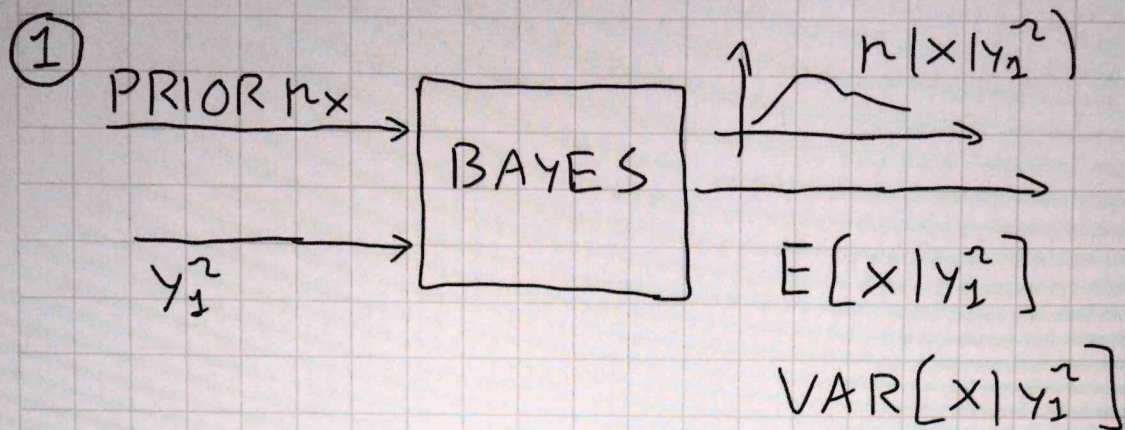
ADDITIONAL REMARKS

$X, Y \sim r_{xy}$ KNOWN, X RANDOM VAR. FOR SIMPLICITY!

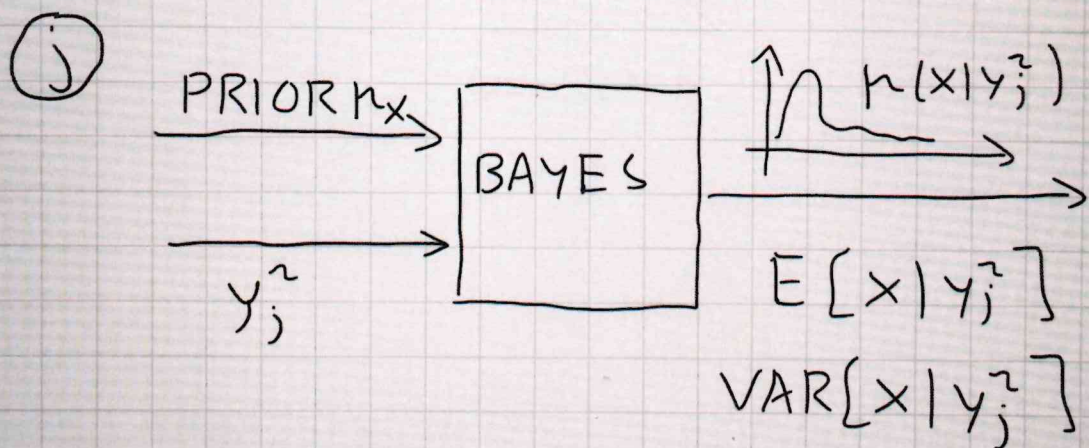
$Y =$ RANDOM VECTOR, EACH EXPERIMENT GIVES DIFFERENT REALIZATIONS

1 ST	EXPERIMENT	y_1^2
!		!
j ^{-TH}		y_j^2

EACH EXPERIMENT LEADS TO A DIFFERENT POSTERIOR



⋮



WE CAN ALSO THINK OF THE POSTERIOR DENSITY AS RANDOM

$\mu(x|y)$
↓
RANDOM

ANY THINK COMPLETE BY

SO, ANY THING COMPUTE BY
 $p(x|y)$ IS RANDOM, E.G.

$$E[x|y]$$

BAYES

ESTIMATOR

$$\text{VAR}[x|y]$$

POSTERIOR

VARIANCE

REGARDING $\text{VAR}[x|y]$

x SCALAR FOR SIMPLICITY

$$\text{VAR}[x|y] = E[(x - E[x|y])^2 | y]$$

$$= \int (x - E[x|y])^2 p_{x|y}(x) dx$$

$$= h(y) \quad \text{AND BECOMES RANDOM} \\ \text{IF WE THINK } y \\ \text{AS RANDOM VECTOR}$$

SO, SUCH VARIANCE IN GENERAL
DEPENDS ON y (RECALL E.G.
STUDENT SCORE)

IN THE GAUSSIAN CASE

$$h(y) = \text{CONSTANT!}$$

$$h(y) = \text{CONSTANT!}$$

NOTE:

IF $\text{VAR}(x|y)$ IS RANDOM,
WHAT IS $E[\text{VAR}(x|y)]$?

$$E[\text{VAR}(x|y)] = E(x - E[x|y])^2$$

AND SO IT IS THE MSE
OF THE BAYES ESTIMATOR.

IT IS COMPUTED BEFORE
SEEING THE DATA (UNCONDITIONAL)

(IN THE GAUSSIAN CASE
IT DOES NOT MAKE ANY
DIFFERENCE TO COMPUTE
IT CONDITIONAL OR UNCOND.)

EXERCISE

PROVE THAT

EXERCISE

PROVE THAT

$$\tilde{x} = x - E[x|y] \perp \hat{x} = E[x|y]$$

\perp = UNCORRELATION

SOL.

$$E[E[x|y]] = E\hat{x} = Ex \Rightarrow E\tilde{x} = 0$$

\Downarrow

$$\text{cov}(\tilde{x}, \hat{x}) = E(x - \hat{x})(\hat{x} - Ex)$$

$$= E\left[E\left[(x - \hat{x})(\hat{x} - Ex) \mid y\right]\right]$$

DETERMINISTIC
WHEN y

$$= E\left[(\hat{x} - Ex) \underbrace{E[x - \hat{x} \mid y]}_{= E[x|y] - E[x|y] = 0}\right]$$

$$= E[0] = 0$$

EXERCISE

EXERCISE

$$x \sim U(0, y)$$

$$y \sim U(0, 2)$$

(X IS NOT
UNIFORM SINCE
Y IS RANDOM)

COMPUTE

$$\hat{x} = E[x|y], \text{MSE } \hat{x}$$

SOL.

IF $z \sim U(0, b)$ ONE HAS

$$E[z] = b/2, \quad E[z^2] = b^2/3$$

$$\text{VAR } z = b^2/12$$

$$x|y \sim U(0, y)$$

↓
DETERMINISTIC

$$E[x|y] = \frac{y}{2} \quad \text{NOW WE SEE } y \text{ AS RANDOM}$$

$$\text{VAR}(x|y) = \frac{y^2}{12} \quad \text{NOW WE SEE } y \text{ AS}$$

$$\text{VAR}(x|y) = \frac{y^2}{12} \quad \text{NOW WE SEE } y \text{ AS RANDOM}$$

$$\text{MSE} = E \text{VAR}(x|y)$$

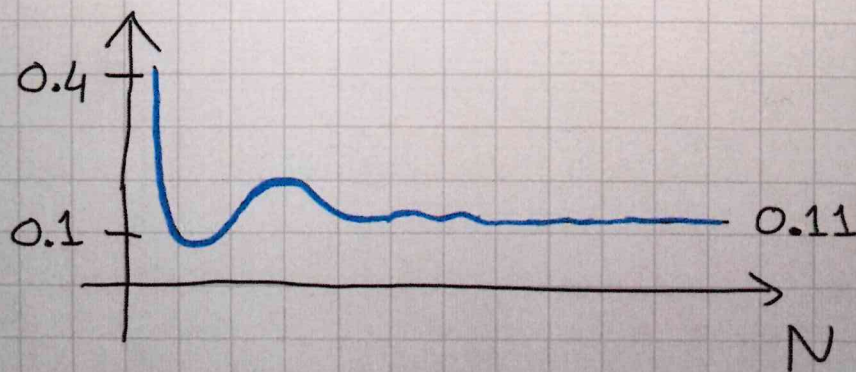
$$= \frac{E y^2}{12} = \left(\frac{2^2}{3} \right) / 12 = \frac{1}{9}$$

MATLAB TEST USING LAW OF LARGE NUMBERS

$$y^i \sim U(0, 2), \quad y(i) = 2 \cdot \text{RAND}$$

$$x^i \sim U(0, y), \quad x(i) = y(i) \cdot \text{RAND}$$

$$\text{MSE} \underset{\substack{\text{LARGE} \\ N}}{\simeq} \frac{\sum_{i=1}^N \left(x(i) - \frac{y(i)}{2} \right)^2}{N}$$



EXAMPLE

GAUSSIAN CASE:

SUMMARY

$$\begin{bmatrix} X \\ Y \end{bmatrix} \sim N \left(\begin{bmatrix} \mu_x \\ \mu_y \end{bmatrix}, \begin{bmatrix} \Sigma_x & \Sigma_{xy} \\ \Sigma_{yx} & \Sigma_y \end{bmatrix} \right)$$

$$E[X|Y] = \mu_x + \Sigma_{xy} \Sigma_y^{-1} (Y - \mu_y) =: \hat{X}$$

$$\tilde{X} := X - E[X|Y]$$

$$\text{VAR}[\tilde{X}] = \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

$$\text{MSE}_{\hat{X}} = \text{TRACE}(\text{VAR} \tilde{X})$$

CONSIDER NOW $X|Y=y^2$

$p(X|Y=y^2)$ COMPUTABLE BY
BAYES RULE



$X|Y=y^2$ IS GAUSSIAN WITH

$$E[X|Y=y^2] = \mu_x + \Sigma_{xy} \Sigma_y^{-1} (y^2 - \mu_y)$$

$$\text{VAR}[X|Y=y^2] = \underbrace{\Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}}_{\text{DOES NOT DEPEND ON } y^2}$$

DOES NOT DEPEND ON y^2

$$\text{VAR}(X - E[X|Y]) = \text{VAR}(X - E[X|Y] | Y=y^2)$$

$$= \text{VAR}(X|Y=y^2)$$

EXAMPLE

$$y = x + e, \quad x \sim N(0, \lambda)$$

$$e \sim N(0, \sigma^2)$$

$$x \perp e$$

$$\hat{x} = ?$$

$$\hat{x} = \Sigma_{xy} \Sigma_y^{-1} y$$

$$\Sigma_{xy} = \text{COV}(x, y)$$

$$= \text{COV}(x, x + e)$$

$$= \text{COV}(x, x) + \text{COV}(x, e)$$

$$= \text{VAR}(x) + 0$$

$$= \lambda$$

$$\Sigma_y = \text{VAR} y = \text{COV}(y, y)$$

$$= \text{COV}(x + e, x + e)$$

$$= \text{COV}(x, x) + \text{COV}(e, e)$$

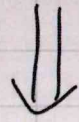
$$+ \text{COV}(x, e) + \text{COV}(e, x)$$

$$= \text{VAR} x + \text{VAR} e$$

$$= \lambda + \sigma^2$$

(SINCE $y = \text{SUM OF TWO INDEP. GAUSSIANS,}$

THEN y IS GAUSSIAN



$$y \sim N(0, \lambda + \sigma^2)$$

$$\hat{x} = \frac{\lambda}{\lambda + \sigma^2} y, \quad \hat{x} \text{ IS GAUSSIAN}$$

$$\left(\text{VAR } \hat{x} = \frac{\lambda^2}{(\lambda + \sigma^2)^2} \cdot \text{VAR } y = \frac{\lambda^2}{\lambda + \sigma^2} \right)$$

$$\text{VAR}(\hat{x} - x) = ?$$

$$\text{VAR}(\hat{x} - x) = \underbrace{\Sigma_x - \Sigma_{xy} \Sigma_y^{-2} \Sigma_{yx}}_{\text{VAR } \hat{x}}$$

$$\text{VAR}(\hat{x}) = \lambda - \frac{\lambda^2}{\lambda + \sigma^2}$$

$$= \frac{\lambda \sigma^2}{\lambda + \sigma^2}$$

"LIMITS"

$$\lambda \rightarrow +\infty \Rightarrow \hat{x} = y, \text{VAR}(\hat{x}) = \sigma^2$$

LIKE SAYING

$$"x \sim U(-\infty, +\infty)"$$

$$\begin{array}{ll} \sigma^2 \rightarrow +\infty \Rightarrow \hat{x} = 0 = Ex, & \\ \text{(NO MEASUREMENT)} & \text{VAR} \hat{x} = \lambda = \text{VAR} X \\ & \text{(POSTERIOR} \\ & \text{PRIOR)} \end{array}$$

HILBERT SPACES

\mathbb{R}^n CONTAINS VECTORS, I.E.

ORDERED n -UPLES OF REAL

NUMBERS $v = [v_1 \dots v_n]$

\mathbb{R}^n IS A VECTOR SPACE SINCE,
WHEN EQUIPPED WITH

$$+ : \mathbb{R}^n \times \mathbb{R}^n \rightarrow \mathbb{R}^n$$

\mathbb{R}^n IS A VECTOR SPACE SINCE,
WHEN EQUIPPED WITH

$$+ : \mathbb{R}^n \times \mathbb{R}^n \rightarrow \mathbb{R}^n$$

$v + h$ HAS COMPONENTS
SUM OF THE COMPONENTS

$$\bullet : \mathbb{R} \times \mathbb{R}^n \rightarrow \mathbb{R}^n$$

$$\lambda v \text{ IS } [\lambda v_1 \dots \lambda v_n],$$

SATISFIES THE 8 AXIOMS.

\mathbb{R}^n IS ALSO HILBERT IF
EQUIPPED WITH

$$\langle v, h \rangle = \sum_{i=1}^n v_i h_i$$

$$\mathbb{R}^n \times \mathbb{R}^n \rightarrow \mathbb{R}$$

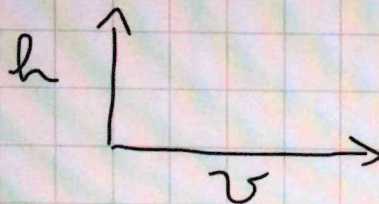
$\langle \cdot, \cdot \rangle$ ABOVE IS AN INNER-PRODUCT

SINCE IT SATISFIES 3 AXIOMS

(SYMMETRY,
BILINEARITY,
POSITIVITY)

USEFULNESS OF INNER-PRODUCTS

a) NOTION OF ORTHOGONALITY


$$\Leftrightarrow \langle v, h \rangle = 0$$

(EASY TO SEE IN \mathbb{R}^3)

b) NOTION OF DISTANCE

$$\begin{aligned} \|v - h\|^2 &= \langle v - h, v - h \rangle \\ &= \sum_{i=1}^n (v_i - h_i)^2 \end{aligned}$$

(ALLOWS TO DEFINE LIMITS,
HILBERT IF ANY CAUCHY SEQ.

IS A CONVERGENT SEQUENCE)

OTHER HILBERT SPACES

BEYOND \mathbb{R}^n :

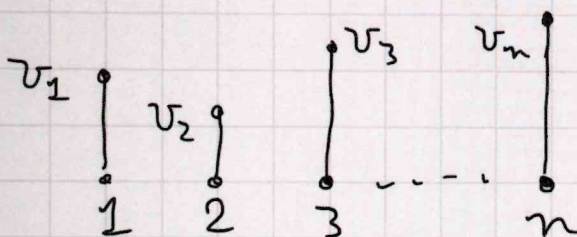
OTHER HILBERT SPACES

BEYOND \mathbb{R}^n :

SPACES OF FUNCTIONS

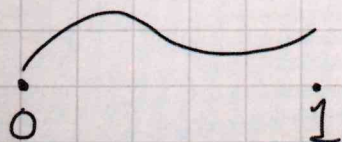
A VECTOR CAN BE SEEN AS

A FUNCTION $\{1, \dots, n\} \rightarrow \mathbb{R}$



$$\langle v, h \rangle = \sum_{i=1}^n v_i h_i$$

IF WE CONSIDER MORE GENERAL DOMAINS, WE OBTAIN GENERAL FUNCTIONS, E.G. $[0, 1] \rightarrow \mathbb{R}$

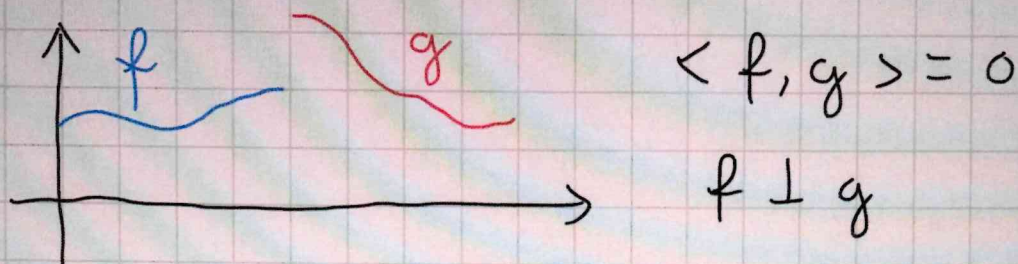


AND WE GENERALIZE THE INNER-PRODUCT AS FOLLOWS

$$\langle f, g \rangle_2 = \int f(x)g(x)dx$$

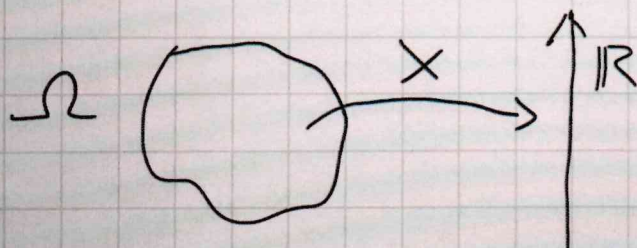
NOW I HAVE ALSO THE CONCEPT OF ORTHOGONAL FUNCTIONS AND DISTANCE BETWEEN FUNCTIONS

NOW I HAVE ALSO THE CONCEPT OF ORTHOGONAL FUNCTIONS AND DISTANCE BETWEEN FUNCTIONS



$$\|f - g\|_2^2 = \int (f(x) - g(x))^2 dx$$

ALSO THE RANDOM VARIABLES ARE FUNCTION EVEN IF ON AN ABSTRACT DOMAIN Ω AND A "STRANGE" INTEGRAL ON Ω



$H =$ SPACE OF ALL RANDOM VARIABLES WITH

- ZERO MEAN
- FINITE VARIANCE

H IS HILBERT WITH

$$\langle x, y \rangle_H = E xy$$

$$= \int_{\Omega} X(\omega) Y(\omega) P(d\omega)$$

$$= \int_{\mathbb{R}^2} xy \mu_{xy}(x, y) dx dy$$

$$\|x\|_H^2 = E x^2 = \text{VAR } x$$

$$\left(\|x\|_H = 0 \Rightarrow X(\omega) = 0 \text{ FOR ALMOST ALL } \omega \text{ W.R.T. } P \right)$$

ADVANTAGES RELATED TO THIS VIEW

- 1) CONCEPTS OF ORTHOGONALITY AND DISTANCE BETWEEN R.V.

$$x \perp y \Leftrightarrow \langle x, y \rangle_H = 0 \quad (= E xy)$$

- 2) IN MANY CASES $E[x|y]$ IS NOT EASILY COMPUTABLE, NOT LINEAR IN y AND AVAILABLE IN CLOSED FORM AS IN THE GAUSSIAN CASE

$$\hat{x} = \Sigma_{xy} \Sigma_y^{-1} y$$

THE GEOMETRICAL VIEW WILL ALLOW US TO FIND AN ALTERNATIVE SUBOPTIMAL ESTIMATOR LINEAR IN y

MINIMUM VARIANCE

LINEAR ESTIMATORS

$E x = E y = 0$ FOR SIMPLICITY
(BUT THEN WE GENERALIZE)

$E[x|y]$ CAN BE COMPLEX AND NOT LINEAR IN y

WE LOOK FOR THE BEST SUBOPTIMAL SOLUTION LINEAR IN y

PROBLEM

$$\hat{g}_y = \underset{\text{LINEAR } g}{\text{ARG MIN}} \|x - g(y)\|_H^2$$

EQUIVALENT TO SOLVING

$$\hat{A} = \underset{A}{\text{ARG MIN}} \|x - Ay\|_H^2$$

($Ay =$ SOMETHING IN THE SUBSPACE GENERATED BY THE $\{y_i\}$)

($Ay = \text{SOMETHING IN THE SUBSPACE}$
GENERATED BY THE $\{y_i\}$)

OBSERVATIONS:

1) IF WE LOOK FOR $Ay + b$ (AFFINE)
WE WOULD FIND $b = 0$ IF $E_x = E_y = 0$

2) IF $E_x \neq 0$ AND $E_y \neq 0$, WE
CAN JUST CONSIDER $\bar{x} = x - \mu_x$,
 $\bar{y} = y - \mu_y$ AND THEN IN THE FINAL
RESULT WE REPLACE \bar{x}, \bar{y} WITH
 $x - \mu_x, y - \mu_y$

3) IN THE GAUSSIAN CASE WE
ALREADY KNOW

$$\hat{c}_y = \underset{c_y}{\text{ARG MIN}} \dots = \underset{\text{LINEAR } c_y}{\text{ARG MIN}} \dots$$

$$\text{AND } \hat{c}_y(y) = \Sigma_{xy} \Sigma_y^{-1} y$$

↓

$$\hat{A} = \Sigma_{xy} \Sigma_y^{-1}$$

SOLUTION USING HILBERT SPACES

DEFINITION:

$$\text{IF } y = [y_1 \dots y_m],$$

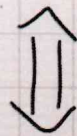
$$\begin{aligned} H(y) &= \text{SPAN} \{y_1, \dots, y_m\} \\ &= \left\{ \sum_{i=1}^m a_i y_i, a_i \in \mathbb{R} \right\} \end{aligned}$$

= SUBSPACE GENERATED
BY $\{y_i\}$

NOTE 1:

$$\{y_i\}_{i=1}^m \text{ IF } \left(\sum_{i=1}^m a_i y_i = 0 \iff \begin{matrix} a_i = 0 \\ \forall i \end{matrix} \right)$$

INDEPENDENT



$$\sum y_i = \text{VARY} > 0$$

NOTE 2: $H(y)$ IS ALWAYS CLOSED

SINCE ITS DIMENSION IS FINITE

! ANY CAUCHY SEQUENCE IN

(ANY CAUCHY SEQUENCE IN $H(Y)$ IS ALSO CONVERGENT IN $H(Y)$, $\forall \varepsilon \exists n$ S.T. $\|z_i - z_j\|_H \leq \varepsilon \forall i, j \geq n$ WITH $z_i \in H(Y) \forall i$, THEN $z_i \rightarrow z \in H(Y)$).

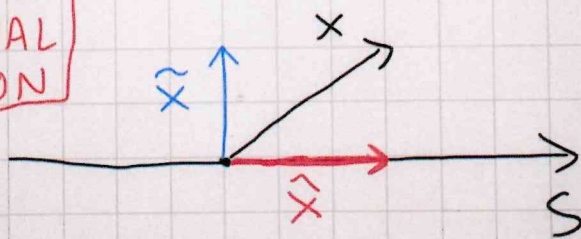
SO, NO HOLES IN $H(Y)$

PROJECTION THEOREM

$H =$ HILBERT SPACE, $S =$ CLOSED SUBSPACE

$x \in H$

$\hat{x} =$ ORTHOGONAL PROJECTION OF x ONTO S



- $\forall x \in H$, ONE CAN WRITE

$$x = \hat{x} + \tilde{x}, \quad \hat{x} \in S, \quad \tilde{x} \in S^\perp$$

$\forall v \in S$
 $\hat{x} \perp v$

- THE DECOMPOSITION IS UNIQUE
- $\hat{x} = \underset{v \in S}{\text{ARGMIN}} \|x - v\|_H$

NOTE 1:

AFTER PROVING THAT THE DECOMP. \exists ,
UNICITY IS OBTAINED BY CONTRADICTION:

$$X = \underset{\in S}{\hat{X}'} + \underset{\in S^\perp}{\tilde{X}'} = \underset{\in S}{\hat{X}} + \underset{\in S^\perp}{\tilde{X}}, \quad \hat{X} \neq \hat{X}'$$

\Downarrow

$$\underset{\in S}{\hat{X}'} - \underset{\in S}{\hat{X}} = \underset{\in S^\perp}{\tilde{X}} - \underset{\in S^\perp}{\tilde{X}'} =: v$$

SO, $v \in S$ AND $v \in S^\perp$

$$\Rightarrow \langle v, v \rangle = 0 = \|v\|^2$$

$\Rightarrow v = 0$ BY AXIOMS
ON INNER PRODUCT

NOTE 2:

IN THE THEOREM $\dim(S)$ CAN BE ∞

HILBERT
Now $H =$ SPACE OF
RANDOM VARIABLES

IF $S = H(y)$ THEN

$$\hat{X} = \sum_{x,y} \sum_y^{-1} y$$

\triangle PROOF:

PROOF:

TO PROVE THAT \hat{x} IS THE PROJECTION,
ONE CAN JUST PROVE THAT

$$x - \hat{x} \perp z, \quad \forall z \in H(y)$$

$$\text{|||}$$
$$x - \hat{x} \perp y_i \quad \forall i = 1, \dots, m$$

WE HAVE ALREADY SEEN THIS
TREATING GAUSSIAN ESTIMATION,
BUT LET US REDO THE CALCULATION:

$$x - \hat{x} = x - \Sigma_{xy} \Sigma_y^{-1} y$$

$$\text{COV}(x - \Sigma_{xy} \Sigma_y^{-1} y, y) = E x y^T$$
$$- E \Sigma_{xy} \Sigma_y^{-1} y y^T$$

$$= \Sigma_{xy} - \Sigma_{xy} \Sigma_y^{-1} E y y^T$$

$$= \Sigma_{xy} - \Sigma_{xy} \Sigma_y^{-1} \Sigma_y = 0$$

(NULL
MATRIX)



0, FROM THE PROJECTION TH.



SO, FROM THE PROJECTION TH.,

$$\hat{x} = \underset{v \in H(y)}{\text{ARG MIN}} \|v - x\|_H^2 = \underset{\text{SCALAR } x}{\text{E}} (v - x)^2$$

SO \hat{x} IS THE LINEAR MINIMUM VARIANCE ESTIMATOR!

NOTE 3: $x = \text{RANDOM VECTOR}$,

$\hat{x} = \Sigma_{xy} \Sigma_y^{-1} y$ HAS THUS THE PROPERTY

$$\tilde{x}_i = (\hat{x}_i - x_i) \perp H(y) \quad \forall i,$$

I.E. EACH COMPONENT OF THE ERROR \tilde{x} IS UNCORRELATED FROM ANY COMPONENT OF THE DATA VECTOR y AND FROM ANY LINEAR COMBINATION OF THE y_i

THE LINEAR MODEL

$$y = Sx + w, \quad x \perp w$$

THE LINEAR MODEL

$$y = Sx + w, \quad x \perp w$$

$$\Sigma_x = \text{VAR} x = P$$

$$\Sigma_w = \text{VAR} w = R$$

$$E x = E w = 0$$

LET US COMPUTE $\hat{E}[x|y]$ WHERE
 \hat{E} = PROJECTION (MINIMUM VARIANCE LINEAR ESTIMATOR)

$$\begin{aligned} \Sigma_{xy} &= \text{COV}(x, y) = \text{COV}(x, Sx + w) \\ &= \text{COV}(x, Sx) = \text{COV}(x, x) S^T = P S^T \end{aligned}$$

$$\begin{aligned} \Sigma_y &= \text{VAR} y = \text{COV}(y, y) \\ &= \text{COV}(Sx + w, Sx + w) \\ &= \text{VAR}(Sx) + \text{VAR}(w) \\ &= S P S^T + R \end{aligned}$$

HENCE

$$\begin{aligned} \hat{E}[x|y] &= \Sigma_{xy} \Sigma_y^{-1} y \\ &= P S^T (S P S^T + R)^{-1} y \end{aligned}$$

$$\text{VAR}(\hat{x}) = S (S P S^T + R)^{-1} S^T$$

$$\hat{E}[x|y] = \Sigma_{xy} \Sigma_y^{-1} y$$

$$= P S^T (S P S^T + R)^{-1} y$$

$$\text{VAR}(\tilde{x}) = \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

$$= P - P S^T (S P S^T + R)^{-1} S P$$

ALTERNATIVE FORMULA

A, C SQUARE AND INVERTIBLE

THEN INVERSION LEMMA SAYS

$$(A + B C D)^{-1} = A^{-1} - A^{-1} B (C^{-1} + D A^{-1} B)^{-1} D A^{-1}$$

THEN IN OUR
CONTEXT:

$$\hat{E}[x|y] = (P^{-1} + S^T R^{-1} S)^{-1} S^T R^{-1} y$$

$$\text{VAR} \tilde{x} = (P^{-1} + S^T R^{-1} S)^{-1}$$

(LITTLE INFORMATIVE PRIOR MEANS

$P^{-1} \rightarrow 0 \dots$)

SUMMARY

$$\hat{x} = \hat{E}[x|y] = \hat{E}[x | H(y)] \quad \text{SOLVES}$$

MIN $E \|x - v\|^2$ SCALAR

SUMMARY

$$\hat{x} = \hat{E}[x|y] = \hat{E}[x | H(y)] \quad \text{SOLVES}$$

$$\min_{v \in H(y)} E \|x - v\|^2 \quad \text{SCALAR CASE}$$

HENCE IT SOLVES

$$\min_A \|x - Ay\|_H^2 \Rightarrow \hat{A} = \Sigma_{xy} \Sigma_y^{-1}$$

(VALID ALSO FOR THE VECTOR CASE!)

WE HAVE ALSO UNDERSTOOD THAT

THE BEST LINEAR ESTIMATOR OF x

BASED ON y ONLY DEPENDS ON THE

1ST AND 2ND MOMENTS OF x, y AND

ONE HAS

$$\hat{E}[x|y] = \mu_x + \Sigma_{xy} \Sigma_y^{-1} (y - \mu_y)$$

$$\tilde{x} = x - \hat{x}$$

$$\text{VAR } \tilde{x} = \Sigma_x - \Sigma_{xy} \Sigma_y^{-1} \Sigma_{yx}$$

IN THE GAUSSIAN CASE (x, y JOINTLY GAUSSIAN)

$$\hat{E}[x|y] = E[x|y]$$

IN THE GAUSSIAN CASE (x, y JOINTLY GAUSSIAN)

- $\hat{E}[x|y] = E[x|y]$
- $x|y$ IS GAUSSIAN WITH MEAN $E[x|y]$, COVARIANCE $\text{VAR} \tilde{x}$
|||
 $\text{VAR}[x|y]$

(IN FACT, $x - E[x|y]$ IS INDEPENDENT OF y , SO THAT $\text{VAR} \tilde{x}$
 $= \text{VAR}[x - E[x|y]]$
 $= \text{VAR}[x - E[x|y] | y]$
 $= \text{VAR}[x|y - E[x|y]]$
 $= \text{VAR}[x|y]$)

EXERCISE

PROVE THAT, IF $E_x = E_y = 0$, THEN

$$\hat{g} = \text{ARG-MIN}_{g(y) \text{ s.t. } g(y) = Ay + b} E \underbrace{\|g(y) - x\|^2}_{\text{EUCLIDEAN CASE, VECTOR CASE!}}$$

EXERCISE

PROVE THAT, IF $E x = E y = 0$, THEN

$$\hat{y} = \underset{\substack{g(y) \text{ s.t.} \\ g(y) = Ay + b}}{\text{ARG MIN}} \underbrace{E \|g(y) - x\|^2}_{\substack{\text{EUCLIDEAN} \\ \text{CASE, VECTOR} \\ \text{CASE!}}}$$

$$= \underset{\substack{g(y) \text{ s.t.} \\ g(y) = Ay}}{\text{ARG MIN}} E \|g(y) - x\|^2$$

WE HAVE TO PROVE THAT $b = 0$.

$$\|x - Ay - b\|^2 = \|x - Ay\|^2 + \|b\|^2 - 2 \langle x - Ay, b \rangle$$

WE TAKE THE MEAN:

$$(E \|x - Ay\|^2) + \|b\|^2 - 2 \underbrace{\langle E(x - Ay), b \rangle}_{= 0}$$

$\Rightarrow b$ MUST BE 0 TO MINIMIZE THE OBJECTIVE

EXERCISE

EXERCISE

n MEASUREMENTS OF A RESISTOR
OF NOMINAL VALUE x_0 AND TOLERANCE
 α

PRIOR: WITH HIGH PROBABILITY

$$x_0 - \alpha x_0 \leq x \leq \alpha x_0 + x_0$$

\Rightarrow USING A GAUSSIAN PRIOR

REASONABLE TO SAY

$$x \sim N(x_0, \sigma_x^2)$$

$$\text{WITH } 2\sigma_x = \alpha x_0 \Rightarrow \sigma_x = \frac{\alpha x_0}{2}$$

THE MODEL IS

$$y_i = x + w_i, \quad i = 1, \dots, n \quad \text{AND}$$

W.H.P. WE KNOW

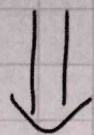
$$-\beta_i x_M \leq w_i \leq \beta_i x_M$$

SO IT IS REASONABLE TO SAY

$$w_i \sim N(0, \sigma_i^2)$$

$$\text{WITH } 2\sigma_i = \beta_i x_M \Rightarrow \sigma_i = \frac{\beta_i x_M}{2}$$

$$\text{WITH } 2\sigma_i = \beta_i x_M \Rightarrow \sigma_i = \frac{\beta_i x_M}{2}$$



$$y = Sx + w, \quad x \perp w$$

$$S = \begin{bmatrix} 1 \\ 1 \\ \vdots \\ 1 \end{bmatrix}, \quad P = \sigma_x^2$$

$$R = \begin{bmatrix} \sigma_1^2 & & & 0 \\ & \ddots & & \\ 0 & & \ddots & \\ & & & \sigma_n^2 \end{bmatrix}$$

SOME CALCULATIONS

$$S^T R^{-1} S = \sum_{i=1}^n \frac{1}{\sigma_i^2}$$

$$\begin{aligned} S^T R^{-1} (y - \mu_y) &= S^T R^{-1} (y - \mu_x) \\ &= S^T R^{-1} \left(y - \begin{bmatrix} x_0 \\ \vdots \\ x_0 \end{bmatrix} \right) \end{aligned}$$

$$= \sum_{i=1}^n \frac{y_i - x_0}{\sigma_i^2}$$

$$= \sum_{i=1}^n \frac{y_i - x_0}{\sigma_i^2}$$

$$E[x|y] = x_0 + \left(\underbrace{\frac{1}{\sigma_x^2}}_{p-1} + \underbrace{\sum_{i=1}^n \frac{1}{\sigma_i^2}}_{S^T R^{-2} S} \right)^{-1} \underbrace{\sum_{i=1}^n \frac{y_i - x_0}{\sigma_i^2}}_{S^T R^{-2} (y - \mu_y)}$$

EXERCISE

$$\{y_t\}_{t=1}^3, \quad y_t = \sin(\omega t), \quad \omega \sim U(-\pi, \pi)$$

I WANT TO ESTIMATE y_3 USING y_1, y_2

$$E y_t = \int_{-\pi}^{\pi} \sin(\omega t) \cdot \underbrace{\frac{1}{2\pi}}_{\text{PRIOR ON } \omega} d\omega = 0$$

$$E y_t y_s = \int_{-\pi}^{\pi} \sin(\omega t) \sin(\omega s) \frac{1}{2\pi} d\omega, \quad t \neq s$$

= 0 (ORTHOGONALITY OF FOURIER BASIS)

$$\hat{E}[y_3 | y_1, y_2] =$$

$$\underbrace{[\text{COV}(y_3, y_1) \quad \text{COV}(y_3, y_2)]}_{\text{COV}(y_3, [y_1, y_2])} (\text{VAR}(y_1, y_2))^{-1} \begin{bmatrix} y_1 \\ y_2 \end{bmatrix}$$

$$\hat{E}[Y_3 | Y_1, Y_2] =$$

$$\underbrace{[\text{COV}(Y_3, Y_1) \quad \text{COV}(Y_3, Y_2)] (\text{VAR}(Y_1, Y_2))^{-1}}_{= [0 \quad 0]} \begin{bmatrix} Y_1 \\ Y_2 \end{bmatrix}$$

$$= 0$$

THE LINEAR ESTIMATOR DOES NOT CHANGE THE PRIOR:

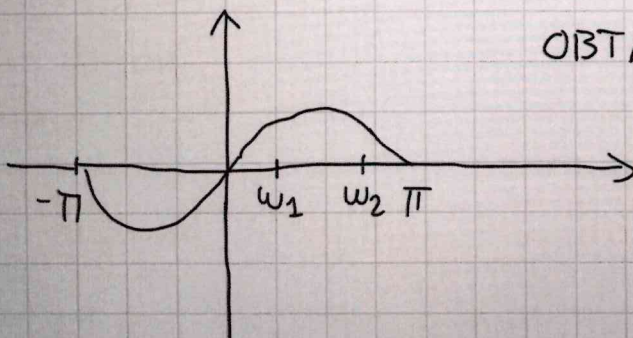
$$\hat{E}[Y_3 | Y_1, Y_2] = \hat{E} Y_3 = 0$$

BUT WE CAN PERFECTLY

RECONSTRUCT w FROM Y_1, Y_2 !

IN FACT:

$$Y_1 = \sin w$$



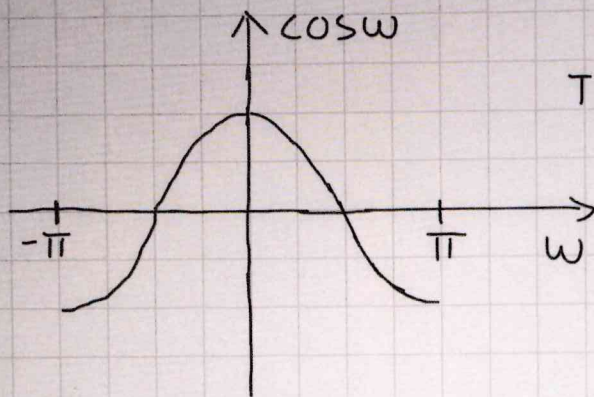
OBTAIN TWO
POSSIBLE
VALUES
FOR w ,
 w_1, w_2

$$Y_2 = \sin 2w = 2 \sin w \cos w$$

$$\rightarrow \cos w = \frac{Y_2}{Y_1}$$

$$y_2 = \sin 2\omega = 2 \sin \omega \cos \omega$$

$$\Rightarrow \cos \omega = \frac{y_2}{2y_1}$$



TWO POSSIBLE VALUES,
ONE POSITIVE,
ONE NEGATIVE,
BUT I KNOW
FROM y_1
THE SIGN!

UNIQUE VALUE OF ω OBTAINED,
PERFECT ESTIMATE

EXERCISE

RADAR / SONAR SIGNAL

$$y(t) = A \sin(\omega_0 t + \theta) + w(t)$$

- $A =$ KNOWN AMPLITUDE, ω_0 KNOWN

- $\theta =$ ECO'S DELAY

$$\theta \sim U(0, 2\pi)$$

- $w(t)$ WHITE NOISE $\perp \theta$,

$$w(t) \sim (0, \sigma^2)$$

ESTIMATE θ FROM $\{y_t\}_{t=1}^N$

ESTIMATE θ FROM $\{y_t\}_{t=1}^N$

SOL. LET US SEE AS A LINEAR ESTIMATION PROBLEM

$$x = \begin{bmatrix} \cos \theta \\ \sin \theta \end{bmatrix}$$

$$\sin(\omega_0 t + \theta) = (\sin \omega_0 t) \cos \theta + (\cos \omega_0 t) \sin \theta$$

\Downarrow

$$y = Sx + w$$

$$y = \begin{bmatrix} y_1 \\ \vdots \\ y_N \end{bmatrix}, \quad S = A \begin{bmatrix} \sin \omega_0 & \cos \omega_0 \\ \vdots & \vdots \\ \sin \omega_0 N & \cos \omega_0 N \end{bmatrix}$$

$$S \in \mathbb{R}^{N \times 2},$$

$$w = \begin{bmatrix} w_1 \\ \vdots \\ w_N \end{bmatrix}$$

$$E x = \begin{bmatrix} 0 \\ 0 \end{bmatrix} \text{ (BY SYMMETRY OF THE INTEGRALS)}$$

$$E \cos^2 \theta = \int_0^{2\pi} \frac{\cos^2 \theta}{2\pi} d\theta$$

$$= \frac{\theta + \sin \theta \cos \theta}{2 \cdot 2\pi} \Big|_0^{2\pi} = \frac{1}{2}$$

$$E \cos \theta \sin \theta = 0$$

$$E \sin^2 \theta = \frac{1}{2}$$

⇓

$$\text{VAR } x = P = \begin{bmatrix} 1/2 & 0 \\ 0 & 1/2 \end{bmatrix}$$

$$\text{VAR } w = R = \sigma^2 I_N$$

BEST LINEAR ESTIMATOR

BEST LINEAR ESTIMATOR

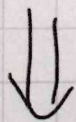
$$S^T R^{-1} S = \frac{S^T S}{\sigma^2}$$

$$= \frac{A^2}{\sigma^2} \begin{bmatrix} \sum_{t=1}^N \sin^2 \omega_0 t & \sum_{t=1}^N \sin \omega_0 t \cos \omega_0 t \\ \sum_{t=1}^N \sin \omega_0 t \cos \omega_0 t & \sum_{t=1}^N \cos^2 \omega_0 t \end{bmatrix}$$

$$S^T R^{-1} y = \frac{A}{\sigma^2} \begin{bmatrix} \sum_{t=1}^N y(t) \sin \omega_0 t \\ \sum_{t=1}^N y(t) \cos \omega_0 t \end{bmatrix},$$

$(E y = E x = 0)$

$$P^{-1} = \begin{bmatrix} 2 & 0 \\ 0 & 2 \end{bmatrix}$$



$$\hat{E}[x|y] = (P^{-1} + S^T R^{-1} S)^{-1} S^T R^{-1} y$$

HOMEWORK: WHAT HAPPENS
FOR LARGE N ?

FROM STATIC TO DYNAMIC ESTIMATE

$\{y_t\}_{t=t_0, t_0+1, \dots}$ COLLECTED OVER TIME,
FLOW OF OBSERVATIONS...

$$y^t := \begin{bmatrix} y_1 \\ \vdots \\ y_t \end{bmatrix}$$

AND WE WANT TO
ESTIMATE x

$$E x = E y^t = 0$$

$$\hat{x}_t = \sum_{x y^t} \sum_{y^t}^{-1} y^t$$

ESTIMATE
OF x AT
INSTANT t

WE APPLY THE
STATIC FORMULA

BUT AS NEW DATA ARRIVE
(t INCREASES) WE HAVE TO

RECOMPUTE \hat{x}_t

• $\sum_{y^t}^{-1}$ REQUIRES $O(t^3)$ OPERATIONS

AND $t \rightarrow +\infty$

AND $t \rightarrow +\infty$

- y^t HAS TO BE STORED IN MEMORY

IT WOULD BE INSTEAD FUNDAMENTAL
TO COMPUTE \hat{x}_{t+1} IN A
RECURSIVE WAY

$$\hat{x}_{t+1} = f \left(\hat{x}_t, y_{t+1} \right)$$

PREVIOUS ESTIMATE ONLY THE LAST MEASUREMENT

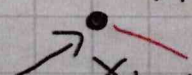
BUT THERE IS AN ADDITIONAL
KEY PROBLEM:

x ITSELF COULD BE DYNAMIC

$x_t = x$ VARIES OVER TIME

$\hat{x}_t =$ ESTIMATE OF x_t BASED
ON y^t

ROBOT'S POSITION
AT $t+1$



TO COMPUTE x_{t+1} IN A

RECURSIVE WAY

$$\hat{x}_{t+1} = f(\hat{x}_t, y_{t+1})$$

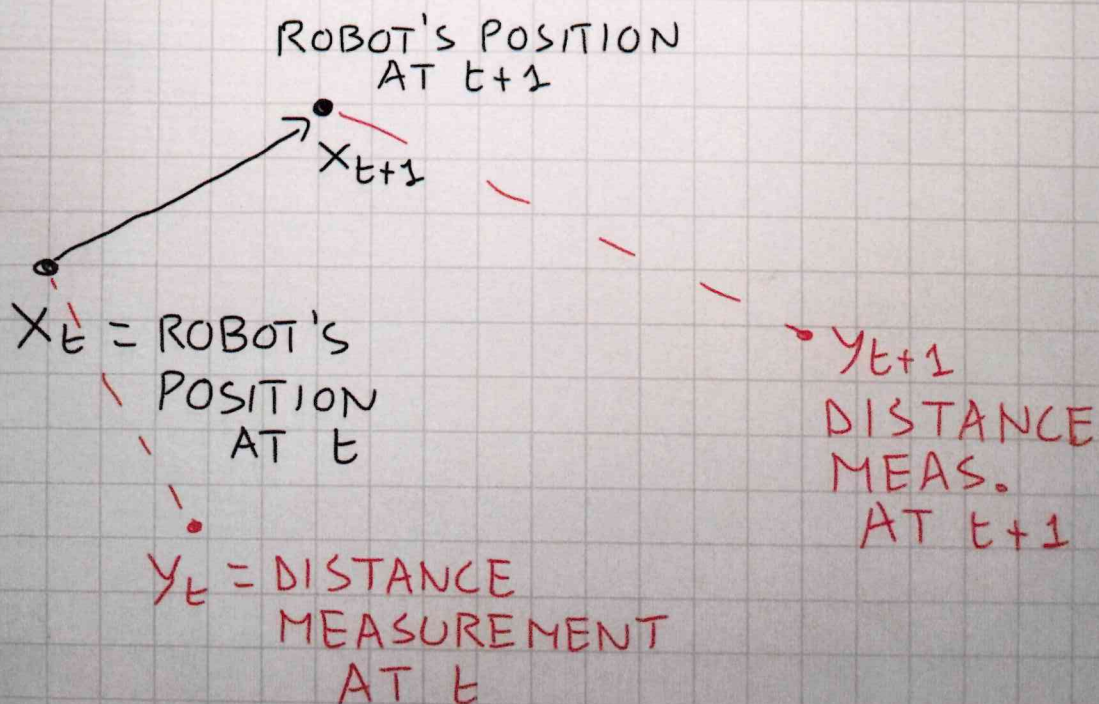
PREVIOUS ESTIMATE ONLY THE LAST MEASUREMENT

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x ITSELF COULD BE DYNAMIC

$x_t = x$ VARIES OVER TIME

$\hat{x}_t =$ ESTIMATE OF x_t BASED ON y^t



FILTERING AND PREDICTION

SCALAR CASE

$\{x_t\}, \{y_t\}$ ARE

STOCHASTIC PROCESSES

DATA ARE $\{y(t), t \in I\}$

FILTERING

$I = [t_0, t]$ AND I WANT

$$E \left[x_t \mid \{y_t\}_{t \in I} \right] =: \hat{x}(t|t)$$

CAUSAL
ESTIMATOR

PREDICTION

$I = [t_0, t]$ AND I WANT

$$E \left[x_{t+h} \mid \{y_t\}_{t \in I} \right] =: \hat{x}(t+h|t), h > 0$$

IT IS A PREDICTOR,
STILL A
CAUSAL ESTIMATOR

SOMETIMES, AS TARGET ALSO

SOMETIMES, AS TARGET ALSO

$$E[y_{t+h} | \{y_k\}_{k \in I}] =: \hat{y}(t+h|t)$$

OUTPUT
PREDICTOR

\hat{E} IN PLACE OF E WILL BE THE
TARGET FOR THE MOMENT

WIENER APPROACH

$\{x_t\}, \{y(t)\}$ JOINTLY STATIONARY
(IN A WEAK SENSE)

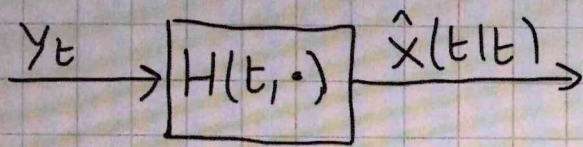
AND $t_0 \rightarrow -\infty$, SO THAT

$$I = [-\infty, t]$$

WIENER SEES THE LINEAR
ESTIMATOR / FILTER

$$\hat{x}(t|t) = \sum_{k=t_0}^t H(t, k) y(k)$$

AS A DYNAMIC SYSTEM



UNDER WIENER ASSUMPTIONS

$H(t, k)$ BECOMES $H(t-k)$

(TIME-INVARIANT SYSTEM).

WE COMPUTE H JUST ONCE

AND WE USE IT $\forall t$.

IT CAN BE OBTAINED BY SOLVING
THE WIENER-HOPF EQUATION

COMPUTATION IS EASY

ONLY IF THE SPECTRA OF

THE PROCESSES ARE RATIONAL

(SUCH PROCESSES ARE THE OUTPUTS

OF FINITE-DIMENSIONAL TIME-INVARIANT
LINEAR SYSTEMS

WITH WHITE NOISE AS INPUT

STARTING AT $t_0 = -\infty$)

SCALAR EXAMPLE

x, y SCALAR

$$\xi_x(\tau) = E x_t x_{t+\tau}$$

$$\downarrow \text{Z-TRANSFORM } \left(\sum_k \xi_x(k) z^{-k} \right)$$

$$S_x(z) = \frac{\text{POLYNOMIAL}}{\text{POLYNOMIAL}}$$

SIMILAR DEFINITIONS THEN FOR

$$S_{xy}(z), S_y(z), S_{yx}(z)$$

IN THE MATRIX CASE

$$S_{xy}(z) = S_{yx}^T(z^*) = (S_{xy}(z))^*$$

$$z = \sigma + j\omega$$

$$z^* = \sigma - j\omega$$

FROM WIENER TO KALMAN

IMPORTANCE OF WIENER: TO SEE THE

SIGNALS OF INTEREST AS STOCHASTIC

WIENER PREDICTOR

LET $s(t)$ BE A STATIONARY STOCHASTIC PROCESS OVER $t \in \mathbb{Z}$ WITH

$$R_s(\tau) = E s(t) s(t-\tau)$$

ITS POWER SPECTRUM IS

$$\bar{\Phi}_s(\omega) = \sum_{\tau=-\infty}^{+\infty} R_s(\tau) e^{-i\tau\omega}, \quad \omega \in \mathbb{R}$$

IF $s(t)$ IS SCALAR (AS WE HEREBY ASSUME),

THEN:

$$1) \bar{\Phi}_s(\omega) \in \mathbb{R} \quad \forall \omega$$

$$2) \bar{\Phi}_s(\omega) \geq 0 \quad \forall \omega$$

$$3) \bar{\Phi}_s(\omega) = \bar{\Phi}_s(-\omega)$$

OFTEN, WE REASON IN TERMS OF

z -TRANSFORM AND

$$\bar{\Phi}_s(z) = \sum_{\tau=-\infty}^{+\infty} R_s(\tau) z^{-\tau}, \quad z \in \mathbb{C}$$

SO THAT THE POWER SPECTRUM IS GIVEN BY ITS EVALUATION ON THE UNIT CIRCLE

SO THAT THE POWER SPECTRUM IS GIVEN BY ITS EVALUATION ON THE UNIT CIRCLE

ALSO, WE OFTEN CONSIDER

$$y(t) = G(z)u(t) + H(z)e(t) \quad \text{⊗}$$

TO INDICATE THAT y IS THE SUM OF THE OUTPUTS OF TWO LINEAR SYSTEMS WITH TRANSFER FUNCTIONS $G(z)$, $H(z)$ FED RESPECTIVELY WITH INPUTS $u(t)$, $e(t)$

THEOREM: LET $y(t)$ BE GIVEN BY ⊗

WITH u A STATIONARY STOCHASTIC PROCESS WITH SPECTRUM $\Phi_u(\omega)$

WHILE e IS WHITE NOISE OF VARIANCE λ . THEN, IF G AND H

ARE STABLE, y IS STATIONARY WITH

$$\Phi_y(\omega) = |G(e^{j\omega})|^2 \Phi_u(\omega) + \lambda |H(e^{j\omega})|^2$$

SPECTRAL FACTORIZATION

LET $\bar{\Phi}(\omega)$ BE RATIONAL IN $e^{j\omega}$
WITH $\bar{\Phi}(\omega) > 0 \forall \omega$.

THEN, THERE EXISTS A MONIC
RATIONAL TRANSFER FUNCTION
OF z , $R(z)$, WITH POLES AND
ZEROS ALL INSIDE THE UNIT CIRCLE
S.T.

$$\bar{\Phi}(\omega) = \lambda |R(e^{j\omega})|^2 \quad \blacksquare$$

THIS MEANS THAT

A STATIONARY DISTURBANCE

v CAN BE WRITTEN AS

$$v(t) = R(z)e(t),$$

WITH $R(z)$ BEING

MINIMUM PHASE AND e WHITE

NOISE OF VARIANCE λ ,

WITH $R(z)$ BEING
MINIMUM PHASE AND e WHITE
NOISE OF VARIANCE λ ,
FROM INFORMATION ABOUT ITS
SPECTRUM ONLY.

SUMMARY

$$y(t) = G(z)u(t) + H(z)e(t)$$

IS THE BASIC DESCRIPTION
OF A LINEAR SYSTEM S.T.

DISTURBANCES WITH e WHITE NOISE.

G AND H ARE RATIONAL TRANSFER FUNCTIONS:

$$G(z) = \sum_{k=1}^{\infty} g(k)z^{-k}$$

$$H(z) = 1 + \sum_{k=1}^{\infty} h(k)z^{-k}, \text{ AND}$$

G STABLE, I.E. $\sum_{k=1}^{\infty} |g(k)| < \infty$,

H STABLE AND MINIMUM PHASE

IMPORTANCE OF MINIMUM-PHASE

NOISE REPRESENTATION:

NOISE MODEL INVERTIBILITY

LET

$$v(t) = H(z) e(t) \quad (*)$$

$$= \sum_{k=0}^{\infty} h(k) e(t-k), \quad h(0) = 1$$

A CRUCIAL PROPERTY FOR PREDICTION

IS THAT $(*)$ SHOULD BE INVERTIBLE,

I.E. WE SHOULD BE ABLE TO COMPUTE

$e(t)$ FROM $\{v(s)\}_{s \leq t}$

(MORE SPECIFICALLY, CAUSALLY
INVERTIBLE).

IF $H(z)$ IS STABLE AND INVERSELY

STABLE (MINIMUM PHASE), WE

JUST HAVE

JUST HAVE

$$\begin{aligned} e(t) &= \tilde{H}(z) v(t) \\ &= \sum_{k=0}^{\infty} \tilde{h}(k) v(t-k) \end{aligned}$$

WITH

$$\tilde{H}(z) = \frac{1}{H(z)} =: H^{-1}(z)$$

SO, $H(z)$ MUST HAVE NO ZEROS ON OR OUTSIDE THE UNIT CIRCLE.

THIS WELL RELATES TO THE SPECTRAL FACTORIZATION RESULT WHICH SAYS THAT, FOR RATIONAL STRICTLY POSITIVE SPECTRA, WE CAN ALWAYS FIND A REPRESENTATION $H(z)$ SATISFYING THIS.

ONE-STEP AND K-STEP

AHEAD OUTPUT PREDICTION

IF

$$y(t) = G(z) w(t) + H(z) e(t)$$

G STABLE, H MINIMUM PHASE



ONE-STEP AND K-STEP

AHEAD OUTPUT PREDICTION

IF

$$y(t) = G(z)u(t) + H(z)e(t)$$

G STABLE, H MINIMUM PHASE

IT TURNS OUT THAT

$$\hat{y}(t|t-1) = H^{-1}(z)G(z)u(t) + [1 - H^{-1}(z)]y(t)$$

IN ADDITION, LETTING

$$\bar{H}_k(z) = \sum_{i=0}^{k-1} h(i)z^{-i}$$

$$W_k(z) = \bar{H}_k(z)H^{-1}(z)$$

ONE HAS

$$\hat{y}(t|t-k) = W_k(z)G(z)u(t) + [1 - W_k(z)]y(t)$$

A New Approach to Linear Filtering and Prediction Problems¹

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The classical filtering and prediction problem is re-examined using the Bode-Shannon representation of random processes and the "state transition" method of analysis of dynamic systems. New results are:

(1) *The formulation and methods of solution of the problem apply without modification to stationary and nonstationary statistics and to growing-memory and infinite-memory filters.*

(2) *A nonlinear difference (or differential) equation is derived for the covariance matrix of the optimal estimation error. From the solution of this equation the coefficients of the difference (or differential) equation of the optimal linear filter are obtained without further calculations.*

(3) *The filtering problem is shown to be the dual of the noise-free regulator problem.*

The new method developed here is applied to two well-known problems, confirming and extending earlier results.

The discussion is largely self-contained and proceeds from first principles; basic concepts of the theory of random processes are reviewed in the Appendix.

Introduction

AN IMPORTANT class of theoretical and practical problems in communication and control is of a statistical nature. Such problems are: (i) Prediction of random signals; (ii) separation of random signals from random noise; (iii) detection of signals of known form (pulses, sinusoids) in the presence of random noise.

In his pioneering work, Wiener [1]³ showed that problems (i) and (ii) lead to the so-called Wiener-Hopf integral equation; he also gave a method (spectral factorization) for the solution of this integral equation in the practically important special case of stationary statistics and rational spectra.

Many extensions and generalizations followed Wiener's basic work. Zadeh and Ragazzini solved the finite-memory case [2]. Concurrently and independently of Bode and Shannon [3], they also gave a simplified method [2] of solution. Booton discussed the nonstationary Wiener-Hopf equation [4]. These results are now in standard texts [5-6]. A somewhat different approach along these main lines has been given recently by Darlington [7]. For extensions to sampled signals, see, e.g., Franklin [8], Lees [9]. Another approach based on the eigenfunctions of the Wiener-Hopf equation (which applies also to nonstationary problems whereas the preceding methods in general don't), has been pioneered by Davis [10] and applied by many others, e.g., Shinbrot [11], Blum [12], Pugachev [13], Solodovnikov [14].

In all these works, the objective is to obtain the specification of a linear dynamic system (Wiener filter) which accomplishes the prediction, separation, or detection of a random signal.⁴

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³ Numbers in brackets designate References at end of paper.

⁴ Of course, in general these tasks may be done better by nonlinear filters. At present, however, little or nothing is known about how to obtain (both theoretically and practically) these nonlinear filters.

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NOTE: Statements and opinions advanced in papers are to be understood as individual expressions of their authors and not those of the Society. Manuscript received at ASME Headquarters, February 24, 1959. Paper No. 59-IRD-11.

Present methods for solving the Wiener problem are subject to a number of limitations which seriously curtail their practical usefulness:

(1) The optimal filter is specified by its impulse response. It is not a simple task to synthesize the filter from such data.

(2) Numerical determination of the optimal impulse response is often quite involved and poorly suited to machine computation. The situation gets rapidly worse with increasing complexity of the problem.

(3) Important generalizations (e.g., growing-memory filters, nonstationary prediction) require new derivations, frequently of considerable difficulty to the nonspecialist.

(4) The mathematics of the derivations are not transparent. Fundamental assumptions and their consequences tend to be obscured.

This paper introduces a new look at this whole assemblage of problems, sidestepping the difficulties just mentioned. The following are the highlights of the paper:

(5) *Optimal Estimates and Orthogonal Projections.* The Wiener problem is approached from the point of view of conditional distributions and expectations. In this way, basic facts of the Wiener theory are quickly obtained; the scope of the results and the fundamental assumptions appear clearly. It is seen that all statistical calculations and results are based on first and second order averages; no other statistical data are needed. Thus difficulty (4) is eliminated. This method is well known in probability theory (see pp. 75-78 and 148-155 of Doob [15] and pp. 455-464 of Loève [16]) but has not yet been used extensively in engineering.

(6) *Models for Random Processes.* Following, in particular, Bode and Shannon [3], arbitrary random signals are represented (up to second order average statistical properties) as the output of a linear dynamic system excited by independent or uncorrelated random signals ("white noise"). This is a standard trick in the engineering applications of the Wiener theory [2-7]. The approach taken here differs from the conventional one only in the way in which linear dynamic systems are described. We shall emphasize the concepts of *state* and *state transition*; in other words, linear systems will be specified by systems of first-order difference (or differential) equations. This point of view is

natural and also necessary in order to take advantage of the simplifications mentioned under (5).

(7) *Solution of the Wiener Problem.* With the state-transition method, a single derivation covers a large variety of problems: growing and infinite memory filters, stationary and nonstationary statistics, etc.; difficulty (3) disappears. Having guessed the "state" of the estimation (i.e., filtering or prediction) problem correctly, one is led to a nonlinear difference (or differential) equation for the covariance matrix of the optimal estimation error. This is vaguely analogous to the Wiener-Hopf equation. Solution of the equation for the covariance matrix starts at the time t_0 when the first observation is taken; at each later time t the solution of the equation represents the covariance of the optimal prediction error given observations in the interval (t_0, t) . From the covariance matrix at time t we obtain at once, without further calculations, the coefficients (in general, time-varying) characterizing the optimal linear filter.

(8) *The Dual Problem.* The new formulation of the Wiener problem brings it into contact with the growing new theory of control systems based on the "state" point of view [17-24]. It turns out, *surprisingly*, that the Wiener problem is the *dual* of the noise-free optimal regulator problem, which has been solved previously by the author, using the state-transition method to great advantage [18, 23, 24]. The mathematical background of the two problems is identical—this has been suspected all along, but until now the analogies have never been made explicit.

(9) *Applications.* The power of the new method is most apparent in theoretical investigations and in numerical answers to complex practical problems. In the latter case, it is best to resort to machine computation. Examples of this type will be discussed later. To provide some feel for applications, two standard examples from nonstationary prediction are included; in these cases the solution of the nonlinear difference equation mentioned under (7) above can be obtained even in closed form.

For easy reference, the main results are displayed in the form of theorems. Only Theorems 3 and 4 are original. The next section and the Appendix serve mainly to review well-known material in a form suitable for the present purposes.

Notation Conventions

Throughout the paper, we shall deal mainly with *discrete* (or *sampled*) dynamic systems; in other words, signals will be observed at equally spaced points in time (*sampling instants*). By suitable choice of the time scale, the constant intervals between successive sampling instants (*sampling periods*) may be chosen as unity. Thus variables referring to time, such as t, t_0, τ, T will always be integers. The restriction to discrete dynamic systems is not at all essential (at least from the engineering point of view); by using the discreteness, however, we can keep the mathematics rigorous and yet elementary. Vectors will be denoted by small bold-face letters: $\mathbf{a}, \mathbf{b}, \dots, \mathbf{u}, \mathbf{x}, \mathbf{y}, \dots$. A *vector* or more precisely an *n-vector* is a set of n numbers x_1, \dots, x_n ; the x_i are the *co-ordinates* or *components* of the vector \mathbf{x} .

Matrices will be denoted by capital bold-face letters: $\mathbf{A}, \mathbf{B}, \mathbf{Q}, \Phi, \Psi, \dots$; they are $m \times n$ arrays of elements $a_{ij}, b_{ij}, q_{ij}, \dots$. The *transpose* (interchanging rows and columns) of a matrix will be denoted by the prime. In manipulating formulas, it will be convenient to regard a vector as a matrix with a single column.

Using the conventional definition of matrix multiplication, we write the *scalar product* of two n -vectors \mathbf{x}, \mathbf{y} as

$$\mathbf{x}'\mathbf{y} = \sum_{i=1}^n x_i y_i = \mathbf{y}'\mathbf{x}$$

The scalar product is clearly a scalar, i.e., not a vector, quantity.

Similarly, the quadratic form associated with the $n \times n$ matrix \mathbf{Q} is,

$$\mathbf{x}'\mathbf{Q}\mathbf{x} = \sum_{i,j=1}^n x_i q_{ij} x_j$$

We define the expression $\mathbf{x}\mathbf{y}'$ where \mathbf{x}' is an m -vector and \mathbf{y} is an n -vector to be the $m \times n$ matrix with elements $x_i y_j$.

We write $E(\mathbf{x}) = E\mathbf{x}$ for the expected value of the random vector \mathbf{x} (see Appendix). It is usually convenient to omit the brackets after E . This does not result in confusion in simple cases since constants and the operator E commute. Thus $E\mathbf{x}\mathbf{y}' =$ matrix with elements $E(x_i y_j)$; $E\mathbf{x}E\mathbf{y}' =$ matrix with elements $E(x_i)E(y_j)$.

For ease of reference, a list of the principal symbols used is given below.

Optimal Estimates

t	time in general, present time.
t_0	time at which observations start.
$x_1(t), x_2(t)$	basic random variables.
$y(t)$	observed random variable.
$x_1^*(t_1 t)$	optimal estimate of $x_1(t_1)$ given $y(t_0), \dots, y(t)$.
L	loss function (non random function of its argument).
ϵ	estimation error (random variable).

Orthogonal Projections

$\mathcal{Y}(t)$	linear manifold generated by the random variables $y(t_0), \dots, y(t)$.
$\bar{x}(t_1 t)$	orthogonal projection of $x(t_1)$ on $\mathcal{Y}(t)$.
$\tilde{x}(t_1 t)$	component of $x(t_1)$ orthogonal to $\mathcal{Y}(t)$.

Models for Random Processes

$\Phi(t+1; t)$	transition matrix
$\mathbf{Q}(t)$	covariance of random excitation

Solution of the Wiener Problem

$\mathbf{x}(t)$	basic random variable.
$\mathbf{y}(t)$	observed random variable.
$\mathcal{Y}(t)$	linear manifold generated by $\mathbf{y}(t_0), \dots, \mathbf{y}(t)$.
$\mathcal{Z}(t)$	linear manifold generated by $\mathbf{y}(t_0), \dots, \mathbf{y}(t-1)$.
$\mathbf{x}^*(t_1 t)$	optimal estimate of $\mathbf{x}(t_1)$ given $\mathcal{Y}(t)$.
$\tilde{\mathbf{x}}(t_1 t)$	error in optimal estimate of $\mathbf{x}(t_1)$ given $\mathcal{Y}(t)$.

Optimal Estimates

To have a concrete description of the type of problems to be studied, consider the following situation. We are given signal $x_1(t)$ and noise $x_2(t)$. Only the sum $y(t) = x_1(t) + x_2(t)$ can be observed. Suppose we have observed and know exactly the values of $y(t_0), \dots, y(t)$. What can we infer from this knowledge in regard to the (unobservable) value of the signal at $t = t_1$, where t_1 may be less than, equal to, or greater than t ? If $t_1 < t$, this is a *data-smoothing* (*interpolation*) problem. If $t_1 = t$, this is called *filtering*. If $t_1 > t$, we have a *prediction* problem. Since our treatment will be general enough to include these and similar problems, we shall use hereafter the collective term *estimation*.

As was pointed out by Wiener [1], the natural setting of the estimation problem belongs to the realm of probability theory and statistics. Thus signal, noise, and their sum will be random variables, and consequently they may be regarded as random processes. From the probabilistic description of the random processes we can determine the probability with which a particular sample of the signal and noise will occur. For any given set of measured values $\eta(t_0), \dots, \eta(t)$ of the random variable $y(t)$ one can then also determine, in principle, the probability of simultaneous occurrence of various values $\xi_1(t)$ of the random variable $x_1(t)$. This is the conditional probability distribution function

New Results in Linear Filtering and Prediction Theory¹

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A nonlinear differential equation of the Riccati type is derived for the covariance matrix of the optimal filtering error. The solution of this "variance equation" completely specifies the optimal filter for either finite or infinite smoothing intervals and stationary or nonstationary statistics.

The variance equation is closely related to the Hamiltonian (canonical) differential equations of the calculus of variations. Analytic solutions are available in some cases. The significance of the variance equation is illustrated by examples which duplicate, simplify, or extend earlier results in this field.

The Duality Principle relating stochastic estimation and deterministic control problems plays an important role in the proof of theoretical results. In several examples, the estimation problem and its dual are discussed side-by-side.

Properties of the variance equation are of great interest in the theory of adaptive systems. Some aspects of this are considered briefly.

1 Introduction

AT PRESENT, a nonspecialist might well regard the Wiener-Kolmogorov theory of filtering and prediction [1, 2]³ as "classical"—in short, a field where the techniques are well established and only minor improvements and generalizations can be expected.

That this is not really so can be seen convincingly from recent results of Shinbrot [3], Steeg [4], Pugachev [5, 6], and Parzen [7]. Using a variety of time-domain methods, these investigators have solved some long-standing problems in *nonstationary* filtering and prediction theory. We present here a unified account of our own independent researches during the past two years (which overlap with much of the work [3-7] just mentioned), as well as numerous new results. We, too, use time-domain methods, and obtain major improvements and generalizations of the conventional Wiener theory. In particular, our methods apply without modification to multivariate problems.

The following is the historical background of this paper.

In an extension of the standard Wiener filtering problem, Follin [8] obtained relationships between time-varying gains and error variances for a given circuit configuration. Later, Hanson [9] proved that Follin's circuit configuration was actually optimal for the assumed statistics; moreover, he showed that the differential equations for the error variance (first obtained by Follin) follow rigorously from the Wiener-Hopf equation. These results were then generalized by Bucy [10], who found explicit relationships between the optimal weighting functions and the error variances; he also gave a rigorous derivation of the variance equations and those of the optimal filter for a wide class of nonstationary signal and noise statistics.

Independently of the work just mentioned, Kalman [11] gave

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³ Numbers in brackets designate References at the end of paper.

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a new approach to the standard filtering and prediction problem. The novelty consisted in combining two well-known ideas:

- (i) the "state-transition" method of describing dynamical systems [12-14], and
- (ii) linear filtering regarded as orthogonal projection in Hilbert space [15, pp. 150-155].

As an important by-product, this approach yielded the *Duality Principle* [11, 16] which provides a link between (stochastic) filtering theory and (deterministic) control theory. Because of the duality, results on the optimal design of linear control systems [13, 16, 17] are directly applicable to the Wiener problem. Duality plays an important role in this paper also.

When the authors became aware of each other's work, it was soon realized that the principal conclusion of both investigations was identical, in spite of the difference in methods:

Rather than to attack the Wiener-Hopf integral equation directly, it is better to convert it into a nonlinear differential equation, whose solution yields the covariance matrix of the minimum filtering error, which in turn contains all necessary information for the design of the optimal filter.

2 Summary of Results: Description

The problem considered in this paper is stated precisely in Section 4. There are two main assumptions:

(A₁) A sufficiently accurate model of the message process is given by a linear (possibly time-varying) dynamical system excited by white noise.

(A₂) Every observed signal contains an additive white noise component.

Assumption (A₂) is unnecessary when the random processes in question are sampled (discrete-time parameter); see [11]. Even in the continuous-time case, (A₂) is no real restriction since it can be removed in various ways as will be shown in a future paper. Assumption (A₁), however, is quite basic; it is analogous to but somewhat less restrictive than the assumption of rational spectra in the conventional theory.

Within these assumptions, we seek the best linear estimate of the message based on past data lying in either a finite or infinite time-interval.

The fundamental relations of our new approach consist of five equations:

$$\mathcal{E}\{\mu(t) - \hat{\mu}(t)\}^2 \geq \|\mathbf{x}^*\|^2_{\mathbf{M}^{-1}(t_0, t)} \quad (32)$$

Every costate \mathbf{x}^* has a minimum-variance unbiased estimator for which the equality sign holds in (32) if and only if \mathbf{M} is positive definite. This motivates the use of condition (A₄') in Theorem 3 and the term "completely observable."

(i) It can be shown [17] that in the constant case complete observability is equivalent to the easily verified condition:

$$\text{rank}[\mathbf{H}', \mathbf{F}'\mathbf{H}', \dots, (\mathbf{F}')^{n-1}\mathbf{H}'] = n \quad (33)$$

where the square brackets denote a matrix with n rows and np columns.

(9) *Stability of the optimal filter.* It should be realized now that the optimality of the filter (I) does not at the same time guarantee its stability. The reader can easily check this by constructing an example (for instance, one in which (10-11) consists of two non-interacting systems). To establish weak sufficient conditions for stability entails some rather delicate mathematical technicalities which we shall bypass and state only the best final result currently available.

First, some additional definitions.

We say that the model (10-11) is *uniformly completely observable* if there exist fixed constants, α_1, α_2 , and σ such that

$$\alpha_1 \|\mathbf{x}^*\|^2 \leq \|\mathbf{x}^*\|^2_{\mathbf{M}(t-\sigma, t)} \leq \alpha_2 \|\mathbf{x}^*\|^2 \quad \text{for all } \mathbf{x}^* \text{ and } t.$$

Similarly, we say that a model is *completely controllable* [uniformly completely controllable] if the dual model is completely observable [uniformly completely observable]. For a discussion of these notions, the reader may refer to [17]. It should be noted that the property of "uniformity" is always true for constant systems.

We can now state the central theorem of the paper:

THEOREM 4. Assume that the model of the message process is

- (A₄') uniformly completely observable;
- (A₅') uniformly completely controllable;
- (A₆') $\alpha_3 \leq \|\mathbf{Q}(t)\| \leq \alpha_4, \quad \alpha_5 \leq \|\mathbf{R}(t)\| \leq \alpha_6$ for all t ;
- (A₇') $\|\mathbf{F}(t)\| \leq \alpha_7$.

Then the following is true:

- (i) The optimal filter is uniformly asymptotically stable;
- (ii) Every solution $\mathbf{II}(t; \mathbf{P}_0, t_0)$ of the variance equation (IV) starting at a symmetric nonnegative matrix \mathbf{P}_0 converges to $\bar{\mathbf{P}}(t)$ (defined in Theorem 3) as $t \rightarrow \infty$.

Remarks. (j) A filter which is not uniformly asymptotically stable may have an unbounded response to a bounded input [21]; the practical usefulness of such a filter is rather limited.

(k) Property (ii) in Theorem 4 is of central importance since it shows that the variance equation is a "stable" computational method that may be expected to be rather insensitive to roundoff errors.

(l) The speed of convergence of $\mathbf{P}_0(t)$ to $\bar{\mathbf{P}}(t)$ can be estimated quite effectively using the second method of Lyapunov; see [17].

(10) *Solution of the classical Wiener problem.* Theorems 3 and 4 have the following immediate corollary:

THEOREM 5. Assume the hypotheses of Theorems 3 and 4 are satisfied and that $\mathbf{F}, \mathbf{G}, \mathbf{H}, \mathbf{Q}, \mathbf{R}$, are constants.

Then, if $t_0 = -\infty$, the solution of the estimation problem is obtained by setting the right-hand side of (IV) equal to zero and solving the resulting set of quadratic algebraic equations. That solution which is nonnegative definite is equal to $\bar{\mathbf{P}}$.

To prove this, we observe that, by the assumption of constancy, $\bar{\mathbf{P}}(t)$ is a constant. By Theorem 4, all solutions of (IV) starting at nonnegative matrices converge to $\bar{\mathbf{P}}$. Hence, if a matrix \mathbf{P} is found for which the right-hand side of (IV) vanishes and if this matrix is nonnegative definite, it must be identical

with $\bar{\mathbf{P}}$. Note, however, that the procedure may fail if the conditions of Theorems 3 and 4 are not satisfied. See Example 4.

(11) *Solution of the Dual Problem.* For details, consult [17]. The only facts needed here are the following: The optimal control law is given by

$$\mathbf{u}^*(t^*) = -\mathbf{K}^*(t^*)\mathbf{x}(t^*) \quad (34)$$

where $\mathbf{K}^*(t^*)$ satisfies the duality relation

$$\mathbf{K}^*(t^*) = \mathbf{K}'(t) \quad (35)$$

and is to be determined by duality from formula (III). The value of the performance index (20) may be written in the form

$$\min_{\mathbf{u}^*} V(\mathbf{x}^*; t^*, t_0^*, \mathbf{u}^*) = \|\mathbf{x}^*\|^2_{\mathbf{II}^*(t^*; \mathbf{x}^*, t_0^*)}$$

where $\mathbf{II}^*(t^*; \mathbf{x}^*, t_0^*)$ is the solution of the dual of the variance equation (IV).

It should be carefully noted that the hypotheses of Theorem 4 are invariant under duality. Hence essentially the same theory covers both the estimation and the regular problem, as stated in Section 5.

The vector-matrix block diagram for the optimal regulator is shown in Fig. 11.

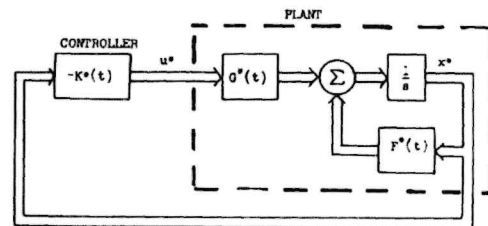


Fig. 11 General block diagram of optimal regulator

(12) *Computation of the covariance matrix for the message process.* To apply Theorem 1, it is necessary to determine $\text{cov}[\mathbf{x}(t), \mathbf{x}(t_0)]$. This may be specified as part of the problem statement as in Example 5. On the other hand, one might assume that the message model has reached steady state (see (A₃)), in which case from (13) and (12) we have that

$$\mathbf{S}(t) = \text{cov}[\mathbf{x}(t), \mathbf{x}(t)] = \int_{-\infty}^t \Phi(t, \tau) \mathbf{G}(\tau) \mathbf{Q}(\tau) \mathbf{G}'(\tau) \Phi'(t, \tau) d\tau$$

provided the model (10) is asymptotically stable. Differentiating this expression with respect to t we obtain the following differential equation for $\mathbf{S}(t)$

$$d\mathbf{S}/dt = \mathbf{F}(t)\mathbf{S} + \mathbf{S}\mathbf{F}'(t) + \mathbf{G}(t)\mathbf{Q}(t)\mathbf{G}'(t) \quad (36)$$

This formula is analogous to the well-known lemma of Lyapunov [21] in evaluating the integrated square of a solution of a linear differential equation. In case of a constant system, (36) reduces to a system of linear algebraic equations.

8 Derivation of the Fundamental Equations

We first deduce the matrix form of the familiar Wiener-Hopf integral equation. Differentiating it with respect to time and then using (10-11), we obtain in a very simple way the fundamental equations of our theory.

Much cumbersome manipulation of integrals can be avoided by recognizing, as has been pointed out by Pugachev [27], that the Wiener-Hopf equation is a special case of a simple geometric principle: *orthogonal projection*.

Consider an abstract space \mathcal{X} such that an inner product (X, Y) is defined between any two elements X, Y of \mathcal{X} . The norm is defined by $\|X\| = (X, X)^{1/2}$. Let \mathcal{U} be a subspace of \mathcal{X} . We

FROM WIENER TO KALMAN

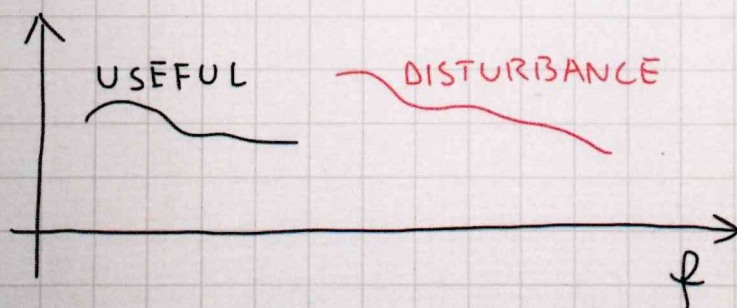
IMPORTANCE OF WIENER: TO SEE THE SIGNALS OF INTEREST AS STOCHASTIC PROCESSES \Rightarrow STATISTICAL SIGNAL PROCESSING

MORE POWERFUL THAN DETERMINISTIC,
MANY TIMES WE CAN MEASURE

(SIGNAL OF INTEREST) + (DISTURBANCE NOISE)

AND THESE TWO SIGNALS DO NOT LIVE ON DIFFERENT FREQUENCIES

IF ONE HAD



JUST A LOW-PASS FILTER IS NEEDED

BUT ONE COULD WELL HAVE



IF STATISTICAL PROPERTIES OF THE SIGNALS ARE KNOWN WE CAN STILL OBTAIN A GOOD ESTIMATE OF THE SIGNAL OF INTEREST

LIMITATIONS OF THE WIENER APPROACH

WE HAVE SAID THAT IT RELIES ON STATIONARY ASSUMPTIONS.

EVEN IF IT CAN ALSO (IN PRINCIPLE) HANDLE NON RATIONAL SPECTRA, STATIONARITY IS AN ASSUMPTION OFTEN VIOLATED IN REAL APPLICATIONS

IN ROBOTICS, BIOMEDICINE, BIOINFORMATICS STRONGLY NONSTATIONARY SIGNALS ARISE.

IN INDUSTRIAL PROCESSES LOAD DISTURBANCES ARE IMPULSIVE INPUTS TO THE SYSTEM.

MEASUREMENT NOISES CAN BE OF THE FORM

$$e_i \sim N(0, \sigma_i^2), \text{ E.G. } \sigma_i^2 \sim (CV \cdot y_i)^2$$

WIENER CAN ALSO HANDLE

∞ -DIM LINEAR SYSTEMS

BUT THE FORMULA TO OBTAIN THE

FILTER IN PRACTICE CAN BE USED

ONLY WHEN SPECTRA ARE RATIONAL

FILTER IN PRACTICE CAN BE USED ONLY WHEN SPECTRA ARE RATIONAL, AND TO HAVE A FINITE-MEMORY FILTER WE NEED TO START FROM RATIONAL SPECTRA

KALMAN

SINCE RATIONAL SPECTRA ARE GENERATED BY FINITE-DIMENSIONAL LINEAR SYSTEMS, THE FUNDAMENTAL IDEA OF KALMAN WAS TO START USING SUCH SYSTEMS AND IN STATE SPACE FORM

VERY STRONG CONSEQUENCES

- THE PROBLEM OF THE EXPLICIT CALCULATION OF THE TRANSFER FUNCTION OF THE FILTER DISAPPEARS.
THE ESTIMATOR HAS A KNOWN ALGORITHMIC STRUCTURE: FINITE-DIMENSIONAL LINEAR SYSTEM IN STATE SPACE DRIVEN BY THE OBSERVATIONS (REACHING MOON WITH NO POWERFUL COMPUTERS)
- NO NEED TO ASSUME TIME-INVARIANT AND STABLE SYSTEMS AC

- NO NEED TO ASSUME TIME-INVARIANT AND STABLE SYSTEMS AS SIGNAL GENERATORS
- NO NEED TO START FROM $t_0 = -\infty$ UNCERTAINTY ON SYSTEM INITIAL CONDITION CAN BE HANDLED
- CALCULATIONS MADE VIA PROJECTIONS ONTO FINITE-DIMENSIONAL SUBSPACES OF RANDOM VARIABLES
("SIMPLER" BUT MORE POWERFUL THEORY)
- DUALITY WITH OPTIMAL CONTROL

FINITE-DIMENSIONAL STOCHASTIC STATE-SPACE MODEL IN DISCRETE-TIME

$$\Sigma \begin{cases} x(t+1) = A(t)x(t) + B(t)n(t), & t \geq t_0 \\ y(t) = C(t)x(t) + D(t)n(t) \\ x(t_0) = x_0 \end{cases}$$

- $n(t)$ IS μ -DIMENSIONAL WHITE NOISE OF UNIT VARIANCE

$$E[n(t)n^T(s)] = I_\mu \delta(t-s)$$

FINITE-DIMENSIONAL STOCHASTIC STATE-SPACE MODEL IN DISCRETE-TIME

$$\Sigma \begin{cases} x(t+1) = A(t)x(t) + B(t)n(t), & t \geq t_0 \\ y(t) = C(t)x(t) + D(t)n(t) \\ x(t_0) = x_0 \end{cases}$$

- $n(t)$ IS μ -DIMENSIONAL WHITE NOISE
OF UNIT VARIANCE

$$E[n(t)n^T(s)] = I_\mu \delta(t-s)$$

$$E[n(t)] = 0 \quad \forall t$$

- $\{x(t)\}$ ARE THE STATES,

n -DIM. RANDOM VECTORS

- x_0 IS THE INITIAL STATE, A
RANDOM VECTOR

$$E x_0 = \mu_0, \quad \Rightarrow x_0 \sim (\mu_0, \Sigma_0)$$

$$\text{VAR } x_0 = \Sigma_0$$

IT IS UNCORRELATED FROM $n(t)$, I.E.

$$\text{COV}(x_0, n(t)) = E[(x_0 - \mu_0)n^T(t)] = 0_{n \times \mu}$$

$\forall t$

- $\{y(t)\}$ ARE THE m -DIM OUTPUTS,
THEY ARE RANDOM VECTORS

DEFINITION:

A PROCESS $\{y(t)\}$ IS SAID TO HAVE
FINITE DIMENSION IF IT ADMITS
A REPRESENTATION WITH THE Σ
ABOVE. Σ CAN BE NON UNIQUE,
ANY Σ THAT REPRESENTS $\{y(t)\}$ IS
SAID A REALIZATION OF $\{y(t)\}$

ASSUMPTION ONLY DONE TO
SIMPLIFY NOTATION

A, B, C, D DO NOT DEPEND ON t

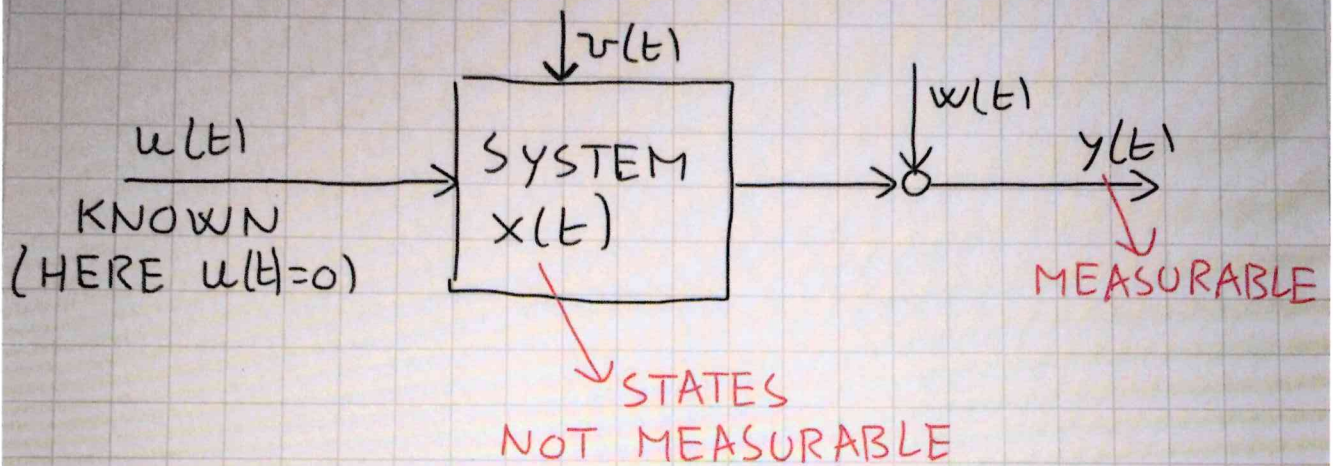
$$A \in \mathbb{R}^{n \times n}, B \in \mathbb{R}^{n \times r}, C \in \mathbb{R}^{m \times n}, \\ D \in \mathbb{R}^{m \times r}$$

FUNDAMENTAL ENGINEERING
PARADIGM FOR MANY
APPLICATIONS

$$y(t) = B u(t) + D u(t) + C x(t)$$

FUNDAMENTAL ENGINEERING PARADIGM FOR MANY APPLICATIONS

$$v(t) := B n(t), \quad w(t) := D n(t)$$



WE WANT TO PREDICT FUTURE
OUTPUTS OR STATES ACCOUNTING
FOR UNEXPECTED EVENTS DESCRIBED
BY DISTURBANCES $v(t)$
(VERY HARD TO CAST THEM
AS DETERMINISTIC
SYSTEM INPUTS)

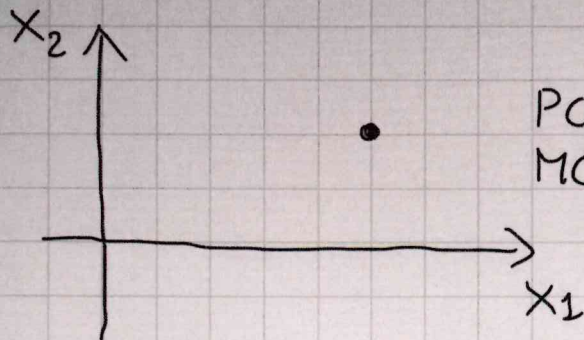
NOTE: MANY TIMES $v(t) \perp w(s)$

$\forall t, s$

EXAMPLE:

EXAMPLE :

ROBOT LOCALIZATION



POSITION OF A
MOBILE AGENT ON
THE PLANE

$$x_1(t+1) = x_1(t) + u_1(t) + v_1(t)$$

$$x_2(t+1) = x_2(t) + u_2(t) + v_2(t)$$

KNOWN
INPUTS
TO MAKE
THE ROBOT
MOVE

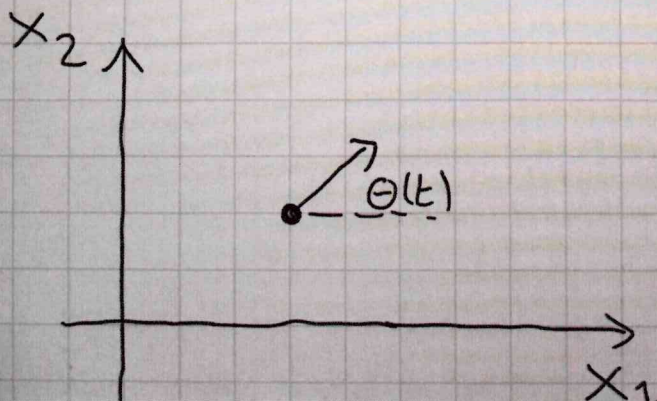
OBSTACLES,
WALLS, ...

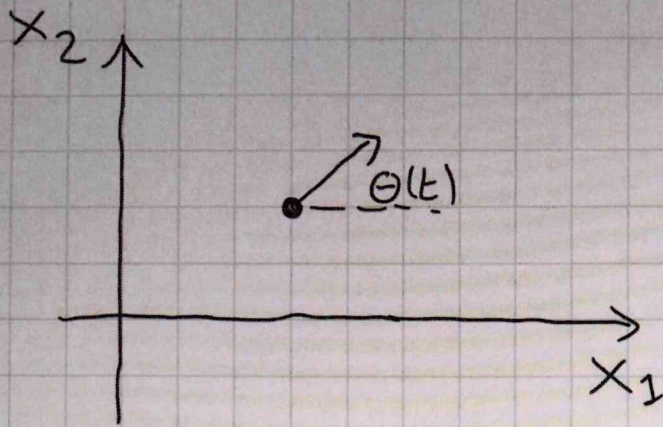
$$y_1(t) = x_1(t) + w_1(t)$$

$$y_2(t) = x_2(t) + w_2(t)$$

LINEAR

POSE ESTIMATION (BEARING)





$T(t)$ = TRANSLATION SPEED

$\omega(t)$ = ROTATION SPEED

$$x_1(t+1) = x_1(t) + \frac{T(t)}{\omega(t)} \cdot \sin(\theta(t) + \omega(t))$$

$$- \frac{T(t)}{\omega(t)} \cdot \sin(\theta(t)) + v_1(t)$$

$$x_2(t+1) = x_2(t) + \frac{T(t)}{\omega(t)} \cdot \cos \theta(t)$$

$$- \frac{T(t)}{\omega(t)} \cdot \cos(\theta(t) + \omega(t)) + v_2(t)$$

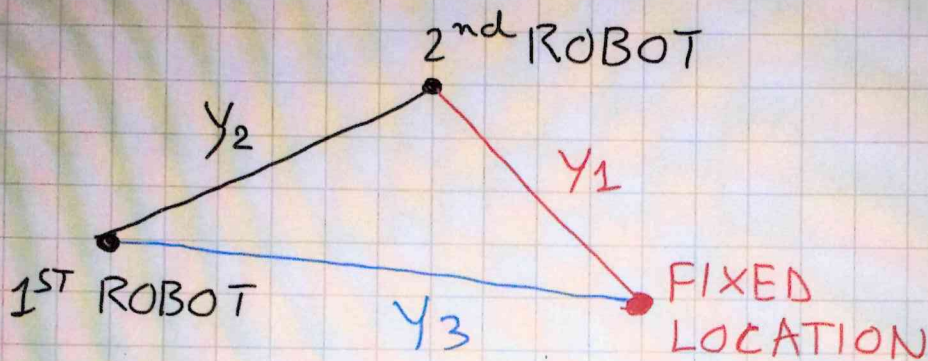
$$\theta(t+1) = \theta(t) + \omega(t)$$

NONLINEAR

THEN GROUP OF ROBOTS CAN
BE CONSIDERED WITH ALTERNATE

NONLINEAR

THEN GROUP OF ROBOTS CAN BE CONSIDERED, WITH DISTANCE MEASUREMENTS



RATIONAL TRANSFER

FUNCTIONS: REALIZATION

THEORY

LET US RECALL WHICH I/O RELATIONS

$\Sigma = (A, B, C, D)$ CAN DESCRIBE

(THINK $w(t)$ AS THE DETERMINISTIC

INPUT $w(t)$ SEEN IN PREVIOUS

COURSES)

$$Y(z) = W_{\Sigma}(z) U(z)$$

RATIONAL TRANSFER FUNCTIONS: REALIZATION THEORY

LET US RECALL WHICH I/O RELATIONS
 $\Sigma = (A, B, C, D)$ CAN DESCRIBE
 (THINK $w(t)$ AS THE DETERMINISTIC
 INPUT $w(t)$ SEEN IN PREVIOUS
 COURSES)

$$Y(z) = W_{\Sigma}(z) U(z)$$

$$W_{\Sigma}(z) = C(zI - A)^{-1}B + D$$

↓
 MATRIX WITH ENTRIES GIVEN BY RATIONAL
 TRANSFER FUNCTIONS

LET A BE STABLE ($|x_i| < 1 \forall i$)

IF $w(t)$ IS STOCHASTIC, IN PARTICULAR

$$w(t) = \text{WHITE NOISE } n(t) \\ \text{OF UNIT VARIANCE}$$

AND $t_0 = -\infty$ (THE INPUT STARTS
 IN THE REMOTE PAST)

THEN $y(t)$ IS A STATIONARY

THEN $y(t)$ IS A STATIONARY
STOCHASTIC PROCESS WITH SPECTRUM

$$S_y(z) = W_z(z) W_z^T(z^{-1})$$

NOW, CONSIDER THE INVERSE
PROBLEM:

GIVEN A

RTF $W(z)$

\Rightarrow

FIND A SYSTEM Σ
WHICH IS A
REALIZATION
OF $W(z)$

THEOREM: GIVEN A PROPER

$$\text{RTF } W(z) \left(= \frac{\text{POLYNOMIAL}}{\text{POLYNOMIAL}} = \frac{n(s)}{d(s)}, \right. \\ \left. \deg n(s) \leq \deg d(s) \right)$$

THERE ALWAYS EXISTS $\Sigma = (A, B, C, D)$
WHICH IS A REALIZATION OF $W(z)$.

SO, RATIONAL SPECTRA ARE
ASSOCIATED WITH FINITE-DIM.

STATE SPACE MODELS

SKETCH

$$W(z) = \bar{W}(z) + \lim_{z \rightarrow \infty} W(z)$$

$$= \underbrace{\bar{W}(z)}_{\text{STRICTLY PROPER}} + 1$$

AND E.G. IN THE SISO CASE, IF

$$\bar{W}(z) = \frac{b_0 + b_1 z + \dots + b_{n-1} z^{n-1}}{a_0 + a_1 z + \dots + z^n}$$

ONE REALIZATION E.G. IS

$$A = \begin{bmatrix} 0 & 1 & & \\ & & \ddots & \\ & & & 1 \\ -a_0 & -a_1 & \dots & -a_{n-1} \end{bmatrix}, B = \begin{bmatrix} 0 \\ 0 \\ \vdots \\ 0 \\ 1 \end{bmatrix},$$

$$C = [b_0 \ b_1 \ \dots \ b_{n-1}]$$

$$C = [c_0 \ c_1 \ \dots \ c_{n-1}]$$

THIS IS THE CANONICAL
CONTROL FORM

THE STATE PROCESS

$$x(t+1) = Ax(t) + Bn(t)$$

PROPOSITION:

$\{x(t)\}$ IS A MARKOV PROCESS
IN WEAK SENSE, I.E.

$$\begin{aligned}\hat{E}[x(t) \mid x_0, \dots, x_s] &= \hat{E}[x(t) \mid H_s(x)] \\ &= \hat{E}[x(t) \mid x(s)] \\ &\quad \forall t \geq s\end{aligned}$$

PROOF:

IT SUFFICES WRITING

PROOF:

IT SUFFICES WRITING

$$x(s+1) = Ax(s) + Bn(s)$$

$$\begin{aligned} x(s+2) &= A(Ax(s) + Bn(s)) + Bn(s+1) \\ &= A^2x(s) + [B \ AB] \begin{bmatrix} n(s+1) \\ n(s) \end{bmatrix} \end{aligned}$$

|
|
|

$$x(t) = \underbrace{A^{t-s}x(s)}_{\in H_s(x)} + \underbrace{\sum_{k=s}^{t-1} A^{t-1-k} Bn(k)}_{\perp x_0, \dots, x_s, \text{ i.e. } \perp H_s(x)}$$

$$= f_{H_s} + f_{H_s^\perp}$$

SO WE HAVE OBTAINED THE

UNIQUE DECOMPOSITION RELATED

UNIQUE DECOMPOSITION RELATED
TO THE PROJECTION THEOREM
AND THE PROJECTION OF $x(t)$
ONTO $H_S(x)$ IS $A^{t-s}x(s)$
PROVING THE THEOREM

NOTE: IN THE GAUSSIAN CASE
 $\{x(t)\}$ IS MARKOV SINCE IT ALSO
HOLDS THAT

$$p(x(t) | H_S(x)) = p(x(t) | x(s))$$

HOMEWORKS

PROVE THAT

$$a) \hat{E} [x(t) | H(x^s, y^{s-1})], t \geq s$$

||

$$\hat{E} [x(t) | x(s)]$$

($x^s =$ ALL THE STATES UNTIL
INSTANT s)

NOTE: IN THE GAUSSIAN CASE

$\{x(t)\}$ IS MARKOV SINCE IT ALSO
HOLDS THAT

$$p(x(t) | H_s(x)) = p(x(t) | x(s))$$

HOMEWORKS

PROVE THAT

$$a) \hat{E} [x(t) | H(x^s, y^{s-1})], \quad t \geq s$$

||

$$\hat{E} [x(t) | x(s)]$$

(x^s = ALL THE STATES UNTIL
INSTANT s)

$$b) \hat{E} [y(t) | H(x^s, y^{s-1})], \quad t \geq s$$

||

$$\hat{E} [y(t) | x(s)]$$

SOLUTION OF POINT a)

WE WANT TO PROVE THAT

$$\hat{E} [x(t) | H(x^s, y^{s-1})]$$

||

$$\hat{E} [x(t) | x(s)], \quad t \geq s$$

PRELIMINARY NOTE:

- WE ALREADY PROVED THAT
= $H(x^s)$ OR JUST x^s

$$\hat{E} [x(t) | H(x_0, \dots, x(s))] = \hat{E} [x(t) | x(s)]$$

$$- x(t_0) = x_0$$

$$x(t_0+1) = Ax_0 + Bn(t_0) \Rightarrow x(t_0+1) \text{ IS LINEARLY} \\ \text{GENERATED BY } x_0, n(t_0)$$

$$x(t_0+2) = Ax(t_0+1) + Bn(t_0+1) \Rightarrow x(t_0+2) \\ \text{IS LIN. GEN. BY } x(t_0), n(t_0), n(t_0+1)$$

⋮

$\Rightarrow x_0, x(t_0+1), \dots, x(s)$ ARE LINEARLY

GENERATED BY $x_0, n(t_0), \dots, n(s-1)$

||

GENERATED BY $x_0, n(t_0), \dots, n(s-1)$

\Downarrow

$$H(x^s) \subseteq H(x_0, n(t_0), \dots, n(s-1))$$

$$- y(t_0) = Cx_0 + Dn(t_0)$$

\vdots

$$y(s-1) = Cx(s-1) + Dn(s-1)$$

\Downarrow

$$H(y^{s-1}) \subseteq H(x_0, n(t_0), \dots, n(s-1))$$

\Downarrow

$$H(x^s, y^{s-1}) \subseteq H(x_0, n(t_0), \dots, n(s-1)) \quad (*)$$

NOW WE CAN PROVE a)

WE KNOW THAT

$$x(t) = \underbrace{A^{t-s}}_{\in H(x^s, y^{s-1})} x(s) + \underbrace{\sum_{k=s}^{t-1} A^{t-1-k} B n(k)}_{\subseteq H(n(s), \dots, n(t-1))}$$

$$\in H(x^s, y^{s-1})$$

$$\subseteq H(n(s), \dots, n(t-1))$$

\Downarrow

$$\perp H(x_0, n(t_0), \dots, n(s-1))$$

\Downarrow , USING (*)

$$\Downarrow$$

$$H(y^{s-1}) \subseteq H(x_0, n(t_0), \dots, n(s-1))$$

$$\Downarrow$$

$$H(x^s, y^{s-1}) \subseteq H(x_0, n(t_0), \dots, n(s-1)) \quad (*)$$

NOW WE CAN PROVE a)

WE KNOW THAT

$$x(t) = \underbrace{A^{t-s} x(s)}_{\in H(x^s, y^{s-1})} + \underbrace{\sum_{k=s}^{t-1} A^{t-1-k} B n(k)}_{\subseteq H(n(s), \dots, n(t-1))}$$

$$\subseteq H(n(s), \dots, n(t-1))$$

\Downarrow

$$\perp H(x_0, n(t_0), \dots, n(s-1))$$

\Downarrow USING (*)

$$\perp H(x^s, y^{s-1})$$

SO, BY THE PROJ. TH. ONE HAS

$$\hat{E}[x(t) | H(x^s, y^{s-1})] = A^{t-s} x(s)$$

$$\left(= \hat{E}[x(t) | H(x^s)] = \hat{E}[x(t) | x(s)] \right)$$

FIRST- AND SECOND-ORDER MOMENTS OF $\{x(t)\}, \{y(t)\}$

$$x(t+1) = Ax(t) + Bn(t)$$

$$y(t) = Cx(t) + Dn(t)$$

$$\left. \begin{array}{l} E x_0 = \mu_0 \\ \text{VAR } x_0 = \Sigma_0 \end{array} \right\} \begin{array}{l} \text{HOW DO THEY} \\ \text{CHANGE IN TIME?} \end{array}$$

MEANS

$$\mu_x(t+1) = A\mu_x(t), \quad \mu_x(t_0) = \mu_0$$

$$\mu_y(t) = C\mu_x(t)$$

COVARIANCE

$$\text{COV}(x(s+1), x(s)) =: \Sigma_x(s+1, s)$$

$$= \text{COV}(Ax(s) + Bn(s), x(s))$$

$$= A \text{COV}(x(s), x(s))$$

$$= A \text{VAR}(x(s)) = A\Sigma(s)$$

ONE ALSO OBTAINS

ONE ALSO OBTAINS

$$\Sigma_x(s, s+1) = \Sigma(s) A^T$$

AND IN GENERAL

$$\Sigma_x(t, s) = A^{t-s} \Sigma(s), t \geq s$$

$$\Sigma_x(t, s) = \Sigma(t) (A^T)^{s-t}, t \leq s$$

VARIANCE

$$\text{VAR } x(t+1) = \text{VAR} [Ax(t) + Bn(t)]$$

$$= A \text{VAR } x(t) A^T + B \text{VAR } n(t) B^T$$

$$= A \text{VAR } x(t) A^T + B B^T$$

$$\underbrace{\hspace{10em}}_{\text{!!} \Sigma(t)}$$

⇓

$$\Sigma(t+1) = A \Sigma(t) A^T + B B^T$$

$$\Sigma(t_0) = \Sigma_0$$

STATE-SPACE MODELS AND STATIONARY PROCESSES

$$x(t+1) = Ax(t) + Bn(t), \quad v(t) := Bn(t)$$

$$x(t_0) = x_0 \sim (\mu_0, \Sigma_0), \quad \text{VAR } v(t) = BB^T \\ = Q$$

PROPOSITION:

IF A IS STABLE ($|\lambda_i| < 1 \forall i$
 $\lambda_i = \text{EIGENVALUE}$)

$x(t)$, FOR $t \rightarrow +\infty$, TENDS TO
BECOME A STATIONARY PROCESS
IN WEAK SENSE (I.E. STATIONARY
IN MEAN AND COVARIANCE).

ONE HAS

$$\lim_{t \rightarrow +\infty} \mu(t) = 0, \quad \text{VAR } x(t) = \bar{\Sigma}$$

$$\lim_{t \rightarrow +\infty} \mu(t) = 0, \quad \text{VAR } x(t) = \bar{\Sigma}$$

$$\Sigma_x(t-s) \approx A^{t-s} \bar{\Sigma}, \quad t \geq s$$

WHERE

$$\bar{\Sigma} = \lim_{t \rightarrow +\infty} \Sigma_x(t) \quad \text{IS THE SOLUTION OF}$$

$$\bar{\Sigma} = A \bar{\Sigma} A^T + B B^T \quad \text{LYAPUNOV}$$

IF ONE ALSO HAS

$$\mu_0 = 0, \quad \Sigma_0 = \bar{\Sigma}$$

THEN

$$\mu(t) = 0, \quad \text{VAR } x(t) = \bar{\Sigma} \quad \forall t$$

PROOF:

$$\mu(t_0) = \mu_0, \quad \mu(t) = A^{t-t_0} \mu_0 \xrightarrow{t \rightarrow +\infty} 0$$

(SINCE A IS STABLE)

$$\Sigma(t) = \Sigma$$

PROOF:

$$\mu(t_0) = \mu_0, \quad \mu(t) = A^{t-t_0} \mu_0 \xrightarrow{t \rightarrow +\infty} 0$$

(SINCE A IS STABLE)

$$\Sigma(t_0) = \Sigma_0$$

$$\Sigma(t+1) = A \Sigma(t) A^T + B B^T \quad (*)$$

$$\Sigma(t) = A^{t-t_0} \Sigma_0 (A^{t-t_0})^T \left(\rightarrow 0 \text{ SINCE } A \text{ IS STABLE} \right)$$

$$+ \sum_{k=0}^{t-t_0-1} A^k B B^T (A^T)^k$$

SO

$$\lim_{t \rightarrow +\infty} \Sigma(t) = \sum_{k=0}^{+\infty} A^k B B^T (A^T)^k \quad \left(\text{CONVERGES SINCE } A \text{ IS STABLE} \right)$$

$$\| \dots \| \sum < \infty$$

NOW, IF WE TAKE THE LIMIT ON THE LEFT AND RIGHT OF (*) ONE OBTAINS THAT $\bar{\Sigma}$ SATISFIES

ONE OBTAINS THAT $\bar{\Sigma}$ SATISFIES

$$\bar{\Sigma} = A \bar{\Sigma} A^T + B B^T$$

FINALLY,

$$\mu_0 = 0 \Rightarrow \mu(t) = A^{t-t_0} \mu_0 = 0 \quad \forall t$$

$$\begin{aligned} \Sigma_0 = \bar{\Sigma} \Rightarrow \Sigma(t_0+1) &= A \bar{\Sigma} A^T + B B^T \\ &= \bar{\Sigma} \end{aligned}$$

$$\begin{aligned} \Sigma(t_0+2) &= A \Sigma(t_0+1) A^T \\ &\quad + B B^T \\ &= A \bar{\Sigma} A^T + B B^T \\ &= \bar{\Sigma} \end{aligned}$$

\Downarrow

$$\Sigma(t) = \bar{\Sigma} \quad \forall t$$

NOTE: IF A IS STABLE, THERE IS A UNIQUE SOLUTION OF THE LYAPUNOV EQUATION.

AND THE $\bar{\Sigma}$ OBTAINED IN THE

$$\Sigma_0 = \bar{\Sigma} \Rightarrow \Sigma(t_0+1) = A \bar{\Sigma} A^T + B B^T \\ = \bar{\Sigma}$$

$$\Sigma(t_0+2) = A \Sigma(t_0+1) A^T \\ + B B^T \\ = A \bar{\Sigma} A^T + B B^T \\ = \bar{\Sigma}$$

$$\Downarrow \\ \Sigma(t) = \bar{\Sigma} \quad \forall t$$

NOTE: IF A IS STABLE, THERE IS A UNIQUE SOLUTION OF THE LYAPUNOV EQUATION.

SO, THE $\bar{\Sigma}$ OBTAINED IN THE PROOF WITH A LIMIT IS THE UNIQUE $\bar{\Sigma}$ SOLVING

$$\bar{\Sigma} = A \bar{\Sigma} A^T + B B^T$$



UNIQUENESS OF $\bar{\Sigma}$

LET (BY CONTRADICTION) $\bar{\Sigma} \neq \bar{\Sigma}$ ANOTHER SOLUTION OF THE LYAPUNOV EQUATION.

WE PROVED THAT, FOR ANY Σ_0 ,

$$\lim_{t \rightarrow +\infty} \Sigma(t) = \bar{\Sigma}$$

LET $\Sigma_0 = \bar{\Sigma}$. BUT THEN

$$\Sigma(t) = \bar{\Sigma} \quad \forall t, \text{ so}$$

$\bar{\Sigma}$ MUST BE EQUAL TO $\bar{\Sigma}$

AND THE CONTRADICTION IS OBTAINED

KALMAN FILTER EQUATIONS

$$x_0 \sim (\mu_0, P_0)$$

$$x(t+1) = Ax(t) + v(t)$$

$$y(t) = Cx(t) + w(t)$$

HANDLING NOISE CORRELATION

$$x_0 \perp v(t), w(t) \quad \text{AND} \quad v(t) \perp w(s) \\ t \neq s$$

BUT

$$\text{VAR} \begin{bmatrix} v(t) \\ w(t) \end{bmatrix} = \begin{bmatrix} Q & S \\ S^T & R \end{bmatrix}, \quad R > 0 \text{ BY ASSUMPTION}$$

IT IS NOW CONVENIENT TO WORK WITH UNCORRELATED NOISES

$$v(t), w(t) \rightarrow \tilde{v}(t), w(t) \\ \text{UNCORRELATED}$$

WITH

$$\tilde{v}(t) = v(t) - E[v(t) | H(w)]$$

SUBSPACE
GENERATED
BY $\{w(t)\}$

$$\tilde{v}(t) = v(t) - E[v(t) | H(w)]$$

SUBSPACE
GENERATED
BY $\{w(t)\}$

$$= v(t) - E[v(t) | w(t)]$$

$$= v(t) - \text{COV}(v(t), w(t)) (\text{VAR } w(t))^{-1} w(t)$$

$$= v(t) - SR^{-1} w(t)$$

$$= v(t) - SR^{-1} y(t) + SR^{-1} C x(t)$$



$$x(t+1) = A x(t) + SR^{-1} y(t) - SR^{-1} C x(t) + \tilde{v}(t)$$

$$= \underbrace{(A - SR^{-1} C)}_F x(t) + \underbrace{SR^{-1} y(t)}_{\text{OUTPUT INJECTION}} + \tilde{v}(t)$$

WITH

$$\text{VAR} \begin{bmatrix} \tilde{v}(t) \\ w(t) \end{bmatrix} = \begin{bmatrix} \tilde{Q} & 0 \\ 0 & R \end{bmatrix}$$

$$\tilde{Q} = Q - SR^{-1} S^T$$

$$\tilde{Q} = Q - SR^{-1}S^T$$

\downarrow \downarrow
 VAR v VAR w
 \downarrow
 COV(v, w)

KALMAN FILTER

NOW, WE CAN START FROM

$$x(t+1) = Fx(t) + SR^{-1}y(t) + \tilde{v}(t)$$

$$y(t) = Cx(t) + w(t)$$

$$x(t_0) = x_0 \sim (\mu_0, P_0)$$

$$\text{VAR } \tilde{v} = \tilde{Q}$$

$$\text{VAR } w = R$$

PREDICTION AND FILTERING

$$\hat{x}(t+1|t) = \hat{E}[x(t+1) | H_t(y)]$$

$$\hat{x}(t+1|t+1) = \hat{E}[x(t+1) | H_{t+1}(y)]$$

PREDICTION AND FILTERING

ERRORS

$$\tilde{x}(t+1|t) = x(t+1) - \hat{x}(t+1|t)$$

PREDICTION AND FILTERING ERRORS

$$\tilde{x}(t+1|t) = x(t+1) - \hat{x}(t+1|t)$$

$$\tilde{x}(t|t) = x(t) - \hat{x}(t|t)$$

COVARIANCES OF THE ERRORS

$$P(t+1|t) = E[\tilde{x}(t+1|t) (\tilde{x}(t+1|t))^T]$$

$$P(t|t) = E[\tilde{x}(t|t) (\tilde{x}(t|t))^T]$$

INNOVATION

$$e(t) = y(t) - C \hat{x}(t|t-1)$$

ONE HAS (ASSUMING $\mu_0 = 0$,
BUT THEN WE REPLUG
IT IN THE INITIAL
CONDITION AND OBTAIN
THE SAME EQUATIONS)

NOTES ON \perp IN STATE SPACE

$$1) S = S_a \oplus S_e, \quad S_a \perp S_e$$



$$E[x|S] = E[x|S_a] + E[x|S_e]$$

$$2) \tilde{v}(t) \perp x(t) \quad (\tilde{v}(t) \text{ FORMS } x(t+1))$$

$$3) \tilde{v}(t) \perp x_0, \tilde{v}^{t-1}, w^t \quad (\text{GENERATE ALSO } y(t_0), \dots, y(t))$$

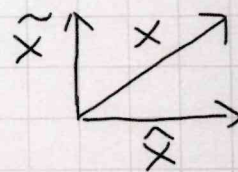


$$\tilde{v}(t) \perp H_t(y) \quad (\text{CONTAINS } \hat{x}(t|t))$$



$$\tilde{v}(t) \perp \hat{x}(t|t)$$

$$4) \tilde{v}(t) \perp \tilde{x}(t|t)$$



SINCE

$$\tilde{x}(t|t) = x(t) - \hat{x}(t|t) \quad \text{AND}$$

$$\tilde{v}(t) \perp x(t) \quad \text{AND} \quad \tilde{v}(t) \perp \hat{x}(t|t)$$

BY ②

BY ③

$$5) e(t+1) = y(t+1) - C \hat{x}(t+1|t) = y(t+1) - \hat{y}(t+1|t)$$

$\Rightarrow e(t+1) = \text{ONE-STEP AHEAD OUTPUT PROJECTION ERROR}$

$$\Rightarrow e(t+1) \perp H_t(y), \quad H_{t+1}(y) = H_t(y) \oplus H(e(t+1))$$

TIME UPDATE

$$\hat{x}(t+1|t) = F \hat{x}(t|t) + SR^{-1} y(t)$$

$$P(t+1|t) = FP(t|t)F^T + \tilde{Q}$$

PROOF

$$x(t+1) = Fx(t) + SR^{-1}y(t) + \tilde{v}(t)$$

AND WE WANT TO PROJECT

ONTO $H_t(y)$. NOTE THAT

$$\tilde{v}(t) \perp \text{SPAN} \{x_0, \tilde{v}(s-1), w(s); s \leq t\} \\ \supset H_t(y)$$

IN FACT

$$x(t+1) = \dots + \tilde{v}(t) \text{ DOES NOT "BUILD" } y(t)$$

$$y(t) = Cx(t) + \dots$$

THEN $\tilde{v}(t) \perp H_t(y)$ AND

$$E[x(t+1) | H_t(y)] = F \hat{x}(t+1|t) + SR^{-1}y(t)$$

SO THAT

SO THAT

$$\begin{aligned}\tilde{x}(t+1|t) &= x(t+1) - \hat{x}(t+1|t) \\ &= Fx(t) + \tilde{v}(t) + SR^{-1}y(t) \\ &\quad - F\hat{x}(t|t) - SR^{-1}y(t) \\ &= F(\underbrace{x(t) - \hat{x}(t|t)}_{\perp \tilde{v}(t)}) + \underbrace{\tilde{v}(t)}_{\perp H_t(y)} \\ &= F\tilde{x}(t|t) + \tilde{v}(t) \quad (\text{SUM OF TWO} \\ &\quad \perp \text{R.V.})\end{aligned}$$

$$\text{VAR } \tilde{x}(t+1|t) = F \text{VAR } \tilde{x}(t|t) F^T + \text{VAR } \tilde{v}(t)$$

$$\Rightarrow P(t+1|t) = F P(t|t) F^T + \tilde{Q}$$

↓
JUST
NOTATION

MEASUREMENT UPDATE

$$\hat{x}(t+1|t+1) = \hat{x}(t+1|t) + L(t+1) \cdot e(t+1)$$

$$e(t+1) = y(t+1) - C \hat{x}(t+1|t)$$

|||
INNOVATION

$$L(t+1) = P(t+1|t) C^T \Lambda^{-1}(t+1)$$

$$\Lambda(t) = \text{VAR } e(t) = C P(t|t-1) C^T + R$$

INNOVATION

$$L(t+1) = P(t+1|t) C^T \Lambda^{-1}(t+1)$$

$$\Lambda(t) = \text{VAR } e(t) = C P(t|t-1) C^T + R$$

$$P(t+1|t+1) = P(t+1|t)$$

$$- P(t+1|t) C^T \Lambda^{-1}(t+1) C P(t+1|t)$$

PROOF

KEY POINT: USE OF INNOVATION
TO PROJECT ONTO

$$H_{t+1}(y) = H_t(y) \oplus H(e(t+1))$$

USING THE FACT THAT PROJECTION
ONTO TWO ORTHOGONAL SUBSPACES
(LIKE $H_t(y)$ AND $H(e(t+1))$)
IS THE SUM OF THE TWO
PROJECTIONS.

NOTE ALSO THAT:

$e(t+1)$ BY DEFINITION IS $\perp H_t(y)$

SINCE $C \hat{x}(t+1|t)$ IS THE OPTIMAL

LINEAR ESTIMATE OF $y(t+1)$ BASED

ON $y(t_0), \dots, y(t)$.

LINEAR ESTIMATE OF $y(t+1)$ BASED
ON $y(t_0), \dots, y(t)$.

$$\begin{aligned} e(t+1) &= y(t+1) - C \hat{x}(t+1|t) \\ &= C [x(t+1) - \hat{x}(t+1|t)] + w(t+1) \\ &= C \tilde{x}(t+1|t) + w(t+1) \end{aligned}$$

$$\begin{aligned} \hat{E} [x(t+1) | H_t(y)] &= \hat{E} [x(t+1) | H_t(y) \oplus H(e(t+1))] \\ &= \hat{E} [x(t+1) | H_t(y)] + \hat{E} [x(t+1) | H(e(t+1))] \end{aligned}$$

⇓

$$\hat{x}(t+1|t+1) = \hat{x}(t+1|t)$$

$$+ \text{COV}(x(t+1), e(t+1)) [\text{VAR}(e(t+1))]^{-1} e(t+1)$$

(a)

(b)

$$\text{(a) } \text{COV}(x(t+1), e(t+1))$$

$$= \text{COV}(x(t+1), y(t+1) - C \hat{x}(t+1|t))$$

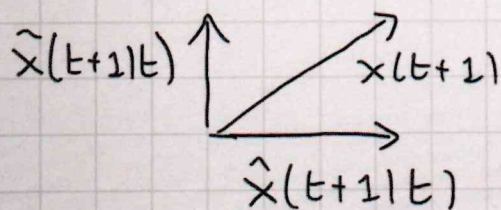
$$= \text{COV}(x(t+1), C(x(t+1) - \hat{x}(t+1|t)) + w(t+1))$$

$$= \text{COV}(x(t+1), C \tilde{x}(t+1|t))$$



$$\begin{aligned}
 \textcircled{a} \quad & \text{COV}(x(t+1), e(t+1)) \\
 &= \text{COV}(x(t+1), y(t+1) - C \hat{x}(t+1|t)) \\
 &= \text{COV}(x(t+1), C(x(t+1) - \hat{x}(t+1|t)) + w(t+1)) \\
 &= \text{COV}(x(t+1), C \tilde{x}(t+1|t)) \\
 &\quad + \underbrace{\text{COV}(x(t+1), w(t+1))}_{=0 \text{ SINCE } x(t+1) \perp w(t+1)}
 \end{aligned}$$

$$= \text{COV}\left(\underbrace{\tilde{x}(t+1|t)}_{\perp H_t(y)} + \underbrace{\hat{x}(t+1|t)}_{\perp \tilde{x}(t+1|t)}, \hat{x}(t+1|t) \right) C^T$$



$$= \text{COV}\left(\tilde{x}(t+1|t), \tilde{x}(t+1|t)\right) C^T$$

$$= P(t+1|t) C^T$$

$$\textcircled{b} \quad \text{VAR}(e(t+1))$$

$$\textcircled{b} \text{ VAR}(e(t+1))$$

$$= \text{VAR}(C \tilde{x}(t+1|t) + w(t+1))$$

$$\perp \hat{x}(t+1|t), \perp x(t+1)$$

$$\Rightarrow \perp \tilde{x}(t+1|t)$$

$$= \text{VAR}(C \tilde{x}(t+1|t)) + \text{VAR} w(t+1)$$

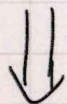
$$= C \text{VAR} \tilde{x}(t+1|t) C^T + R$$

$$= C P(t+1|t) C^T + R$$

FINALLY, TO UPDATE THE
COVARIANCES:

$$x(t+1) - \hat{x}(t+1|t+1)$$

$$= x(t+1) - \hat{x}(t+1|t) - L(t+1)e(t+1)$$



$$\tilde{x}(t+1|t+1) = \tilde{x}(t+1|t) - L(t+1)e(t+1)$$



$$\underbrace{\tilde{x}(t+1|t+1)} + \underbrace{L(t+1)e(t+1)} = \tilde{x}(t+1|t)$$



$$\underbrace{\hat{x}(t+1|t+1)}_{\perp H_{t+1}(y)} + \underbrace{L(t+1)e(t+1)}_{\in H_{t+1}(y)} = \hat{x}(t+1|t)$$

THEN WE TAKE THE VARIANCES
TO OBTAIN

$$P(t+1|t+1) + L(t+1) \underbrace{\Lambda(t+1)}_{\text{VAR}e(t+1)} L(t+1)^T$$
$$= P(t+1|t)$$

INITIAL CONDITION
(NOTATION)

$$\hat{x}(t_0|t_0-1) = \mu_0, \quad P(t_0|t_0-1) = P_0$$

PROJECTION OF
 x_0 ONTO A SPACE
WITH NULL INFORMATION

(NOTATION)

$$\hat{x}(t_0 | t_0 - 1) = \mu_0, \quad P(t_0 | t_0 - 1) = P_0$$

PROJECTION OF
 x_0 ONTO A SPACE
WITH NULL INFORMATION

OUTPUT PREDICTION

$$y(t+1) = Cx(t+1) + w(t+1)$$

PROJECTION
ONTO $H_t(y)$

$$\hat{y}(t+1|t) = C \hat{x}(t+1|t) + 0$$

SINCE
 $w(t+1) \perp H_t(y)$

IT IS
RECURSIVELY
GIVEN BY THE
KF THAT THUS ALSO
GIVES THE OUTPUT PRED.

MULTI-STEP AHEAD

PREDICTOR

$$\hat{x}(t+s+1|t) = F^s \hat{x}(t+1|t)$$

PREDICTOR AND FILTER DYNAMICS

USEFUL TO OBTAIN MORE COMPACT
EQUATIONS

PREDICTOR

$$\hat{x}(t+1|t) = F \hat{x}(t|t-1) + SR^{-1} y(t) \\ + K(t) (y(t) - C \hat{x}(t|t-1))$$

$$K(t) = FL(t) = F P(t|t-1) C^T \Lambda^{-1}(t)$$

$$\Lambda(t) = \text{VAR } e(t)$$

RICCATI EQUATION

$$P(t+1|t) = F \left[P(t|t-1) - P(t|t-1) C^T \Lambda^{-1}(t) \right. \\ \left. \cdot C P(t|t-1) \right] F^T + Q$$

$$P(t_0 | t_0 - 1) = P_0$$

PREDICTOR AS SYSTEM

IN FEEDBACK

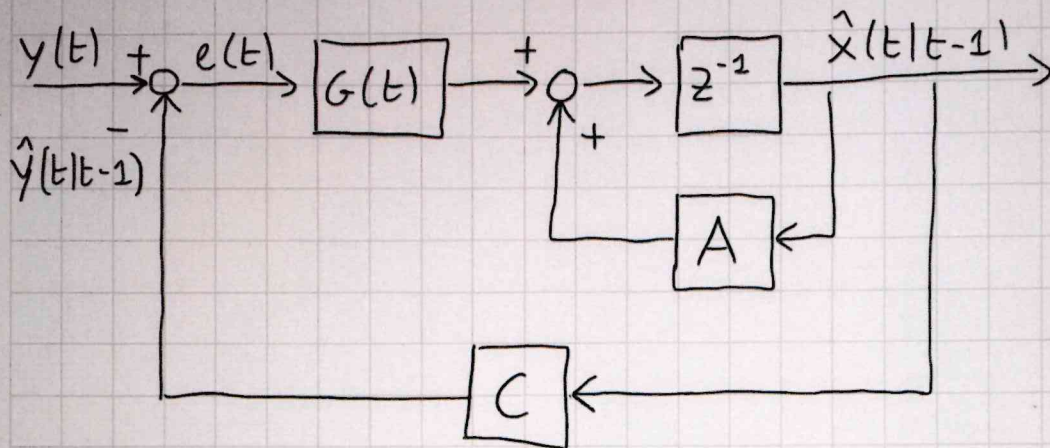
$$\hat{x}(t+1|t) = A \hat{x}(t|t-1) + G(t) \cdot e(t)$$

$$G(t) = K(t) + SR^{-1}$$

$$e(t) = y(t) - C \hat{x}(t|t-1)$$

$$G(t) = K(t) + SR^{-1}$$

$$e(t) = y(t) - C \hat{x}(t|t-1)$$



THE STATE IS THE
PREDICTION
THE SYSTEM IS FED
BY THE INNOVATION

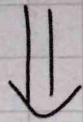
THE STATE TRANSITION MATRIX

WE CAN ALSO WRITE

$$\hat{x}(t+1|t) = \underbrace{(F - K(t)C)}_{P(t)} \hat{x}(t|t-1) + (SR^{-1} + K(t))y(t)$$

$$P(t+1|t) = P(t)P(t|t-1)P^T(t) + K(t)R K^T(t) + \tilde{Q}$$

$$P(t_0 | t_0 - 1) = P_0$$

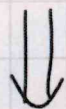


$\Gamma(t)$ ALSO REGULATES THE
ERROR DYNAMICS

$$\tilde{x}(t+1|t) = \Gamma(t) \tilde{x}(t|t-1) - K(t)w(t) + \tilde{v}(t)$$

IF $\Gamma(t)$ IS STABLE

- SMALL SENSITIVITY TO
PARAMETER VARIATION
- SMALL SENSITIVITY TO
MODELING ERRORS
- DISTURBANCE REJECTION



FUNDAMENTAL ISSUES

$$\lim_{t \rightarrow +\infty} \Gamma(t)$$

FUNDAMENTAL ISSUES

$$\lim_{t \rightarrow +\infty} \Gamma(t)$$

(IF IT EXISTS)

WHICH IN TURN REQUIRES
STUDY OF $K(t)$ AND $P(t|t-1)$
FOR $t \rightarrow +\infty$

FILTER EQUATIONS (COMPACT)

LET $S=0$ (NO NOISE CORRELATION)

$$\hat{x}(t+1|t+1) = A \hat{x}(t|t)$$

$$+ P(t+1|t) C^T \Lambda^{-1}(t+1) (y(t+1) - CA \hat{x}(t|t))$$

$$\Lambda(t+1) = C P(t+1|t) C^T + R$$

FROM THE
RICCATI

$$P(t+1|t+1) = [I - P(t+1|t) C^T \Lambda^{-1}(t+1) C] P(t+1|t)$$

NOTE:

NOTE:

$$\text{LET } L(t+1) = P(t+1|t)C^T \Lambda^{-1}(t+1)$$

THEN THE TRANSITION
MATRIX FOR $\hat{x}(t|t)$ IS

$$[I - L(t+1)C]A$$

WHILE THAT FOR $\hat{x}(t|t-1)$ IS

$$A[I - L(t)C] \quad (\text{WITH } S=0)$$

SO THAT THEY SHARE THE
SAME EIGENVALUES

EXAMPLES

①

$$x(t+1) = Ax(t) + v(t)$$

$$y(t) = w(t)$$

$$v(t) \perp w(t)$$

$$\hat{x}(t+1|t) = ?$$

SOL.

$$\hat{x}(t+1|t) = F \hat{x}(t|t-1) + SR^{-1}y(t)$$

SOL.

$$\hat{x}(t+1|t) = F \hat{x}(t|t-1) + SR^{-1}y(t) \\ + FP(t|t-1)C^T \Lambda^{-1}(t)(y(t) - C \hat{x}(t|t-1))$$

$$F = A, \quad S = 0, \quad C = 0$$



$$\hat{x}(t+1|t) = A \hat{x}(t|t-1)$$

AND THIS IS OBVIOUS SINCE
 $y(t)$ DOES NOT CARRY ANY
INFORMATION ON $x(t)$

②

$$x(t+1) = v(t)$$

$$y(t) = Cx(t) + w(t)$$

$$v(t) \perp w(t)$$

$$\text{NOW } A = 0, \quad S = 0, \quad F = A = 0$$



$$\hat{x}(t+1|t) = 0$$



$$\hat{x}(t+1|t) = 0$$

AND THIS IS DUE TO THE FACT THAT $y(t)$ CARRIES INFORMATION ON $x(t)$ BUT NOT ON $x(t+1)$ SINCE $\{x(t)\}$ IS WHITE NOISE (NO DYNAMICS)

③

$$x(t+1) = x(t)$$

$$y(t) = x(t) + w(t), \quad w(t) \sim N(0, 1)$$

$$t_0 = 1$$

$$x(1) \sim N(x_1, 1), \quad x_1 = \hat{x}(1|0)$$

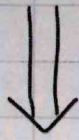
$$P_0 = P(1|0) = 1$$

RICCATI:

$$P(t+1|t) = F \left[P(t|t-1) - P(t|t-1) C^T \Lambda^{-1}(t) \cdot C P(t|t-1) \right] F^T + \tilde{Q},$$

$$F = 1, \quad \tilde{Q} = 0, \quad C = 1$$

$$F=1, \tilde{G}=0, c=1$$



$$P(t+1|t) = P(t|t-1) - \frac{P^2(t|t-1)}{1+P(t|t-1)}$$

$$= \frac{P(t|t-1)}{1+P(t|t-1)}$$



$$P(1|0) = 1,$$

$$P(2|1) = \frac{1}{2},$$

$$P(3|2) = \frac{\frac{1}{2}}{1+\frac{1}{2}} = \frac{1}{3}$$

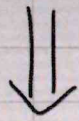
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$$P(t+1|t) = \frac{1}{t+1}$$

||

$$P(t+1|t) = \frac{1}{t+1}$$



$$\hat{x}(t+1|t) = \hat{x}(t|t-1) + \frac{P(t|t-1)[y(t) - \hat{x}(t|t-1)]}{P(t|t-1) + 1}$$

$$= \hat{x}(t|t-1) + \frac{y(t) - \hat{x}(t|t-1)}{t+1}$$

$$= x_1 + \frac{\sum_{i=1}^t y_i}{t+1}$$

(OBVIOUSLY
THE MEAN IN
VIEW OF THE
SYSTEM EQUATIONS!)

⇓
IN FACT:

$$\hat{x}(2|1) = x_1 + \frac{y(1) - x_1}{2} = \frac{x_1 + y(1)}{2}$$

$$\hat{x}(3|2) = \hat{x}(2|1) + \frac{y(2) - \hat{x}(2|1)}{3}$$

$$= \frac{x_1 + y(1)}{2} + \frac{y(2) - \frac{x_1 + y(1)}{2}}{3}$$

$$\hat{x}(3|2) = \hat{x}(2|1) + \frac{y(2) - \hat{x}(2|1)}{3}$$

$$= \frac{x_1 + y(1)}{2} + \frac{y(2) - \frac{x_1 + y(1)}{2}}{3}$$

$$= \frac{x_1 + y(1) + y(2)}{3}$$

SEMIDEFINITE POSITIVE MATRIX

$$A \in \mathbb{R}^{n \times n}$$

DEFINITION: $A = A^T \geq 0$

IF

$$x^T A x \geq 0 \quad \forall x \in \mathbb{R}^n,$$

$$A = A^T > 0$$

IF

$$x^T A x > 0 \quad \forall x \in \mathbb{R}^n$$

CONSEQUENCES

$$x^T A x > 0 \quad \forall x \in \mathbb{R}^n$$

CONSEQUENCES

- IF $A \geq B$, I.E. $A - B \geq 0$, THEN

$$[A]_{ii} \geq [B]_{ii} \quad \forall i$$

(BUT NOT $[A]_{ij} \geq [B]_{ij} \quad \forall i, j$)

$$- B \in \mathbb{R}^{m \times n} \Rightarrow B B^T \in \mathbb{R}^{m \times m}$$

$$\text{AND } B B^T \geq 0$$

$$- A \geq 0, B \in \mathbb{R}^{m \times n} \Rightarrow B A B^T \in \mathbb{R}^{m \times m}$$

$$\text{AND } B A B^T \geq 0$$

$$- A > 0 \Rightarrow \exists \rho_1, \rho_2 > 0 \text{ s.t.}$$

$$\rho_1 I \leq A \leq \rho_2 I$$

TO PROVE THIS RECALL THAT

$$A = A^T > 0 \Leftrightarrow \lambda_i(A) > 0 \quad \forall i$$

EIGENVALUES

$$(A = A^T \geq 0 \Leftrightarrow \lambda_i(A) \geq 0 \quad \forall i)$$

$$- B \in \mathbb{R}^{m \times n} \Rightarrow BB^T \in \mathbb{R}^{m \times m}$$

$$\text{AND } BB^T \geq 0$$

$$- A \geq 0, B \in \mathbb{R}^{m \times n} \Rightarrow BAB^T \in \mathbb{R}^{m \times m}$$

$$\text{AND } BAB^T \geq 0$$

$$- A > 0 \Rightarrow \exists \rho_1, \rho_2 > 0 \text{ s.t.}$$

$$\rho_1 I \leq A \leq \rho_2 I$$

TO PROVE THIS RECALL THAT

$$A = A^T > 0 \Leftrightarrow \lambda_i(A) > 0 \quad \forall i$$

EIGENVALUES

$$(A = A^T \geq 0 \Leftrightarrow \lambda_i(A) \geq 0 \quad \forall i)$$

$$\text{AND USE } \rho_1 = \min(\lambda_i)$$

$$\rho_2 = \max(\lambda_i)$$

OPTIMAL ONE-STEP
AHEAD PREDICTOR:
SIMPLIFIED NOTATION

$$\Sigma: \begin{cases} x(t+1) = Ax(t) + v(t) \\ y(t) = Cx(t) + w(t) \end{cases}$$

$$v(t) \perp w(s) \quad \forall s, t$$

$$\text{VAR} \begin{bmatrix} v(t) \\ w(t) \end{bmatrix} = \begin{bmatrix} Q & 0 \\ 0 & R \end{bmatrix}, \quad R > 0$$

$$t_0 = 1$$

$$x(1) \sim (x_1, P(1))$$

$$x(1) \perp \{v(t), w(t)\}$$

$$E x(1) = x_1, \quad \text{VAR } x(1) = P(1)$$

$$\hat{x}(1|0) = x_1$$

$$P(1|0) =: P(1) = P_0$$

$$P(t|t-1) =: P(t)$$



$$P(t|t-1) =: P(t)$$

|||

|||

COVARIANCE MATRIX OF THE
ONE-STEP AHEAD PREDICTOR ERROR

KALMAN PREDICTOR

$$\begin{cases} \hat{x}(t+1|t) = A\hat{x}(t|t-1) + K(t)e(t) \\ \hat{y}(t+1|t) = C\hat{x}(t+1|t) \end{cases}$$

$$\uparrow = (A - K(t)C)\hat{x}(t|t-1) + K(t)y(t)$$

$K(t)$ = KALMAN GAIN

$$= AP(t)C^T (C P(t)C^T + R)^{-1}$$

WITH

$$\text{DRE: } P(t+1) = AP(t)A^T + Q$$

$$- AP(t)C^T (C P(t)C^T + R)^{-1} C P(t)A^T$$

$e(t)$ = INNOVATION

$$= y(t) - C\hat{x}(t|t-1)$$

ASYMPTOTIC ISSUES

Σ IS TIME INVARIANT SUBJECT TO STATIONARY NOISES BUT THE PREDICTOR $\hat{\Sigma}$ IS TIME VARYING SINCE $K(t)$ VARIES IN TIME

MOTIVATIONS FOR STUDYING THE ASYMPTOTIC PREDICTOR

① IF $\lim_{t \rightarrow +\infty} K(t) = \bar{K}$

I COULD ALREADY USE \bar{K} OBTAINING A SUBOPTIMAL PREDICTOR CONVERGING TO THE OPTIMAL ONE

THIS MOTIVATES THE STUDY OF

$P(t)$ FOR $t \rightarrow +\infty$ (ASYMPTOTICS OF DRE)

IF $P(t) \rightarrow \bar{P}$, $\bar{P} = \bar{P}^T \geq 0$ AND

\bar{P} HAS TO SOLVE THE

ALGEBRAIC RICCATI EQUATION
(ARE)

$$P = AP^T + Q - APC^T(CPC^T + R)^{-1}CPA^T$$

QUESTIONS:

- WHEN $\lim_{t \rightarrow +\infty} P(t)$ EXISTS?
- DOES THE LIMIT \bar{P} (IF IT EXISTS) DEPEND ON P_0 ?
- CAN WE ADMIT MORE THAN ONE SOLUTION? WHICH IS THE RIGHT \bar{P} TO WHICH $P(t)$ CONVERGES?

② BEYOND COMPUTATIONAL CONSIDERATIONS, STUDY OF

$$\lim_{t \rightarrow +\infty} P(t)$$

MAKES UNDERSTAND IF THE $\hat{\Sigma}$ CAN PREDICT THE STATE WITH AN ERROR VARIANCE $P(t)$ WHICH REMAINS SMALL (BOUNDED BY U)

③ ASYMPTOTIC BEHAVIOUR STUDY

③ ASYMPTOTIC BEHAVIOUR STUDY PERMITS TO ASSESS IF THE $\hat{\Sigma}$ TENDS TO A STABLE SYSTEM

$$\hat{\Sigma}_{\infty} : \hat{x}(t+1|t) = (A - \bar{K}C) \hat{x}(t|t-1) + \bar{K}y(t)$$

$$A - \bar{K}C = A - A\bar{P}C^T(C\bar{P}C^T + R)^{-1}C$$

THE PREDICTOR $\hat{\Sigma}_{\infty}$ IS STABLE

$$\text{IF } \max_i |\lambda_i(A - \bar{K}C)| < 1$$

WE WILL SEE THAT

$$\lim_{t \rightarrow +\infty} P(t) = \bar{P} \not\Rightarrow \hat{\Sigma}_{\infty} \text{ STABLE}$$

DEFINITION: A SOLUTION \bar{P} OF THE ARE IS STABILIZING IF

$$\max_i |\lambda_i(A - \bar{K}C)| < 1$$



DRE CONVERGENCE

AND

PREDICTOR STABILITY

WE WILL CONSIDER 3 DIFFERENT SCENARIOS GIVING 3 DIFFERENT THEOREMS. WE WILL PROVE ONLY THE LAST ONE SINCE IT WILL CONTAIN THE FIRST TWO AS SPECIAL CASES. THIS WILL GIVE CRUCIAL INSIGHTS ON THE DRE (NON TRIVIAL MATRIX EQUATION SINCE IT IS HIGHLY NON LINEAR)

SCENARIO # 1:

STABLE \leq

$$x(t+1) = Ax(t) + v(t), \quad |\lambda_i| < 1 \quad \forall i$$

SCENARIO # 1:

STABLE Σ

$$x(t+1) = Ax(t) + v(t), \quad |\lambda_i| < 1 \quad \forall i$$

WE KNOW THAT $x(t)$ TENDS TO BECOME A STATIONARY PROCESS WITH COVARIANCE $\bar{\Sigma}$ THAT IS THE UNIQUE SOLUTION OF

$$\bar{\Sigma} = A\bar{\Sigma}A^T + Q$$

IS IT POSSIBLE THAT $P(t)$ DIVERGES?

NO: IN THE WORST CASE NO DATA $y(t)$ ARE AVAILABLE, THUS $K(t) = 0$

AND

$$\hat{x}(t+1|t) = A\hat{x}(t|t-1)$$

$$\hat{x}(t+1|t) \rightarrow 0 = \lim_{t \rightarrow +\infty} E x(t)$$

SO THAT

$$P(t) \rightarrow \bar{\Sigma} \quad \text{SINCE}$$

AND

$$\hat{x}(t+1|t) = A \hat{x}(t|t-1)$$

$$\hat{x}(t+1|t) \rightarrow 0 = \lim_{t \rightarrow +\infty} E x(t)$$

SO THAT

$$P(t) \rightarrow \bar{\Sigma} \quad \text{SINCE}$$

$$\text{VAR} (x(t+1) - \hat{x}(t+1|t)) \underset{\substack{\approx \\ \text{LARGE} \\ t}}{\approx} \text{VAR } x(t+1) \\ \approx \bar{\Sigma}$$

CAN $\hat{\Sigma}$ BE UNSTABLE?

$\hat{\Sigma}$ STABLE \Rightarrow x, y STATIONARY

AND $y(t)$ IS THE INPUT
TO THE PREDICTOR $\hat{\Sigma}$

$$\hat{\Sigma}_{\infty} : \hat{x}(t+1|t) = (A - \bar{K}C) \hat{x}(t|t-1) + \bar{K} y(t)$$

IF UNSTABLE, \hat{x} NON STATIONARY,

IT SEEMS ABSURD TO
PREDICT A STATIONARY
PROCESS x WITH A NON
STATIONARY \hat{x}

FIRST CONVERGENCE THEOREM

$$\text{STABLE } \hat{\Sigma} \Rightarrow 1) \forall P_0 = P_0^T \geq 0$$

$$\lim_{t \rightarrow +\infty} P(t) = \bar{P}$$

$$2) \hat{\Sigma}_{\infty} \text{ IS STABLE}$$

$$3) \bar{P} \text{ IS THE ONLY}$$

$$\bar{P} = \bar{P}^T \geq 0 \text{ WHICH}$$

SOLVES THE ARE

NOTE: REMOVING THE ASSUMPTION

$R > 0$, $A - \bar{K}C$ COULD HAVE

EIGENVALUES λ_i ON THE

UNIT CIRCLE ($S_y(e^{j\omega}) \geq 0$ ONLY)

RELATIONSHIP WITH WIENER

IF $\hat{\Sigma}$ IS STABLE, IF $\tau = t - t_0$:

$$\lim_{\tau \rightarrow \infty} \left\| \underbrace{\hat{x}(\tau | \tau - 1)}_{\text{KALMAN}} - \underbrace{\hat{x}_{\infty}(\tau | \tau - 1)}_{\text{WIENER-KALMAGORAN}} \right\| = 0$$

< Note



Fine

 $t \rightarrow +\infty$ 2) $\hat{\Sigma}_{\infty}$ IS STABLE3) \bar{P} IS THE ONLY

$$\bar{P} = \bar{P}^T \geq 0 \text{ WHICH}$$

SOLVES THE ARE

NOTE: REMOVING THE ASSUMPTION

 $R > 0$, $A - \bar{K}C$ COULD HAVEEIGENVALUES λ_i ON THEUNIT CIRCLE ($S_y(e^{j\omega}) \geq 0$ ONLY)

RELATIONSHIP WITH WIENER

IF $\hat{\Sigma}$ IS STABLE, IF $\tau = t - t_0$:

$$\lim_{\tau \rightarrow \infty} \left\| \underbrace{\hat{x}(\tau | \tau - 1)}_{\text{KALMAN PREDICTOR}} - \underbrace{\hat{x}_{\infty}(\tau | \tau - 1)}_{\text{WIENER-KOLMOGOROV PREDICTOR WHICH USES MEASUREMENTS FROM } t_0 = -\infty} \right\| = 0$$

$$\|x\| = \text{TRACE OF VAR } x$$

SCENARIO #2: OBSERVABILITY AND CONTROLLABILITY

STABILITY IS SUFFICIENT BUT NOT
NECESSARY FOR PREDICTOR
CONVERGENCE!

- POSSIBLE CONVERGENCE EVEN WITH
UNSTABLE Σ

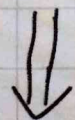
- $P(t)$ COULD SATISFY $P(t) \leq U \forall t$
EVEN IF $\hat{\Sigma}_{\infty}$ IS UNSTABLE

EXAMPLE 1

$$x(t+1) = x(t)$$

$$y(t) = w(t)$$

$$x(1) = x_1, \quad P(1) = 0$$



$$\hat{x}(t+1|t) = \hat{x}(t|t-1) = x_1$$

NECESSARY FOR PREDICTOR CONVERGENCE!

- POSSIBLE CONVERGENCE EVEN WITH
UNSTABLE $\hat{\Sigma}$

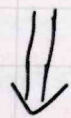
- $P(t)$ COULD SATISFY $P(t) \leq U \forall t$
EVEN IF $\hat{\Sigma}_{\infty}$ IS UNSTABLE

EXAMPLE 1

$$x(t+1) = x(t)$$

$$y(t) = w(t)$$

$$x(1) = x_1, \quad P(1) = 0$$



$$\hat{x}(t+1|t) = \hat{x}(t|t-1) = x_1$$

$$\text{AND } P(t) = 0 \quad \forall t$$

SO, $P(t)$ CONVERGES, $\hat{\Sigma}$ AND $\hat{\Sigma}$
ARE UNSTABLE

EXAMPLE 2

$$\Sigma \begin{cases} x(t+1) = \alpha x(t) + v(t), & v(t) \sim (0, \beta^2) \\ y(t) = \gamma x(t) + w(t), & w(t) \sim (0, 1) \end{cases}$$

$w \perp v$

$$|\alpha| > 1$$

CASE A: $\beta \neq 0, \gamma \neq 0$

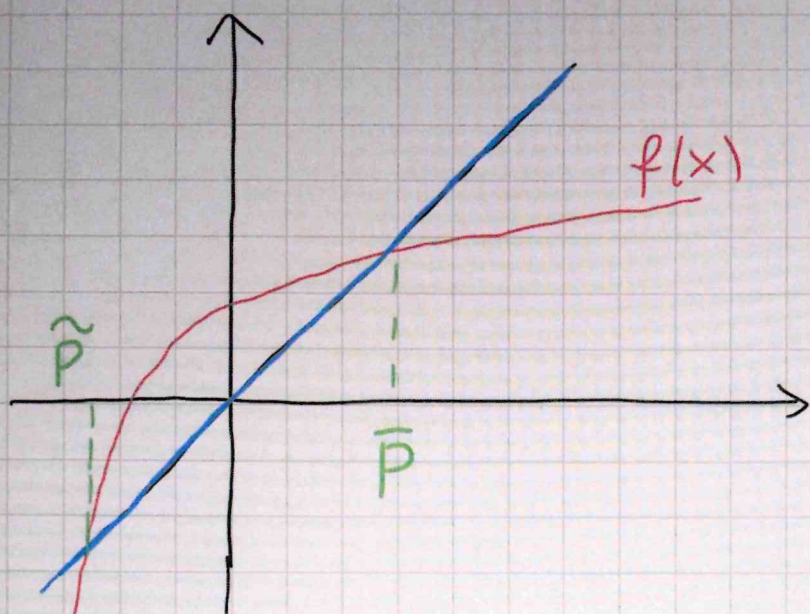
$x(t)$ DIVERGES IN A PROBABILISTIC SENSE ($\text{VAR } x(t) \rightarrow +\infty$), BUT LET US STUDY THE DRE

$$P(t+1) = \frac{\beta^2 + \alpha^2 P(t)}{1 + \gamma^2 P(t)}$$

IT IS USEFUL TO PLOT

$$f(x) = \frac{\beta^2 + \alpha^2 x}{1 + \gamma^2 x}$$

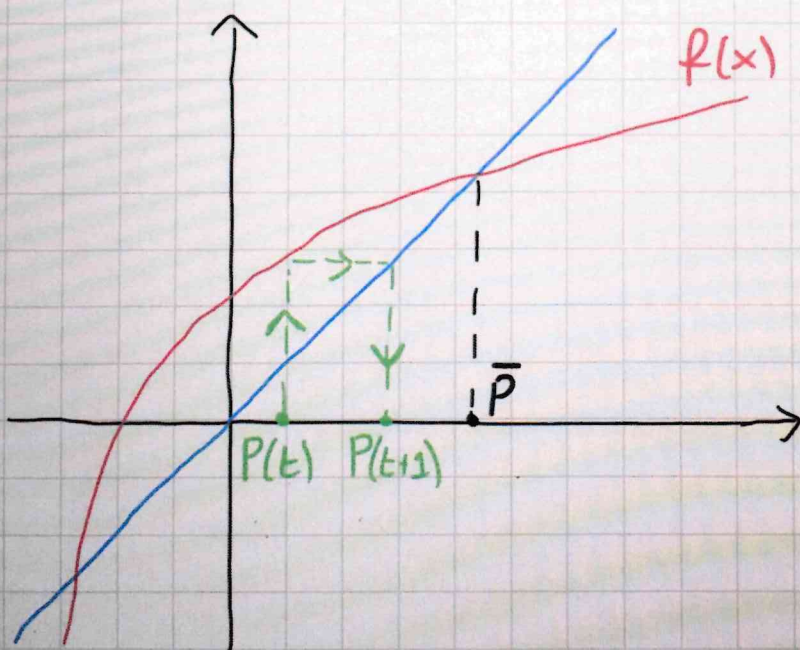
WHOSE \cap WITH $y=x$ GIVES ARE SOLUTION



\bar{P} IS THE ONLY SOLUTION (OF INTEREST SINCE ≥ 0) OF THE ARE

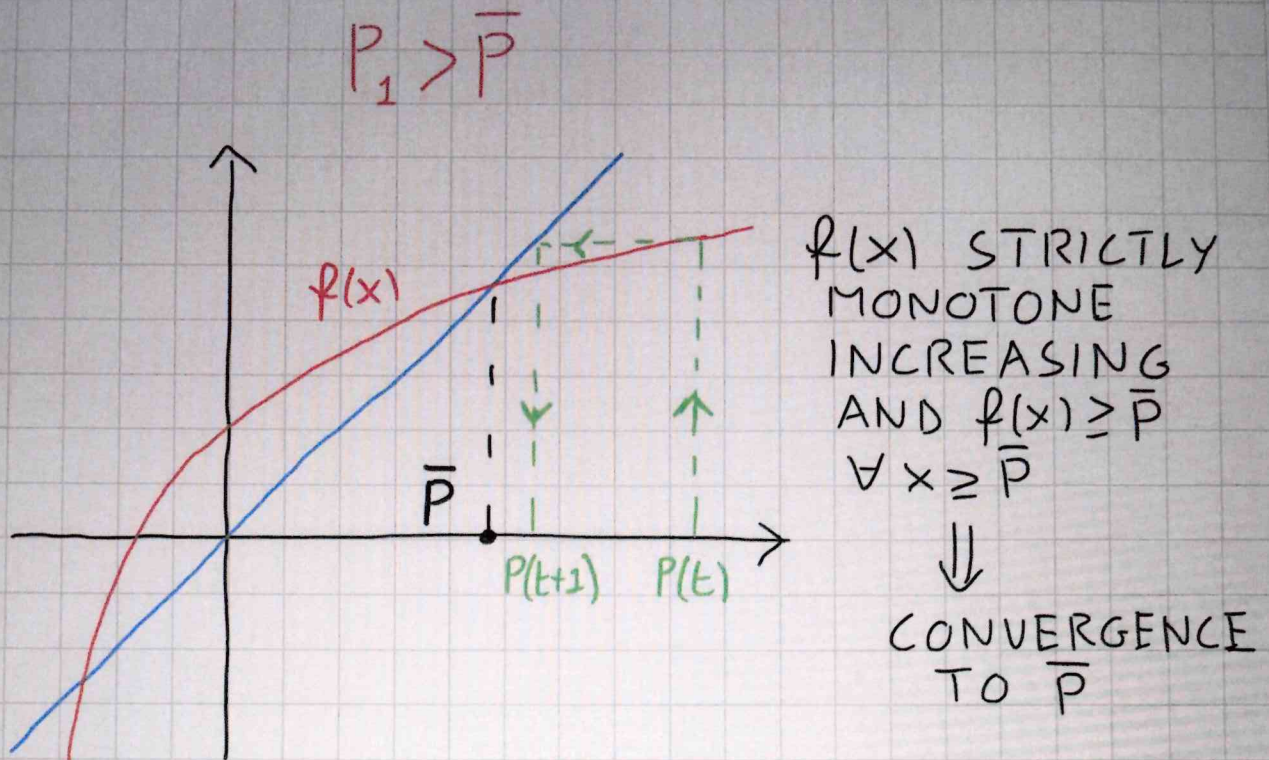
LET US NOW STUDY THE TWO CASES FOR $P(1)$

$$P_1 < \bar{P}$$



$f(x)$ STRICTLY MONOTONE INCREASING AND $f(x) \leq \bar{P} \forall x \leq \bar{P}$
 \Downarrow
 CONVERGENCE TO \bar{P}

$$P_1 > \bar{P}$$



IN CONCLUSION: $\forall P_1 \geq 0$ THERE IS
 CONVERGENCE TO \bar{P}

CHECK ALSO THAT \bar{P} IS STABILIZING
 SINCE $|\alpha - \bar{k}\gamma| < 1$

NOTE THAT THE PREDICTOR IS FED
 WITH AN "UNSTABLE" $y(t)$ AND
 ALSO $\hat{x}(t+1|t)$ DIVERGES BUT THE
 PREDICTION ERROR $P(t+1)$ REMAINS

WELL BOUNDED. IT IS THE NON STATIONARITY OF $y(t)$ THAT ALLOWS THE STABLE PREDICTOR TO FOLLOW THE NONSTATIONARY PROCESS $x(t)$

CASE B: $\beta \neq 0, \gamma = 0$

DRE BECOMES

$$P(t+1) = \beta^2 + \alpha^2 P(t)$$

\Downarrow

$P(t)$ DIVERGES

$K(t)$ IS PROPORTIONAL TO γ

\Downarrow

$$K(t) = 0 \quad \forall t$$

SO, \hat{x} IS IN OPEN LOOP:

$$\hat{x}(t+1|t) = \alpha \hat{x}(t|t-1)$$

AND IS UNSTABLE

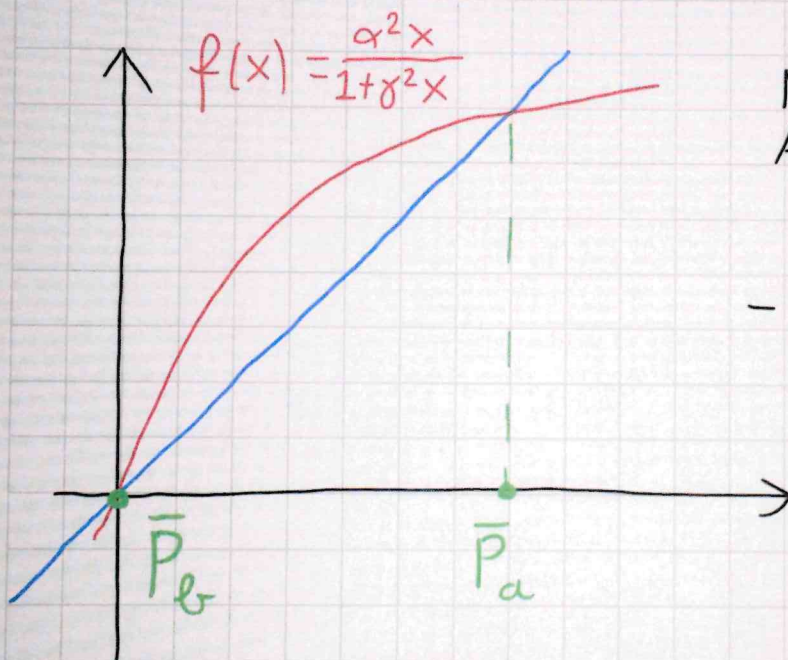
CASE C: $\beta=0, \gamma \neq 0$

SO, NOW

$$x(t+1) = \alpha x(t)$$

$$y(t) = \gamma x(t) + w(t)$$

$$\Rightarrow P(t+1) = \frac{\alpha^2 P(t)}{1 + \gamma^2 P(t)}$$



NOW $f(0) = 0$ AND
ARE HAS TWO
SOLUTIONS:

$$- P_1 = 0 \Rightarrow P(t) = 0 \quad \forall t$$

$$\text{AND } \hat{x}(t+1|t) = \alpha \hat{x}(t|t-1)$$

\bar{P}_e IS NOT
STABILIZING BUT
THE ERROR IS
NULL

$$- P_1 \neq 0 \Rightarrow P(t) \rightarrow \bar{P}_a$$

AS IN THE CASE A

AND \bar{P}_a IS STABILIZING

SO, IN CASE C TWO SOLUTIONS

• FOR ARE, ONE STABILIZING

DISCUSSION ON EXAMPLE #2

FIRST, IT POINTS OUT THE IMPORTANCE OF $y(t)$. Σ IS UNSTABLE, BUT $P(t)$ DIVERGES ONLY IF $\gamma = 0$ (NO INFO FROM THE OUTPUT)

IN GENERAL, CONVERGENCE OF $P(t)$ REQUIRES $y(t)$ TO CARRY INFORMATION (EVEN IF TRIVIAL CASES SUCH AS EXAMPLE #1 EXIST WHERE $\gamma = 0$ AND $P(t) = 0$!)

CASE A WITH $\gamma \neq 0, \beta \neq 0$ ENSURES:

- UNIQUE ARE SOLUTION
- $P(t) \rightarrow \bar{P} \quad \forall P_1 \geq 0$
- STABILIZING \bar{P}

WHICH STRUCTURAL PROPERTIES DOES Σ HAVE IN CASE A?

$\gamma \neq 0 \Rightarrow$ OBSERVABLE Σ

OBSERVABILITY

$$x(t+1) = Ax(t)$$

$$y(1) = Cx(1)$$

$$y(t) = Cx(t)$$

\Rightarrow

$$y(2) = CAx(1)$$

\vdots

$$y(n) = CA^{n-1}x(1)$$

$$n = \text{DIM } x(t)$$

Σ (OR THE COUPLE A, C) IS OBSERVABLE

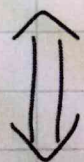
IF THERE ARE NO DISTINCT $x(1)$ THAT

GENERATE THE SAME (FREE) OUTPUTS

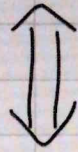
OBSERVABLE Σ



$$\text{RANK} \begin{bmatrix} C^T & A^T C^T & \dots & (A^T)^{n-1} C^T \end{bmatrix} = n$$



$$\text{RANK} \left[C^T \mid A^T C^T \mid \dots \mid (A^T)^{n-1} C^T \right] = n$$



$$\text{RANK} [sI - A^T \mid C^T] = n$$

$\forall s \in \mathbb{C}$ (IT SUFFICES CHECKING $s = \lambda_k$ $k=1, \dots, n$)

THEOREM:

(A, C) OBSERVABLE \implies ARE ADMITS AT LEAST ONE SOLUTION $\bar{P} = \bar{P}^T \geq 0$

$\beta \neq 0 \implies$ CONTROLLABILITY

EVEN IF (A, C) IS OBSERVABLE, ARE CAN HAVE MORE THAN ONE

SOLUTION (RECALL THE CASE $\gamma \neq 0, \beta = 0$)

IF INSTEAD $\gamma \neq 0, \beta \neq 0$ WE OBTAINED

EVEN IF (A, C) IS OBSERVABLE,
 ARE CAN HAVE MORE THAN ONE
 SOLUTION (RECALL THE CASE $\gamma \neq 0, \beta = 0$)

IF INSTEAD $\gamma \neq 0, \beta \neq 0$, WE OBTAINED
 A UNIQUE \bar{P} .

$\beta \neq 0$ ENSURES THAT $x(t)$ IS

INFLUENCED BY THE TRANSITION NOISE.

ARE HAS ONLY ONE SOLUTION IF

$v(t) = Bn(t)$ REACHES ALL THE

COMPONENTS OF $x(t)$. THIS IS

EMBEDDED IN THE CONTROLLABILITY

CONCEPT.

CONTROLLABILITY

$$x(t+1) = Ax(t) + Bn(t)$$

NOW WE HAVE TO CONSIDER THE

CONTROLLABILITY

$$x(t+1) = Ax(t) + Bu(t)$$

NOW WE HAVE TO CONSIDER THE

COUPLE (A, B) OR, EQUIVALENTLY,

(A, Q) OR $(A, Q^{1/2})$ WITH $Q = BB^T$.

$u(t)$ REACHES ALL THE STATE SPACE

IF Σ IS CONTROLLABLE. ONE HAS

CONTROLLABLE Σ



$$\text{RANK} [B \mid AB \mid \dots \mid A^{n-1}B] = n$$



$$\text{RANK} [sI - A \mid B] = n \quad \forall s \in \mathbb{C}$$

$$(s = \lambda_k, k=1, \dots, n)$$

NOTE:

COUPLE (A, B) OR, EQUIVALENTLY,

(A, Q) OR $(A, Q^{1/2})$ WITH $Q = BB^T$.

$x(t)$ REACHES ALL THE STATE SPACE

IF Σ IS CONTROLLABLE. ONE HAS

CONTROLLABLE Σ



$$\text{RANK} [B \mid AB \mid \dots \mid A^{n-1}B] = n$$



$$\text{RANK} [sI - A \mid B] = n \quad \forall s \in \mathbb{C}$$

$$(s = \lambda_k, k=1, \dots, n)$$

NOTE:

OBSERVABLE \iff CONTROLLABLE

(A, C)

(A^T, C^T)

SECOND DRE

CONVERGENCE THEOREM

(A, C)
OBSERVABLE, \implies 1) ARE HAS A
UNIQUE SOLUTION
 $\bar{P} \geq 0$

(A, B)
CONTROLLABLE 2) $P(t) \rightarrow \bar{P} \quad \forall P_0 \geq 0$

3) $\bar{P} > 0$ ANY
STABILIZING

NOTE a: LIKE 1ST TH. EXCEPT $\bar{P} > 0$

NOTE b: 2) \implies 1). IN FACT, IF

$$P(t) \rightarrow \bar{P} \quad \forall P_0 \geq 0,$$

ASSUME $\exists A, B$ SOLUTIONS OF ARE,
WITH $A \geq 0, B \geq 0, A \neq B$.

BUT IF $P_0 = A, P(t) = A \quad \forall t$

IF $P_0 = B, P(t) = B \quad \forall t$

AND THIS CONTRADICTS 2), I.E.



AND THIS CONTRADICTS 2), I.E.
IT WOULD NOT BE TRUE THAT

$$P(L) \rightarrow \text{SAME MATRIX} \quad \forall P_0 \geq 0$$

THE 1ST AND 2ND THEOREM
DO NOT GIVE NECESSARY
AND SUFFICIENT CONDITIONS

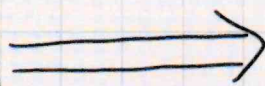
STABLE
 Σ



WE WILL SEE
THAT BOTH OF
THESE CONDITIONS

(A, C) OBS.

(A, B) CONTR.



(TAKEN ALONE)

IMPLY A KEY
STRUCTURE OF Σ

REPRESENTATION OF

REPRESENTATION OF NON CONTROLLABLE Σ

$$\text{DIM} [B \ AB \ \dots \ A^{n-1}B] = p < n$$

$\exists T$ INVERTIBLE DEFINING

$$x = T \bar{x}, \quad \bar{x} = \text{STATE IN THE NEW COORDINATES}$$

S.T. Σ BECOMES (JUST A, B TO SEE):

$$\bar{A} = T^{-1} A T = \begin{matrix} \begin{matrix} \updownarrow p \\ \left[\begin{array}{cc} \bar{A}_{11} & \bar{A}_{12} \\ 0 & \bar{A}_{22} \end{array} \right] \end{matrix} \end{matrix},$$

$$\bar{B} = T^{-1} B = \begin{matrix} \left[\begin{array}{c} \bar{B}_1 \\ 0 \end{array} \right] \end{matrix} \updownarrow p$$

AND THE SYSTEM IS WRITTEN

Note



Fine

AND THE SYSTEM IS WRITTEN
IN CONTROLLABLE FORM.

ONE HAS:

a) $(\bar{A}_{11}, \bar{B}_1)$ IS CONTROLLABLE,

$(\bar{A}_{22}, 0)$ IS THE NON CONTROLLABLE
SUBSYSTEM

b) $\lambda_i(\bar{A}_{22})$ ARE EXACTLY THOSE
 $s \in \mathbb{C}$ S.T.

$$\text{RANK} [A - sI \ B] < n$$

(THE λ_i WHICH MAKE THE
PBH TEST FAIL)

$$c) \bar{x}_1(t+1) = \bar{A}_{11} \bar{x}_1(t) + \bar{A}_{12} \bar{x}_2(t) + \bar{B}_1 u(t)$$

$$\bar{x}_2(t+1) = \bar{A}_{22} \bar{x}_2(t)$$

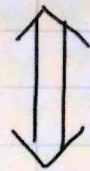
DYNAMICS OF \bar{x}_2 ARE NOT
INFLUENCED BY THE INPUT

STABILITY

STABILIZABILITY

Σ IS SAID TO BE STABILIZABLE

$$|\lambda_i(\bar{A}_{22})| < 1 \quad \forall i$$



$$\exists K \text{ S.T. } |\lambda_i(A+BK)| < 1 \quad \forall i$$

REPRESENTATION OF NON OBSERVABLE Σ

$$\text{RANK} [C^T \ A^T C \ \dots \ (A^T)^{n-1} C^T] = p < n$$

$\exists T$ INVERTIBLE WITH

$$X = T \bar{X} \quad \text{S.T.}$$

$$\bar{A} = T^{-1} A T = \begin{bmatrix} \bar{A}_{11} & 0 \\ \bar{A}_{21} & \bar{A}_{22} \end{bmatrix},$$

$$\bar{B} = T^{-1}B = \begin{bmatrix} \bar{B}_1 \\ \bar{B}_2 \end{bmatrix} \updownarrow \text{?}$$

$$\bar{C} = CT = \begin{bmatrix} \bar{C}_1 & 0 \end{bmatrix} \leftarrow \text{?}$$

AND ONE HAS

a) $(\bar{A}_{11}, \bar{C}_1)$ OBSERVABLE,

$(\bar{A}_{11}, \bar{B}_1, \bar{C}_1)$ OBSERVABLE
SUBSYSTEM

b) $\lambda_i(\bar{A}_{22})$ ARE ALL THE SET
S.T.

$$\text{RANK} [A^T - sI \quad C^T] < n$$

c) $(\bar{A}_{22}, \bar{B}_2, 0)$ IS THE NON OBS.
SUBSYSTEM.

$$\bar{x}_1(t+1) = \bar{A}_{11}x_1(t) + \bar{B}_1u(t)$$

$$\bar{x}_2(t+1) = \bar{A}_{21}x_1(t) + \bar{A}_{22}x_2(t) + \bar{B}_2u(t)$$

$$y(t) = \bar{C}_1x_1(t), \quad \bar{x}_2(t) \text{ DOES NOT}$$

INFLUENCE $y(t)$

DETECTABILITY

Σ IS SAID TO BE DETECTABLE

$$\text{IF } |\lambda_i(\bar{A}_{22})| < 1 \quad \forall i$$



$\exists K$ S.T.

$$|\lambda_i(A + Kc)| < 1 \quad \forall i$$

A SIMPLE OBSERVATION

STABLE $\Sigma \implies$ STABILIZABLE
AND
DETECTABLE Σ

(IN FACT $|\lambda_i(A)| < 1 \quad \forall i$)

ANOTHER SIMPLE OBSERVATION

2

DETECTABLE Σ

(IN FACT $|\lambda_i(A)| < 1 \forall i$)

ANOTHER SIMPLE OBSERVATION

(A, C) OBS. \implies STABILIZABLE
 (A, B) CONTR. AND
DETECTABLE Σ

(IN FACT THERE ARE NO
UNSTABLE λ_i IN THE NON OBS.
AND NON CONTR. SUBSYSTEMS
SINCE SUCH SUBSYSTEMS
DO NOT EXIST!)

IS MAYBE STABILIZABILITY
AND DETECTABILITY THE CONDITION
WEAKER THAN STABILITY AND
THAN (OBS. + CONTR.) WHICH IS
KEY TO CHARACTERIZE THE
ASYMPTOTICS OF $\hat{\Sigma}$?

GENERAL CONVERGENCE THEOREM

CONSIDER

$$x(t+1) = Ax(t) + v(t), \quad \text{VAR } v(t) = Q$$

$$y(t) = Cx(t) + w(t), \quad \text{VAR } w(t) = R > 0$$

$$v \perp w$$

IF $Q = BB^T$, ONE HAS

$$\begin{aligned} (A, C) \text{ DETECTABLE} & \iff 1) P(t) \rightarrow \bar{P} \quad \forall P_0 \\ (A, B) \text{ STABILIZABLE} & \iff 2) \exists ! \bar{P} = \bar{P}^T \geq 0 \\ & \text{ARE SOLUTION} \\ & 3) \bar{P} \text{ IS STABILIZING} \end{aligned}$$

WE WILL PROVE \Rightarrow IN SIX POINTS

POINT 1:

DETECTABILITY $\Rightarrow \forall P_0 \exists U = U^T \geq 0$ S.T.

$$\text{DETECTABILITY} \Rightarrow \forall P_0 \exists U = U^T \geq 0 \text{ s.t.} \\ P(t) \leq U \quad \forall t$$

PROOF: DETECTABILITY IMPLIES

THAT EXISTS \hat{K} S.T. $A - \hat{K}C$ IS STABLE.

WE USE A PREDICTOR WITH GAIN \hat{K} :

$$x(t+1|t) = Ax(t|t-1) + \hat{K}(y(t) - Cx(t|t-1))$$

\Downarrow

$$x(t+1) - x(t+1|t) = (A - \hat{K}C)(x(t) - x(t|t-1)) \\ + v(t) - \hat{K}w(t)$$

SO THE ERROR IS THE STATE OF A STABLE SYSTEM FED WITH A STATIONARY INPUT $v(t) - \hat{K}w(t)$

\Downarrow

$\tilde{P}(t) := \text{VAR}(x(t+1) - x(t+1|t))$ TENDS TO

A FINITE MATRIX. SO THERE

EXISTS U S.T.

$$\tilde{P}(t) < U$$

∩

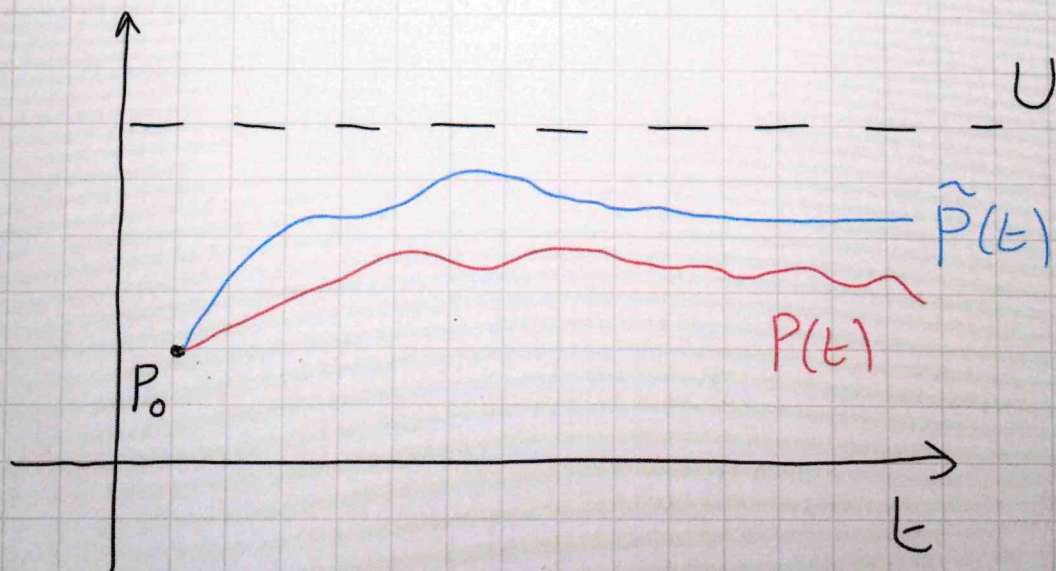
$\tilde{P}(t) := \text{VAR}(x(t+1) - x(t+1|t))$ TENDS TO
A FINITE MATRIX. SO THERE
EXISTS U S.T.

$$\tilde{P}(t) \leq U$$

WHERE $\tilde{P}(t)$ AND U DEPEND ON P_0 .
FOR ANY P_0 , THE DRE GENERATES
VARIANCE ERRORS $P(t)$ S.T.

$$P(t) \leq \tilde{P}(t) \leq U \quad \forall t$$

BY THE OPTIMALITY OF
THE KALMAN FILTER.



NOTE: WE CANNOT SAY THAT $P(t)$ CONVERGES TO SOMETHING AS $\tilde{P}(t)$ DOES. \tilde{K} IS SUBOPTIMAL BUT STABILIZES BY CONSTRUCTION

$$(|\lambda_i(A - \tilde{K}C)| < 1 \forall i).$$

$K(t)$ IS OPTIMAL BUT IT IS NOT KNOWN IF IT CONVERGES TO A \bar{K} (WHICH COULD STABILIZE $A - \bar{K}C$).

POINT 2:

GIVEN TWO INITIAL CONDITIONS

$$P'(1) = M \geq 0, \quad P''(1) = N \geq 0,$$

THEN

$$M \geq N \Rightarrow P'(t) \geq P''(t) \quad \forall t$$

↓ ↓
DRE SOLUTIONS
WITH I.C.

$P'(1), P''(1)$, RESPECTIVELY

PROOF: USEFUL TO WRITE THE R.H.S.

— E THE ARE USING

PROOF: USEFUL TO WRITE THE R.H.S.
OF THE DRE USING

$$\Delta(P, K) = (A - KC)P(A - KC)^T + Q + KRK^T \quad *$$

Δ : 2 MATRICES \rightarrow 1 MATRIX

IF WE USE

$$K = K(P) = AP^T(C^T P^T + R)^{-1},$$

ONE OBTAINS THE DRE

$$\bar{\Pi}(P) = \Delta(P, K(P))$$

$\bar{\Pi}$ = DRE: 1 MATRIX \rightarrow 1 MATRIX

$$P(t+1) = \bar{\Pi}(P(t)) \quad \text{DRE}$$

$$P = \bar{\Pi}(P) \quad \text{ARE}$$

LET US NOW PROVE TWO INEQUALITIES

$$a) M \geq N \Rightarrow \underbrace{\Delta(M, K(M))}_{= \bar{\Pi}(M)} \geq \underbrace{\Delta(N, K(M))}_{\neq \bar{\Pi}(N)}$$

THIS DERIVES FROM $*$. WE CAN

REPLACE THERE $(M, K(M))$ AND THEN

$(N, K(M))$ AND ONE EASILY SEES



$(N, K(M))$ AND ONE EASILY SEES
 THAT $\Delta(M, K(M)) \geq \Delta(N, K(M))$

REDUCES TO

$$(A - KC)M(A - KC)^T \geq (A - KC)N(A - KC)^T$$

$(K = K(M) \text{ EVERYWHERE})$

WHICH HOLDS TRUE IF $M \geq N$

(SINCE $M \geq N \iff x^T M x \geq x^T N x \forall x$)

$$b) \underbrace{\Delta(N, K(M))}_{\neq \Pi(N)} \geq \underbrace{\Delta(N, K(N))}_{= \Pi(N)}$$

THIS INEQUALITY HOLDS SINCE
 EACH Δ DESCRIBES THE ERROR
 COVARIANCES OF TWO FILTERS WITH
 $P(1) = N$ AND ON THE RIGHT THE
 OPTIMAL KALMAN GAIN $K(N)$ IS
 USED.

$$b) \underbrace{\Delta(N, K(M))}_{\neq \pi(N)} \geq \underbrace{\Delta(N, K(N))}_{= \pi(N)}$$

THIS INEQUALITY HOLDS SINCE EACH Δ DESCRIBES THE ERROR COVARIANCES OF TWO FILTERS WITH $P(1) = N$ AND ON THE RIGHT THE OPTIMAL KALMAN GAIN $K(N)$ IS USED.

COMBINING a) AND b) :

$$\Delta(M, K(M)) \underset{a)}{\geq} \Delta(N, K(M)) \underset{b)}{\geq} \Delta(N, K(N))$$

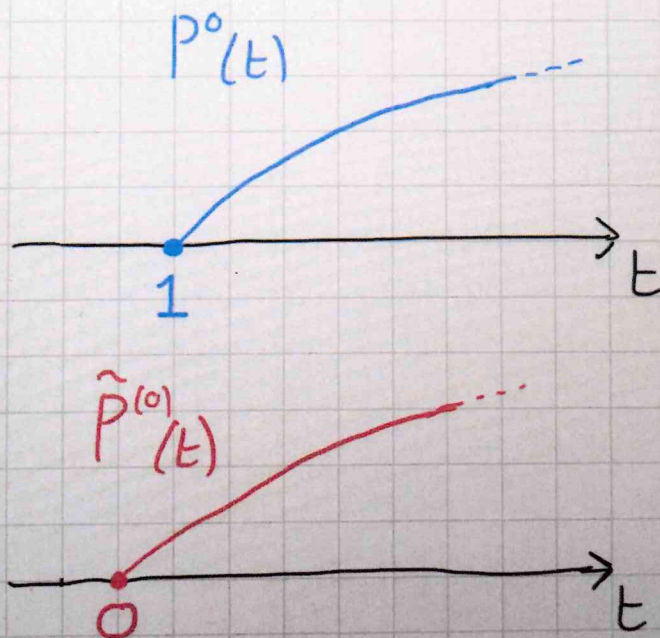
AND THIS CONCLUDES THE PROOF

POINT 3:

THE DRE SOLUTION WITH INITIAL CONDITION $P(1)=0$ IS MONOTONICALLY INCREASING IN MATRIX SENSE, I.E.

$$P^{(0)}(t+1) \geq P^{(0)}(t)$$

PROOF: THE MAIN IDEA IS TO INTRODUCE TWO FILTERS WITH A TIME DELAY AND THEN TO EXPLOIT POINT 2



SO, CONSIDER A FILTER WITH NULL
l.c. AT $t_0=0$ IN PLACE OF $t_0=1$.

$P(0)=0$ AND $\tilde{P}^{(0)}(t)$ IS THE
RELATED DRE SOLUTION. ONE HAS

$$\tilde{P}^{(0)}(t) = P^{(0)}(t+1) \quad (\alpha)$$

IT IS OBVIOUS THAT

$$\tilde{P}^{(0)}(1) \geq P^{(0)}(1) \quad \text{SINCE } P^{(0)}(1)=0$$

BUT THEN, USING POINT 2,

$$\overline{\Pi}(\tilde{P}^{(0)}(t)) \geq \overline{\Pi}(P^{(0)}(t)) \quad \forall t$$

$$\begin{array}{ccc} \text{|||} & & \text{|||} \\ \tilde{P}^{(0)}(t+1) & \geq & P^{(0)}(t+1) \end{array}$$

⇓ USING (α)

$$P^{(0)}(t+2) \geq P^{(0)}(t+1)$$

POINT 4:

THE $P^{(0)}(t)$ SEEN BEFORE
CONVERGES TO A MATRIX \bar{P} , I.E.

$$\lim_{t \rightarrow +\infty} P^{(0)}(t) = \bar{P}$$

PROOF: IT DERIVES FROM THE
FACT THAT $P^{(0)}(t)$ IS MONOTONICALLY
NON DECREASING AND BOUNDED BY U .

IN FACT:

LET $P_i = P_i^T \geq 0$ A SEQUENCE OF
MATRICES S.T.

$$P_1 \leq P_2 \leq \dots \leq U.$$

LET US PROVE THAT

$$\lim_{i \rightarrow +\infty} P_i = \bar{P}$$

BY ASSUMPTION ONE HAS

BY ASSUMPTION ONE HAS

$$x^T P_i x \leq x^T P_{i+1} x \leq x^T U x$$

$\forall x, i$

SO, $\{x^T P_i x\}$ IS A SEQUENCE OF BOUNDED AND MONOTONICALLY NON DECREASING SCALARS. SO, IT CONVERGES.

⑧

LET

$$e_j = [0 \ 0 \ \dots \ 1 \ 0 \ \dots \ 0]^T$$

$\underbrace{\hspace{2cm}}$
j-TH ELEMENT

IT IS EASY TO SEE THAT

$$2 e_a^T P_i e_b = \underbrace{(e_a + e_b)^T P_i (e_a + e_b)}_{\text{CONVERGES BY } \textcircled{8} \text{ WITH } x = e_a + e_b}$$

$\underbrace{[P_i]_{ab}}$

$$= \underbrace{e_a^T P_i e_a} - \underbrace{e_b^T P_i e_b}$$

IT IS EASY TO SEE THAT

$$\underbrace{2 l_a P_i l_b}_{[P_i]_{ab}} = \underbrace{(l_a + l_b)^T P_i (l_a + l_b)}_{\text{CONVERGES BY } \textcircled{\delta} \text{ WITH } x = l_a + l_b}$$

$$- \underbrace{l_a^T P_i l_a}_{\text{CONVERGES BY } \textcircled{\delta} \text{ WITH } x = l_a} - \underbrace{l_b^T P_i l_b}_{\text{CONVERGES BY } \textcircled{\delta} \text{ WITH } x = l_b}$$

SO

$\lim_{i \rightarrow \infty} [P_i]_{ab}$ CONVERGES TOWARDS

A LIMIT IDENTIFYING

$$[\bar{P}]_{ab}$$

$$\Downarrow$$
$$P_i \rightarrow \bar{P}$$

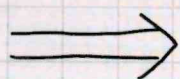
COMMENT AFTER

POINTS 1-4

WE HAVE JUST USED DETECTABILITY.

WE HAVE PROVED THAT

(A, c)
DETECTABLE



1) IF $P(1) = 0$,

DRE SOLUTION $P(t)$

CONVERGES TO A
MATRIX WE DENOTE
WITH $\bar{P} \geq 0$

2) SO, THERE EXISTS

AT LEAST A SOLUTION

TO THE ARE ($\bar{P} \geq 0$)

3) ANY SOLUTION

OF THE DRE (WITH

ANY $P(1)$) IS BOUNDED

NOTE ON EIGENVECTORS

AND EIGENVALUES

NOTE ON EIGENVECTORS AND EIGENVALUES

GIVEN $A \in \mathbb{R}^{n \times n}$, $v \in \mathbb{C}^n$, IF

$$Av = \lambda v, \lambda \in \mathbb{C},$$

THEN v IS A RIGHT EIGENVECTOR
OF A . INSTEAD $w \in \mathbb{C}^n$ S.T.

$$w^T A = \lambda w^T, \lambda \in \mathbb{C}$$

IS A LEFT EIGENVECTOR

(ALSO SATISFYING

$$A^T w = \lambda w)$$

RIGHT AND LEFT EIGENVECTORS
ARE IN GENERAL DIFFERENT BUT
SHARE THE SAME EIGENVALUES.

POINT 5:

POINT 5:

\bar{P} (AS GIVEN BY POINT 4)

IS STABILIZING

PROOF: \bar{P} SATISFIES THE
ARE, I.E.

$$\bar{P} = (A - \bar{K}C) \bar{P} (A - \bar{K}C)^T + \bar{K}R\bar{K}^T + Q \quad (\beta)$$

WHERE \bar{K} IS THE GAIN ASSOCIATED
WITH \bar{P} , I.E.

$$\bar{K} = A \bar{P} C^T (C \bar{P} C^T + R)^{-1}$$

FOR SAKE OF CONTRADICTION,

ASSUME $A - \bar{K}C$ UNSTABLE.

THUS, THERE IS A LEFT EIGENVECTOR
 x S.T.

$$(A - \bar{K}C)^T x = \lambda x, \quad x \neq 0, \quad |\lambda| \geq 1 \quad (\gamma)$$

FROM $\textcircled{\beta}$, IF x^* IS THE CONJUGATE
TRANSPOSE OF x :

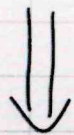
$$x^* \bar{P} x = x^* \underbrace{(A - \bar{K}C)}_{\lambda^* x^*} \bar{P} \underbrace{(A - \bar{K}C)^T}_{\lambda x} x$$

$$|\lambda|^2 x^* \bar{P} x \text{ BY } \textcircled{\delta}$$

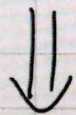
$$+ x^* K R K^T x + x^* Q x$$

SO THAT

$$\underbrace{(1 - |\lambda|^2)}_{\leq 0} \underbrace{x^* \bar{P} x}_{\geq 0} = \underbrace{x^* \bar{K} R \bar{K}^T x}_{\geq 0} + \underbrace{x^* Q x}_{\geq 0}$$

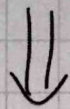


$$" \leq 0 = \geq 0 "$$



$$\begin{cases} x^* K R \bar{K}^T x = 0 \\ x^* Q x = 0 \end{cases}$$

||

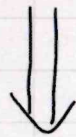


a) $\bar{K}^T x = 0$ (SINCE $R > 0$)

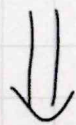
WHICH, COMBINED WITH $\textcircled{\gamma}$,

IMPLIES $A^T x = \lambda x$

b) $Q x = 0$



$$\begin{bmatrix} A^T - \lambda I \\ Q \end{bmatrix} x = \begin{bmatrix} 0 \\ 0 \end{bmatrix}$$



$[A - \lambda I \quad Q]$ DOES NOT

HAVE FULL RANK FOR

λ WITH $|\lambda| \geq 1$ BUT

THIS CONTRADICTS

STABILIZABILITY

POINT 6:

$$\lim_{t \rightarrow +\infty} P(t) = \bar{P} \quad \forall P(1)$$

WITH THE SAME \bar{P} DEFINED ABOVE
SETTING $P(1) = 0$.

PROOF:

WE WILL USE DIFFERENT YET
EQUIVALENT DRE EXPRESSIONS:

$$P(t+1) = (A - K(t)C)P(t)A^T + Q \quad (1)$$

AND

$$P(t+1) = (A - K(t)C)P(t)(A - K(t)C)^T + K(t)R K^T(t) + Q \quad (2)$$

NOW, FIX $P(1) = P_0 > 0$ AND LET

$$- \Psi(t) = (A - K(t-1)C)(A - K(t-2)C) \dots (A - K(1)C)$$

$$- \rho > 0 \text{ s.t. } P_0 \geq \rho I$$

USING THESE DEFINITIONS IN (2)

WE OBTAIN:

PROOF:

WE WILL USE DIFFERENT YET
EQUIVALENT DRE EXPRESSIONS:

$$P(t+1) = (A - K(t)C)P(t)A^T + Q \quad (1)$$

AND

$$P(t+1) = (A - K(t)C)P(t)(A - K(t)C)^T + K(t)R K^T(t) + Q \quad (2)$$

NOW, FIX $P(1) = P_0 > 0$ AND LET

$$- \Psi(t) = (A - K(t-1)C)(A - K(t-2)C) \dots (A - K(1)C)$$

$$- \rho > 0 \text{ s.t. } P_0 \geq \rho I$$

USING THESE DEFINITIONS IN (2)

WE OBTAIN:

$$P(t) = \Psi(t)P_0\Psi^T(t) + [\geq 0]$$

$$\geq \Psi(t)P_0\Psi^T(t) \geq \rho \Psi(t)\Psi^T(t).$$

WE KNOW THAT THERE IS U S.T.

$$P(t) \leq U \quad \forall t$$

SO,

$$U \geq P(t) \geq \rho \Psi(t) \Psi^T(t)$$

AND TAKING THE TRACE

$$\infty > \text{TRACE}[U] \geq \rho \sum_{ij} \Psi_{ij}^2(t)$$



$\Psi(t)$ IS UNIFORMLY
BOUNDED IN t .

NOW, WE USE ①:

$$\begin{aligned} P(t+1) &= (A - K(t)C)P(t)A^T + Q \\ &= P^T(t+1) \\ &= A P(t) (A - K(t)C)^T + Q. \end{aligned}$$

RECALL THAT

$$\bar{P} = (A - \bar{K}C) \bar{P} A^T + Q.$$

NOW WE HAVE:

NOW WE HAVE:

$$\begin{aligned} P(t+1) - \bar{P} &= A P(t) (A - K(t)C)^T - (A - \bar{K}C) \bar{P} A^T \\ &= (A - \bar{K}C) (P(t) - \bar{P}) (A - K(t)C)^T \\ &\quad + \underbrace{\bar{K}C P(t) (A - K(t)C)^T - (A - \bar{K}C) \bar{P} C^T K^T(t)} \end{aligned}$$

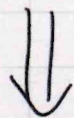
= 0! TEDIOUS CALCULATIONS
USING

$$K(t) = A \left(P(t) - P(t)C^T (C P(t)C^T + R)^{-1} C P(t) \right) C^T R^{-1},$$

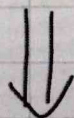
$$\bar{K} = \dots$$



$$P(t+1) - \bar{P} = (A - \bar{K}C) (P(t) - \bar{P}) (A - K(t)C)^T$$



$$P(t) - \bar{P} = \underbrace{(A - \bar{K}C)^{t-1}}_{\substack{\text{STABLE,} \\ \text{SINCE } \bar{P} \\ \text{IS STABILIZING}}} (P_0 - \bar{P}) \underbrace{(A - K(1)C)^T \dots (A - K(t-1)C)^T}_{= \Psi^T(t) \text{ BOUNDED IN } t}$$



$$\lim_{t \rightarrow +\infty} P(t) = \bar{P} \quad \text{IF } P(1) = P_0 > 0$$

BUT IF $P_0 \geq 0$?

LET NOW $P_0 \geq 0$ AND $P_0 \neq 0$

$\exists \lambda > 0$ S.T.

$$0 \leq P(1) = P_0 \leq \lambda I$$

WITH SUCH I.C. WE CONVERGE TO \bar{P} ,
 $\tilde{P}_0(t) \rightarrow \bar{P}$

LEADS TO
 $P(t)$

WE JUST PROVED CONVERGENCE TO \bar{P}
 $\tilde{P}_\lambda(t) \rightarrow \bar{P}$

BY POINT 2, SUCH ORDERING OF I.C. $0 \leq P_0 \leq \lambda I$ PERSISTS IN TIME, I.E.

$$\tilde{P}_0(t) \leq P(t) \leq \tilde{P}_\lambda(t)$$

AND SO ALSO THE $P(t)$ WILL
 CONVERGE TO \bar{P} AS FORMALLY
 PROVED BELOW.

$\forall x$ WE HAVE

$$x^T \tilde{P}_0(t) x \leq x^T P(t) x \leq x^T \tilde{P}_\lambda(t) x$$

$$\begin{array}{ccc} t \rightarrow +\infty \downarrow & & \downarrow & & \downarrow t \rightarrow +\infty \\ & & & & \end{array}$$

$$x^T \bar{P} x \Rightarrow \text{MUST CONVERGE TO } x^T \bar{P} x \Leftarrow x^T \bar{P} x$$



NOW, AS ALREADY SEEN

$$e_i^T P(t) e_j = (i, j)\text{-TH ELEMENT OF } P(t)$$

$$= \frac{(e_i + e_j)^T P(t) (e_i + e_j)}{2} - \frac{e_i^T P(t) e_i}{2} - \frac{e_j^T P(t) e_j}{2}$$

$$\begin{array}{ccc} t \rightarrow +\infty \downarrow & & t \rightarrow +\infty \downarrow & & t \rightarrow +\infty \downarrow \\ & & & & \end{array}$$

$$e_i^T P(t) e_j = (i, j)\text{-TH ELEMENT OF } P(t)$$

$$= \frac{(e_i + e_j)^T P(t) (e_i + e_j)}{2} - \frac{e_i^T P(t) e_i}{2} - \frac{e_j^T P(t) e_j}{2}$$

$t \rightarrow +\infty$

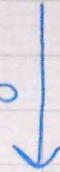


$$\frac{1}{2} (e_i + e_j)^T \bar{P} (e_i + e_j)$$

BY \otimes WITH

$$x = e_i + e_j$$

$t \rightarrow +\infty$

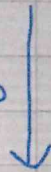


$$\frac{1}{2} e_i^T \bar{P} e_i$$

BY \otimes
WITH

$$x = e_i$$

$t \rightarrow +\infty$



$$\frac{1}{2} e_j^T \bar{P} e_j$$

BY \otimes
WITH

$$x = e_j$$

THUS,

$$[P(t)]_{ij} \rightarrow [\bar{P}]_{ij} \quad \text{SINCE}$$

$$\frac{1}{2} (e_i + e_j)^T \bar{P} (e_i + e_j) - \frac{1}{2} e_i^T \bar{P} e_i - \frac{1}{2} e_j^T \bar{P} e_j$$

|||

$$[\bar{P}]_{ij}$$

CONCLUDING REMARKS

- DETECTABILITY AND STABILIZABILITY ARE ALSO NECESSARY.

E.G. IF ξ IS NOT DETECTABLE, SOME UNSTABLE MODES ARE NOT SEEN IN THE OUTPUT. THEN, AN I.C. WHICH MAKES THOSE MODES ACTIVE WILL MAKE VARIANCE OF PREDICTION ERROR $P(t)$ GROW TO INFINITY.

(EVEN IF THE I.C. OF THE NON OBS. SYSTEM WERE KNOWN, THE ERROR COULD DIVERGE IF THE TRANSITION NOISE EXCITES THE UNSTABLE PART)

- IF $v(t)$ AND $w(t)$ ARE CORRELATED, THE NECESSARY AND SUFFICIENT CONDITIONS BECOME

< Note Fine

- IF $v(t)$ AND $w(t)$ ARE CORRELATED,
THE NECESSARY AND SUFFICIENT
CONDITIONS BECOME

(F, C) DETECTABLE
 $(F, \tilde{Q}^{1/2})$ STABILIZABLE

WITH \tilde{F}, \tilde{Q} DEFINED SOME
LECTURES AGO (THEY ARE
OUTCOMES FROM THE
WHITENING OPERATION).

- THEOREM:

LET (A, B) BE STABILIZABLE.

THEN

(A, C) DETECTABLE $\iff \forall P_0, \exists U$ S.T.
 $P(t) \leq U \quad \forall t$

MEANING: POINT 1 SHOWED THAT

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DETECTABILITY IS SUFFICIENT FOR
EXISTENCE OF U .

STABILIZABILITY MAKES IT ALSO
NECESSARY SINCE IT GUARANTEES
THAT THE TRANSITION NOISE
CAN EXCITE THE UNSTABLE MODES
OF THE NON OBSERVABLE
SUBSYSTEM (IF STABILIZABILITY
HOLDS, THE UNREACHABLE MODES
INCLUDE ONLY STABLE MODES)

- THEOREM:

(A, C) $\implies \exists$ AT LEAST
DETECTABLE ONE ARE SOLUTION
(THE \bar{P} OBTAINED
WITH $P_0 = 0!$)

HOLDS, THE UNREACHABLE MODES
INCLUDE ONLY STABLE MODES)

- THEOREM:

(A, C) $\implies \exists$ AT LEAST
DETECTABLE ONE ARE SOLUTION
(THE \bar{P} OBTAINED
WITH $P_0 = 0!$)

- THEOREM:

(A, B) \implies THERE IS AT
MOST ONE
STABILIZABLE $\bar{P} = \bar{P}^T \geq 0$ ARE
SOLUTION
(NECESSARILY
STABILIZING)

KALMAN SMOOTHING FILTER

MEASUREMENTS ON A FIXED INTERVAL

$$y(1), y(2), \dots, y(N)$$

AND WE WANT

$$\hat{E} [x(t) | y(1), \dots, y(N)] = \hat{x}(t|N)$$

WHERE STILL
WE
HAVE

$$x(t+1) = Ax(t) + v(t), \quad v \sim (0, Q)$$

$$y(t) = Cx(t) + w(t), \quad w \sim (0, R)$$

$$v(t) \perp w(t)$$

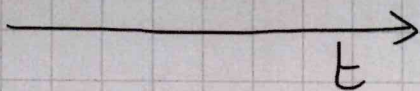
THE KALMAN FILTER

GIVES $\hat{x}(t|t)$ AND

$\hat{x}(t+1|t)$ USING

FORWARD RECURSIONS

FORWARD RECURSIONS



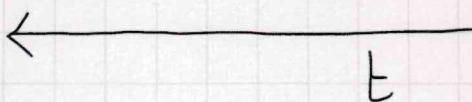
THEOREM:

ONE HAS

$$\hat{x}(t|N) = \hat{x}(t|t) + R_t (\hat{x}(t+1|N) - \hat{x}(t+1|t))$$

$$R_t = P(t|t) A^T P^{-1}(t+1|t)$$

BACKWARD RECURSIONS



PROOF:

LET US COMPUTE TWO PROJECTIONS

①

$$\hat{E} \left[x(t) \mid \underbrace{\{y(k)\}_{k=1}^t}, \underbrace{x(t+1) - \hat{x}(t+1|t)}, \underbrace{\{v(k), w(k)\}_{k=t+1}^N} \right]$$

THREE UNCORRELATED BLOCKS

$$\hat{E} \left[x(t) \mid \{y(k)\}_{k=1}^t \right] + \hat{E} \left[x(t) \mid x(t+1) - \hat{x}(t+1|t) \right]$$

$$= \underbrace{\hat{E} \left[x(t) \mid \{y(k)\}_{k=1}^t \right]}_{\hat{x}(t|t)} + \underbrace{\hat{E} \left[x(t) \mid x(t+1) - \hat{x}(t+1|t) \right]}_{\textcircled{*} \text{ SEE BELOW!}}$$

$$+ \underbrace{E \left[x(t) \mid \{v(k), w(k)\}_{k=t+1}^N \right]}_0$$

LET $m(t+1) := x(t+1) - \hat{x}(t+1|t),$

THEN $\textcircled{*}$
 \Downarrow
 $\text{VAR}[m(t+1)] = P(t+1|t)$
 |||

$$\text{COV}(x(t), m(t+1)) \text{VAR}^{-1}(m(t+1)) m(t+1)$$

$$= \text{COV} \left(x(t) - \hat{x}(t|t) + \hat{x}(t|t), A \underbrace{(x(t) - \hat{x}(t|t))}_{\perp \hat{x}(t|t)} + \underbrace{w(t)}_{\perp \text{ TO ALL THE LHS}} \right)$$

$$\bullet P^{-1}(t+1|t) m(t+1)$$

$$= P(t|t) A^T P^{-1}(t+1|t) m(t+1)$$

SO, LETTING

SO, LETTING

$$S = \text{SPAN} \left\{ \left\{ y(k) \right\}_{k=1}^t, m(t+1), \left\{ v(k), w(k) \right\}_{k=t+1}^N \right\}$$

WE HAVE OBTAINED

$$\hat{E}[x(t) | S] = \hat{x}(t|t) + P(t|t) A^T P^{-1}(t+1|t) m(t+1)$$

② THE SECOND PROJECTION
(FINAL STEP)

$$A = \text{SPAN} \{ y(1), \dots, y(N) \}$$



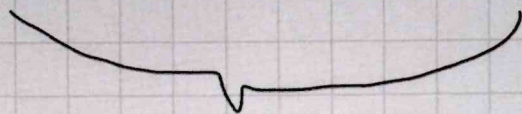
$$A \subseteq S \quad (\text{SEE NOTE AT THE END})$$



$$\hat{x}(t|N)$$

$$\hat{E}[x(t) | A] = \hat{E} \left[\underbrace{E[x(t) | S]}_{\text{}} \middle| A \right]$$

$$\hat{E}^{\parallel} [x(t) | A] = \hat{E} [E[x(t) | S] | A]$$



|||

$$\hat{x}(t|t) + P(t|t) A^T P^{-1}(t+1|t) (\hat{x}(t+1|N) - \hat{x}(t+1|t))$$

NOT AFFECTED
BY THE PROJECTION
ONTO A SINCE IT
ALREADY BELONGS
TO A

$x(t+1)$
PROJECTED
ONTO A

NOT
AFFECTED
BY THE
PROJ.
ONTO
A

$$\hat{x}(t|t), \in \text{SPAN}\{y(1), \dots, y(t)\}$$

$$\hat{x}(t+1|t)$$

WHY $A \subseteq S$?

$$S = \text{SPAN} \left\{ \underbrace{\{y(k)\}_{k=1}^t}_{\in \text{SPAN}\{y(1), \dots, y(t)\}}, \underbrace{x(t+1) - \hat{x}(t+1|t)}_{\in \text{SPAN}\{y(1), \dots, y(t)\}}, \underbrace{\{v(k), w(k)\}_{k=t+1}^N}_{k=t+1} \right\}$$

ABLE TO GENERATE
ALL

$$\text{SPAN}\{y(1), \dots, y(t), x(t+1)\}$$

ONTO A SINCE IT
ALREADY BELONGS
TO A

PROJ.
ONTO
A

$$\hat{x}(t|t), \in \text{SPAN}\{y(1), \dots, y(t)\}$$
$$\hat{x}(t+1|t)$$

WHY $A \subseteq S$?

$$S = \text{SPAN}\left\{ \underbrace{\{y(k)\}_{k=1}^t}_{\in \text{SPAN}\{y(1), \dots, y(t)\}}, x(t+1) - \hat{x}(t+1|t), \underbrace{\{v(k), w(k)\}_{k=t+1}^N}_{k=t+1} \right\}$$

ABLE TO GENERATE
ALL

$$\text{SPAN}\{y(1), \dots, y(t), x(t+1)\}$$

ABLE TO GENERATE
ALL

$$\text{SPAN}\{y(1), \dots, y(N)\} = A$$



$$A \subseteq S$$