formatique et d'Automatique in France, where he worked on the mathematical theory of infinite-dimensional systems. At Georgia Tech, he has been the principal investigator for several research grants dealing with the theory of infinite-dimensional systems (including systems with
time delays), linear time-varying systems, linear two-dimensional systems, and bilinear systems.

Dr. Kamen is a member of Sigma Xi, Eta Kappa Nu, Tau Beta Pi', and Phi Kappa Phi.

# A Stochastic Realization Approach to the Smoothing Problem 

FARIS A. BADAWI, ANDERS LINDQUIST, MEMBER, IEEE, AND MICHELE PAVON


#### Abstract

The purpose of this paper is to develop a theory of smoothing for finite dimensional linear stochastic systems in the context of stochastic realization theory. The basic idea is to embed the given stochastic system in a class of similar systems all having the same output process and the same Kalman-Bucy filter. This class has a lattice structure with a smallest and a largest element; these two elements completely determine the smoothing estimates. This approach enables us to obtain stochastic interpretations of many important smoothing formulas and to explain the relationship between them.


## I. Introduction

LET $\{x(t) ; 0 \leqslant t \leqslant T\}$ and $\{y(t) ; 0 \leqslant t \leqslant T\}$ be two stochastic vector processes, of dimensions $n$ and $m$, respectively, defined as the solution of the linear system of stochastic differential equations

$$
(S) \begin{cases}d x=A(t) x(t) d t+B(t) d w ; & x(0)=\xi  \tag{1.1a}\\ d y=C(t) x(t) d t+D(t) d w ; & y(0)=0\end{cases}
$$

where $w$ is a vector process, of dimension $p \geqslant m$, with orthogonal increments such that

$$
\begin{equation*}
E\{d w\}=0 ; \quad E\left\{d w d w^{\prime}\right\}=I d t \tag{1.2}
\end{equation*}
$$

(prime denotes transposition), $\xi$ is a centered random vector with finite covariance $\Pi:=E\left\{\xi \xi^{\prime}\right\}$ and uncorrelated with $w, R(t):=D(t) D(t)^{\prime}$ is positive definite on $[0, T]$, and $A, B, C, D$, and $R^{-1}$ are matrices of analytic functions defined on $[0, T]$. The model $S$ is usually called a linear stochastic system; $y$ is its output process, $w$ is its input process, and $x$ its state process. We shall assume that the representation $S$ is minimal in the sense that there is

[^0]no other model of the form (1.1) with the process $y$ as its output and with a state process $x$ of smaller dimension than $n$. Clearly, the matrix function $P(t):=E\left\{x(t) x(t)^{\prime}\right\}$ satisfies the differential equation
\[

$$
\begin{equation*}
\dot{P}=A P+P A^{\prime}+B B^{\prime} ; \quad P(0)=\Pi \tag{1.3}
\end{equation*}
$$

\]

on $[0, T]$. We shall call $P$ the state covariance function of $S$.
The following problem is of considerable importance in the systems sciences. For an arbitrary $t \in[0, T]$, find the linear least-squares estimate $\hat{x}(i)$ of the state vector $x(t)$ given the output $\{y(\tau) ; 0 \leqslant \tau \leqslant T\}$, i.e., the wide-sense conditional expectation

$$
\begin{equation*}
\hat{x}(t)=\hat{E}\{x(t) \mid y(\tau) ; 0 \leqslant \tau \leqslant T\} \tag{1.4}
\end{equation*}
$$

in the terminology of Doob [1]. This is the smoothing problem, and it has generated a rather extensive literature [2]-[17]. (See the survey paper [18] for further references.) Here we shall study this problem from a new angle employing concepts and techniques from the stochastic realization theory developed in [20]-[22] and more recently in [23]-[33]. The basic idea consists in embedding the model (1.1) into a class $\delta$ of models $S$ all having the same process $y$ as its output and all having the same Kalman-Bucy filter. Such a representation is called a stochastic realization of $y$. (Note that we only consider proper realizations [20], i.e., models $S$ whose outputs not merely have the same covariance properties, the only requirement in the earlier realization theory [34]-[38], but are equal for each $t$ a.s.) It can be seen that, slightly extended, the class $\mathfrak{S}$ has a lattice structure with a smallest ( $S_{*}$ ) and a largest ( $S^{*}$ ) element, the partial ordering being induced by the "size" of the covariance matrix $P(t)$ in the sense that $P_{1}>P_{2}$ if $P_{1}-P_{2}$ is positive definite. This approach will enable us to obtain stochastic interpretations of many important smoothing formulas and lay the groundwork for a theory of smoothing which so far has been lacking.

Our interest in the smoothing problem was caused by the Mayne-Fraser two-filter formula [5], [6], on which topic a large number of papers have been written [7]-[9], [12]-[17]. In some of these papers the authors have encountered difficulties in motivating this formula, and the many attempts to justify it stochastically have, in our opinion, been less than convincing. ${ }^{1}$ In our stochastic realization setting the two filters have a natural interpretation: they are simply the minimum and maximum variance realizations $S_{*}$ and $S^{*}$, respectively. Hence, the latter is not a "backward filter" as suggested in the literature (although it can be reformulated as such), but a "forward filter" just as its structure suggests.
At first sight some of the technical assumptions above may seem rather stringent, namely the minimality condition and the analyticity of the coefficient matrices. These conditions are introduced to ensure that, for each $S \in \mathcal{\delta}$, the state covariance matrix $P(t)$ is invertible for each $t \in(0, T)$. It is quite probable that these assumptions can be relaxed, ${ }^{2}$ but our objective here is to convey some basic ideas, and we do not want to obscure matters by introducing extra difficulties of a purely technical nature. On the other hand, the model (1.1) is more general than the one usually encountered in the smoothing literature in that $B d w$ and $D d w$ may be correlated. There is a reason for this, too. To limit our analysis to models $S$ for which $B D^{\prime}=0$ would render the class $\delta$ incomplete.
The contents of the paper are as follows. Section II is devoted to some preliminary results. We present a strict sense version of some results on backward Markovian representations developed, for much the same purposes, in [15], [16]. The idea of proof is borrowed from [20]. In Section III we define the stochastic realization setting mentioned above, and in Section IV we apply it to derive and interpret various smoothing procedures.

## II. Preliminaries

Let $H$ be the space of all centered stochastic variables (on an underlying probability space) with finite secondorder moments. Then $H$ is a Hilbert space with inner product $(\xi, \eta)=E\{\xi \eta\}$. For an arbitrary $k$-dimensional stochastic process $\{z(t) ; 0 \leqslant t \leqslant T\}$ with components in $H$, define $H_{t}(z)$ to be the (closed) subspace spanned by the random variables $\left\{z_{1}(t), z_{2}(t), \cdots, z_{k}(t)\right\}$, and let $H(z)$ be the closed linear hull in $H$ of the subspaces $\left\{H_{t}(z)\right.$; $0<t \leqslant T\}$; we shall write this as $H(z)=\vee_{t \in[0, T]} H_{t}(z)$. Similarly define the past space $H_{t}^{-}(z):=\vee_{\tau \in[0, t]} H_{\tau}(z)$ and the future space $H_{t}^{+}(z):=\bigvee_{\tau \in[t, T]} H_{\tau}(z)$. Sometimes we shall be more interested in spaces spanned by the increments of $z$. Hence, we define $H(d z), H_{t}{ }^{-}(d z)$, and $H_{t}{ }^{+}(d z)$ to be the closed linear hulls in $H$ of $\{z(\tau)-z(\sigma) ; \tau, \sigma \in I\}$ where $I$ is the interval $[0, T],[0, t]$, and $[t, T]$, respectively.
For each $\eta \in H$ and subspace $K \subset H$ let $\hat{E}\{\eta \mid K\}$ be the

[^1]projection of $\eta$ onto $K$, i.e., the wide-sense conditional mean [1]. Let $u$ be a stochastic vector with components in $H$, and let $H(u)$ be the closed linear span in $H$ of the components of $u$. Then, for any $\eta \in H$, we shall often write $\hat{E}\{\eta \mid u\}$ in place of $\hat{E}\{\eta \mid H(u)\}$, and, for any subspace $K \subset H, \hat{E}\{u \mid K\}$ will denote the vector with components $\hat{E}\left\{u_{i} \mid K\right\}$. We shall need the following lemma, the proof of which can be found in most standard texts on estimation theory.

Lemma 2.1: Let $u$ and $v$ be two stochastic vectors with components in $H$ and assume that $E\left\{v v^{\prime}\right\}$ is positive definite. Then

$$
\begin{equation*}
\hat{E}\{u \mid v\}=E\left\{u v^{\prime}\right\}\left(E\left\{v v^{\prime}\right\}\right)^{-1} v . \tag{2.1}
\end{equation*}
$$

The state process $x$ defined by (1.la) is a wide-sense Markov process [1], i.e.,

$$
\begin{equation*}
\hat{E}\left\{x(t) \mid H_{s}^{-}(x)\right\}=\hat{E}\{x(t) \mid x(s)\} \quad \text { for } t \geqslant s . \tag{2.2}
\end{equation*}
$$

To see this, merely note that $x(t)$ can be written

$$
\begin{equation*}
x(t)=\Phi(t, s) x(s)+\int_{s}^{t} \Phi(t, \tau) B(\tau) d w \tag{2.3}
\end{equation*}
$$

and that $H_{s}^{+}(d w) \perp H_{0}(x) \oplus H_{s}^{-}(d w) \supset H_{s}^{-}(x)$. (The symbol $\perp$ denotes "orthogonal to.") Here, of course, $\Phi$ is the transition matrix defined by

$$
\begin{equation*}
\frac{\partial \Phi}{\partial t}(t, s)=A(t) \Phi(t, s) ; \quad \Phi(s, s)=I . \tag{2.4}
\end{equation*}
$$

In deriving the main results of this paper we shall need to reverse the direction of time in (1.1). The Markov property is independent of the choice of time direction and therefore we also have

$$
\begin{equation*}
\hat{E}\left\{x(s) \mid H_{t}^{+}(x)\right\}=\hat{E}\{x(s) \mid x(t)\} \quad \text { for } t \geqslant s \tag{2.5}
\end{equation*}
$$

(In the present setting this can be seen by observing that, in view of (2.3), $H_{t}^{+}(x) \ominus H_{t}(x) \subset H_{t}^{+}(d w) \perp x(s)$.) The differential equation (1.1a), however, is not symmetric with respect to time; the two terms in the right member of (2.3) are orthogonal if and only if $t \geqslant s$. Hence, we need to define a backward version of (1.1a). This requires the inversion of the covariance matrix $P(t)$, which is the topic of the following lemma. Here and in the sequel $Q>0$ ( $Q \geqslant 0$ ) means that the symmetric matrix $Q$ is positive (nonnegative) definite.

Lemma 2.2: Let $P$ be the state covariance function of the linear stochastic system $S$ defined in Section I. Then, for any $\epsilon>0, P^{-1}$ exists and is analytic on the interval $[\epsilon, T]$. If $\Pi>0$, the same holds for the complete interval $[0, T]$.

Proof: Integrating (1.2) yields

$$
\begin{align*}
& P(t)=\Phi(t, 0) \Pi \Phi(t, 0)^{\prime} \\
&+\int_{0}^{t} \Phi(t, \tau) B(\tau) B(\tau)^{\prime} \Phi(t, \tau)^{\prime} d \tau \tag{2.6}
\end{align*}
$$

which is positive definite if $\Pi>0$; hence, since $A$ and $B$ are analytic on $[0, T]$, so is $P^{-1}$. Now assume that $\Pi \ngtr 0$. Since $S$ is minimal, ( $A, B$ ) must be completely controlla-
ble. In fact, were this not the case, the input-output map of $S$ could be reduced [39], contradicting minimality. Since in addition $A$ and $B$ are analytic, $(A, B)$ is totally controllable [40], [41]. Therefore, since the second term in (2.6) is the controllability gramian, $P(t)>0$ on any interval $[\epsilon, T]$. The analyticity of $P^{-1}$ then follows in the same way as above.

As we shall see below, it is more convenient to express the backward representation in terms of the process

$$
\begin{equation*}
\bar{x}(t)=P(t)^{-1} x(t) \tag{2.7}
\end{equation*}
$$

rather than $x$. In view of Lemma $2.2, \bar{x}(t)$ is well defined with components in $H$ on any interval $[\epsilon, T]$. Let $\bar{P}$ denote its covariance function, i.e.,

$$
\begin{equation*}
\vec{P}(t)=E\left\{\bar{x}(t) \bar{x}(t)^{\prime}\right\} \tag{2.8}
\end{equation*}
$$

We are now in a position to formulate a backward version of the state equation (1.1a).

Lemma 2.3: Let $x$ be the state process of the linear stochastic system $S$. Then, for any $\epsilon>0$, the process $\bar{x}$ defined by (2.7) satisfies the backward model

$$
\begin{equation*}
d \bar{x}=-A(t)^{\prime} \bar{x}(t) d t+\bar{B}(t) d \bar{w} ; \quad \bar{x}(T)=\bar{\xi} \tag{2.9}
\end{equation*}
$$

on $[\epsilon, T]$, where $\bar{\xi}=P(T)^{-1} x(T), \bar{B}=P^{-1} B$, and $\bar{w}$ is a p-dimensional orthogonal increment process satisfying (1.2) and the condition $H_{t}^{-}(d \bar{w}) \perp H_{t}^{+}(\bar{x})$ for all t. The increments of $\bar{w}$ are given by

$$
\begin{equation*}
d \bar{w}=d w-B(t)^{\prime} P(t)^{-1} x(t) d t \tag{2.10}
\end{equation*}
$$

and the covariance function (2.8) by $\bar{P}=P^{-1}$, it satisfies the Lyapunov equation

$$
\begin{equation*}
\dot{\bar{P}}=-A^{\prime} \bar{P}-\bar{P}_{A}-\bar{B} \bar{B}^{\prime} ; \quad \bar{P}(T)=\bar{\Pi} \tag{2.11}
\end{equation*}
$$

where $\bar{\Pi}=P(T)^{-1}$. If $\Pi>0,(2.9)-(2.11)$ are defined on the whole interval $[0, T]$.

Lemma 2.3 is a strict sense version of a similar result presented in [15], [16]. As explained in [42], an alternative justification of the wide-sense results [15], [16] can be obtained by means of the earlier work [12], [13]. The version given in all these papers is, however, insufficient for our purposes since it provides a deterministic rather than a probabilistic result. Moreover, we have chosen to write the backward equation in terms of $\bar{x}$ rather than $x$ as in [15], [16]. (However, see the "adjoint" formulation in [16].) The reason for this will become evident in Section III. Our choice will yield a backward Kalman-Bucy filter which is invariant over the class $\delta$, the one in [15], [16] will not.

The proof of Lemma 2.3 follows exactly the same lines as in [20]. It is based on the observation that, for $s \leqslant t$, the orthogonal decomposition

$$
\begin{equation*}
\bar{x}(s)=\hat{E}\left\{\bar{x}(s) \mid H_{t}^{+}(x)\right\}+\left[\bar{x}(s)-\hat{E}\left\{\bar{x}(s) \mid H_{t}^{+}(x)\right\}\right] \tag{2.12}
\end{equation*}
$$

can be written in the form

$$
\begin{equation*}
\bar{x}(s)=\Phi(t, s)^{\prime} \bar{x}(t)+\int_{t}^{s} \Phi(\tau, s)^{\prime} \bar{B}(\tau) d \bar{w} \tag{2.13}
\end{equation*}
$$

which is the integral form of (2.9).
Proof of Lemma 2.3: In view of Lemma 2.2, the state covariance function $P$ is invertible on the stated interval. Clearly, $\bar{P}=P^{-1}$. Then, since $\bar{P}=-\bar{P} \dot{P} \bar{P}$, (2.11) follows from (1.3). Then Lemma 2.1 together with (2.5) and (2.7) yields

$$
\begin{equation*}
\hat{E}\left\{\bar{x}(s) \mid H_{t}^{+}(x)\right\}=\Phi(t, s)^{\prime} \bar{x}(t) \tag{2.14}
\end{equation*}
$$

for it follows from (2.3) that $E\left\{x(s) x(t)^{\prime}\right\}=P(s) \Phi(t, s)^{\prime}$ for $s \leqslant t$. Consequently, the process $u(t):=\Phi(t, 0)^{\prime} \bar{x}(t)$ is a wide-sense backward martingale with respect to $H_{t}^{+}(x)$, i.e.,

$$
\begin{equation*}
\hat{E}\left\{u(s) \mid H_{t}^{+}(x)\right\}=u(t) \quad \text { for } s \leqslant t \tag{2.15}
\end{equation*}
$$

and hence it has orthogonal increments. We shall now show that $u$ can be normalized as follows:

$$
\begin{equation*}
u(s)-u(t)=\int_{t}^{s} \Phi(\tau, 0)^{\prime} \bar{B}(\tau) d \bar{w} \tag{2.16}
\end{equation*}
$$

where $\bar{w}$ is defined by (2.10). To this end differentiate $u(t)=\Phi(t, 0)^{\prime} \bar{P}(t) x(t)$ and use (1.1a), (2.4), and (2.11) to obtain $d u=\Phi(t, 0)^{\prime} \bar{B}\left(d w-B^{\prime} \bar{P} x d t\right)$. It remains to show that $\bar{w}$ is an orthogonal increment process satisfying (1.2). This follows from a tedious but straightforward calculation of the incremental covariance function. (If $B$ were full rank, we could conclude this directly from the martingale property ( 2.15 ); this could be achieved by working with the complete system $S$ instead.) The desired representation is then obtained by noting that

$$
\bar{x}(s)=\Phi(0, s)^{\prime}[u(t)+u(s)-u(t)]
$$

into which we insert (2.16) to obtain (2.13). Obviously, $H_{t}^{+}(\bar{x}) \perp H_{t}^{-}(d \bar{w})$, for, by construction, the two terms in (2.13) are orthogonal for all $t$.

## III. Forward and Backward Stochastic Realizations

Let $\{y(t) ; 0 \leqslant t \leqslant T\}$ be an $m$-dimensional vector process defined as the output of the linear stochastic system $S$ introduced in Section I. Any system of type (1.1) [with $\xi \in H, w$ satisfying (1.2) and $\xi \perp H(d w)]$ having the given process $y$ as its output is called a realization of $y$. In particular, by assumption, $S$ is minimal, i.e., there is no other realization of $y$ with a state process of smaller dimension, ${ }^{3}$ and analytic, i.e., its parameter matrices $A, B$, $C, D$, and $R^{-1}$ are analytic on $[0, T]$. Clearly, the components of $x(t)$ and $y(t)$ belong to $H$ for all $t \in[0, T]$, and the same holds for the increments of $w$.

It is well-known that the least-squares estimate

[^2]\[

$$
\begin{equation*}
x_{*}(t)=\hat{E}\left\{x(t) \mid H_{t}^{-}(d y)\right\} \tag{3.1}
\end{equation*}
$$

\]

of the state process $x$ of $S$ is generated on $[0, T]$ by the Kalman-Bucy filter

$$
\begin{equation*}
d x_{*}=A x_{*} d t+B_{*} R^{-1 / 2}\left(d y-C x_{*} d t\right) ; \quad x_{*}(0)=0 \tag{3.2a}
\end{equation*}
$$

where $R^{1 / 2}(t)$ is the symmetric square root of $R(t)=$ $D(t) D(t)^{\prime}$, and the gain function $B_{*}$ is given by

$$
\begin{equation*}
B_{*}=\left(Q_{*} C^{\prime}+B D^{\prime}\right) R^{-1 / 2} \tag{3.2b}
\end{equation*}
$$

the error covariance matrix

$$
\begin{equation*}
Q_{*}(t)=E\left\{\left[x(t)-x_{*}(t)\right]\left[x(t)-x_{*}(t)\right]^{\prime}\right\} \tag{3.2c}
\end{equation*}
$$

being the solution of the matrix Riccati equation

$$
\left\{\begin{align*}
\dot{Q}_{*}= & A Q_{*}+Q_{*} A^{\prime}-\left(Q_{*} C^{\prime}+B D^{\prime}\right) R^{-1}  \tag{3.2d}\\
& \cdot\left(Q_{*} C^{\prime}+B D^{\prime}\right)^{\prime}+B B^{\prime} \\
Q_{*}(0)= & \Pi
\end{align*}\right.
$$

As we shall see shortly there are other realizations which have the same Kalman-Bucy filter (3.2a). Hence, we define $\mathcal{S}$ to be the class of all analytic realizations $S$ of $y$ whose Kalman-Bucy filter, determined as in (3.2), has the same coefficient functions $A, C, R$, and $B_{*}$ as in (3.2a). Then (since we only consider proper [20] realizations) the estimates $x_{*}$ are also the same. (The error covariance $Q_{*}$, however, will of course vary over §.) Clearly, all realizations in $\mathfrak{S}$ are minimal. Moreover, it is well-known that the innovation process $\left\{w_{*}(t) ; 0 \leqslant t \leqslant T\right\}$, whose increments are defined by

$$
\begin{equation*}
d w_{*}=R^{-1 / 2}\left(d y-C x_{*} d t\right) \tag{3.3}
\end{equation*}
$$

is a process with orthogonal increments satisfying (1.2) and $H_{t}^{-}\left(d w_{*}\right)=H_{t}^{-}(d y)$ for all $t \in[0, T]$ (see, e.g., [43]). Then (3.2a) and (3.3) yield

$$
\left(S_{*}\right)\left\{\begin{array}{l}
d x_{*}=A x_{*} d t+B_{*} d w_{*} ; \quad x_{*}(0)=0  \tag{3.4}\\
d y=C x_{*} d t+R^{1 / 2} d w_{*}
\end{array}\right.
$$

Since $B_{*}$ is analytic on $[0, T]$, this is a realization in $\delta$ whose state covariance matrix $P_{*}(t):=E\left\{x_{*}(t) x_{*}(t)^{\prime}\right\}$ satisfies

$$
\begin{equation*}
\dot{P}_{*}=A P_{*}+P_{*} A^{\prime}+B_{*} B_{*}^{\prime} ; \quad P_{*}(0)=0 \tag{3.5}
\end{equation*}
$$

(This can also be seen by subtracting (1.3) from (3.2d), noting that $Q_{*}=P-P_{*}$.) Now define the $n \times m$ matrix function

$$
\begin{equation*}
G=P_{*} C^{\prime}+B_{*} R^{1 / 2} \tag{3.6}
\end{equation*}
$$

which is clearly analytic on $[0, T]$.
Lemma 3.1: Let $\boldsymbol{G}$ be defined by (3.6). Then for any realization $S \in \delta$,

$$
\begin{equation*}
P(t) C(t)^{\prime}+B(t) D(t)^{\prime}=G(t) \tag{3.7}
\end{equation*}
$$

for all $t \in[0, T]$.
Proof: This follows from (3.2b) and the fact that $Q_{*}=P-P_{*}$.

Consequently, $A, C, G$, and $R$ are invariants for the class $\mathcal{S}$-in fact, the covariance function of $y$ is determined by these four matrix functions [37], [44]-whereas $B, D, P, w$, and $x$ will vary with different realizations $S$. Actually, even the dimension $p$ of the process $w$ will vary. However, since $R$ is full rank, we will always have $p \geqslant m$.

The Kalman-Bucy filter realization $S_{*}$ belongs to a class of realizations for which $p$ is minimal, i.e., $p=m$. Define $\delta_{0}$ to be the subclass of all $S \in \mathcal{S}$ such that $p=m$ and $x(0) \in H(d y)$. Note that, since $y(0)=0, H(d y)=H(y)$. We shall use the former notation as we are really only interested in the increments of $y$, the assumption $y(0)=0$ being one of convenience.) Let

$$
\left(S_{0}\right)\left\{\begin{array}{l}
d x_{0}=A x_{0} d t+B_{0} d w_{0} ; \quad x_{0}(0)=\xi_{0}  \tag{3.8}\\
d y=C x_{0} d t+D_{0} d w_{0}
\end{array}\right.
$$

be a realization in $\mathscr{S}_{0}$ with state covariance function $P_{0}$. Then $D_{0}$ is invertible and therefore

$$
\begin{equation*}
d x_{0}=A x_{0} d t+B_{0} D_{0}^{-1}\left(d y-C x_{0} d t\right) ; \quad x_{0}(0)=\xi_{0} . \tag{3.9a}
\end{equation*}
$$

Now let (1.1) be an arbitrary realization in $\delta$ and define

$$
\begin{equation*}
Q_{0}=P-P_{0} \tag{3.9b}
\end{equation*}
$$

Then Lemma 3.1 yields

$$
\begin{equation*}
B_{0}=\left(Q_{0} C^{\prime}+B D^{\prime}\right)\left(D_{0}^{\prime}\right)^{-1} \tag{3.9c}
\end{equation*}
$$

where $Q_{0}$ satisfies the matrix Riccati equation

$$
\left\{\begin{align*}
& \dot{Q}_{0}=A Q_{0}+Q_{0} A^{\prime}-\left(Q_{0} C^{\prime}+B D^{\prime}\right) R^{-1}  \tag{3.9d}\\
& \cdot\left(Q_{0} C^{\prime}+B D^{\prime}\right)^{\prime}+B B^{\prime} \\
& Q_{0}(0)=\Pi-\Pi_{0}
\end{align*}\right.
$$

To see this just insert (3.9c) into the equation (1.3) corresponding to $P_{0}$ and subtract from (1.3). Formally, (3.9) looks precisely like the Kalman-Bucy filtering equations (3.2). In fact, the differential equations are the same, only the initial conditions differ. However, note that, unlike $Q_{*}, Q_{0}$ is in general indefinite due to the definition of $Q_{0}(0)$. In view of the fact that $\xi_{0} \in H(d y)$, (3.9a) implies that $H\left(x_{0}\right) \subset H(d y)$. We shall call a realization $S \in \mathcal{S}$ satisfying the condition $H(x) \subset H(d y)$ internal; if $H(x) \ddagger$ $H(d y)$ we shall say that $S$ is external [20]. Hence we have shown that all $S \in \delta_{0}$ are internal. In Section IV we shall see that, if $\binom{B}{D}$ is full rank, $\delta_{0}$ is precisely the class of all internal realizations. We shall also see that $\delta_{0}$ is a partially ordered set with a smallest element and that it can be slightly extended to also contain a largest element.

Our next task is to establish a backward counterpart $\bar{S}$ to each realization $S \in \mathcal{\delta}$. We shall begin by restricting
our attention to the subclass $\mathcal{S}_{+}$of all realizations $S \in \mathcal{S}$ for which $\Pi>0$.

Lemma 3.2: The class $\mathfrak{S}_{+}$is nonempty.
Proof: It is shown in [44] that, since $y$ is generated by the model (1.1), for some $\epsilon>0$ the covariance function of $y$ can be continuously extended to the interval $[0, T+\epsilon]$ while retaining its nonnegativity property and its "lumped" character. It is not hard to modify the proof of [44, Appendix II] to show that a similar extension, which also preserves analyticity, can be made to the interval [ $-\epsilon, T$ ] for some $\epsilon>0$. Hence, by the main result of [44], there is an (analytic) realization $S_{\epsilon}$ of $y$ on $[-\epsilon, T]$ with state-dimension $n$. Since its restriction to $[0, T]$ belongs to $\delta$, it is minimal. Therefore, $(A, B)$ corresponding to $S_{\epsilon}$ is totally controllable [40], and consequently $P(0)>0$ by the argument of Lemma 2.2. Hence, the restriction of $S_{\epsilon}$ to [ $0, T$ ] belongs to $\mathcal{S}_{+}$.

Let $S \in \mathcal{S}_{+}$. Then, by Lemma 2.3, $\bar{x}=P^{-1} x$ is defined on all of $[0, T]$ and satisfies (2.9) there. Inserting (2.10) into (1.1b) yields $d y=\left(C P+D B^{\prime}\right) \bar{x} d t+D d \bar{w}$, so in view of Lemma 3.1 we have obtained a backward model for $y$ on $[0, T]$, namely,

$$
(\bar{S})\left\{\begin{array}{l}
d \bar{x}=-A^{\prime} \bar{x} d t+\bar{B} d \bar{w} ; \quad \bar{x}(T)=\bar{\xi}  \tag{3.10}\\
d y=G^{\prime} \bar{x} d t+D d \bar{w}
\end{array}\right.
$$

where $\bar{\xi}=P(T)^{-1} x(T) \perp H(d \bar{w})$ and $\bar{B}=P^{-1} B$. Its state covariance function $\bar{P}=P^{-1}$ satisfies (2.11). We shall call any model of type (3.10) with $y$ as its output, $\bar{\xi} \in H, \bar{w}$ satisfying (1.2) and $\bar{\xi} \perp \underline{H}(d \bar{w})$ a backward realization of $y$. In view of Lemma $2.2, \bar{S}$ is also analytic (i.e., $A, \bar{B}, G, D$, and $R^{-1}$ are analytic). Note that $S$ and $\bar{S}$ have the same state space, i.e.,

$$
\begin{equation*}
H_{t}(\bar{x})=H_{t}(x) \tag{3.11}
\end{equation*}
$$

for each $t \in[0, T]$.
By symmetry with the forward setting we can now see that

$$
\begin{equation*}
\bar{x}_{*}(t)=\hat{E}\left\{\bar{x}(t) \mid H_{t}^{+}(d y)\right\} \tag{3.12}
\end{equation*}
$$

is generated by the backward Kalman-Bucy filter

$$
\begin{array}{r}
d \bar{x}_{*}=-A^{\prime} \bar{x}_{*} d t+\bar{B}_{*} R^{-1 / 2}\left(d y-G^{\prime} \bar{x}_{*} d t\right) \\
\bar{x}_{*}(T)=0 \tag{3.13a}
\end{array}
$$

where $\bar{B}_{*}=-\left(\bar{Q}_{*} G-\bar{B} D^{\prime}\right) R^{-1 / 2}$, and the error covariance $\bar{Q}_{*}(t):=E\left\{\left[\bar{x}(t)-\bar{x}_{*}(t)\right]\left[\bar{x}(t)-\bar{x}_{*}(t)\right]\right\}$ satisfies

$$
\left\{\begin{array}{c}
\dot{\bar{Q}}_{*}=-A^{\prime} \bar{Q}_{*}-\bar{Q}_{*} A+\left(\bar{Q}_{*} G-\bar{B} D^{\prime}\right)  \tag{3.13b}\\
\cdot R^{-1}\left(\bar{Q}_{*} G-\bar{B} D^{\prime}\right)-\bar{B} \bar{B}^{\prime} \\
\bar{Q}_{*}(T)=P(T)^{-1}
\end{array}\right.
$$

and that the backward innovation process $\left\{\bar{w}_{*}(t) ; 0 \leqslant t \leqslant\right.$ $T$ \}, given by

$$
\begin{equation*}
d \bar{w}_{*}=R^{-1 / 2}\left(d y-G^{\prime} \bar{x}_{*} d t\right) \tag{3.14}
\end{equation*}
$$

has orthogonal increments and satisfies (1.2) and $H_{t}^{+}\left(d \bar{w}_{*}\right)=H_{t}^{+}(d y)$ for all $t \in[0, T]$ (see [20], [45]). Hence the covariance function $\bar{P}_{*}(t):=E\left\{\bar{x}_{*}(t) \bar{x}_{*}(t)^{\prime}\right\}$ satisfies

$$
\begin{equation*}
\dot{\bar{P}}_{*}=-A^{\prime} \bar{P}_{*}-\bar{P}_{*} A-\bar{B}_{*} \bar{B}_{*}^{\prime} ; \quad \bar{P}_{*}(T)=0 \tag{3.15}
\end{equation*}
$$

The following lemma ensures the invariance of the backward filter (3.13a).

Lemma 3.3: The gain function $\bar{B}_{*}$ is uniquely determined by the four (invariant) matrix functions $A, C, G$, and $R$, i.e., $\bar{B}_{*}$ is invariant for $\mathcal{S}$.

Proof: Since $\bar{Q}_{*}=\bar{P}-\bar{P}_{*}$, it follows from Lemma 3.1 that $\bar{B}_{*}=\left(C^{\prime}-\bar{P}_{*} G\right) R^{-1 / 2}$, which inserted into (3.15) yields an equation for $\bar{P}_{*}$ which only depends on $A, C, G$, and $R$. Hence, the same holds for $\bar{B}_{*}$.

Now define $\overline{\mathcal{S}}$ to be the class of all analytic backward realizations $\bar{S}$ having (3.13a) as its backward Kalman-Bucy filter, and let $\overline{\mathcal{S}}_{+}$be the subclass consisting of those $\bar{S} \in \overline{\mathcal{S}}$ for which $\overline{I I}>0$. In the same way as in the forward setting it is seen that the realization

$$
\left(\bar{S}_{*}\right)\left\{\begin{array}{l}
d \bar{x}_{*}=-A^{\prime} \bar{x}_{*} d t+\bar{B}_{*} d \bar{w}_{*} ; \quad \bar{x}_{*}(T)=0  \tag{3.16}\\
d y=G^{\prime} \bar{x}_{*} d t+R^{1 / 2} d \bar{w}_{*}
\end{array}\right.
$$

belongs to $\overline{\mathcal{S}}$. The state covariance function $\bar{P}_{*}$ of $\bar{S}_{*}$ is given by ( 3.15 ). By Lemma 3.3 the class $\overline{\mathcal{E}}$ is uniquely defined in terms of the invariants $A, C, G$, and $R$, and therefore the backward counterpart $\bar{S}$ of any $S \in \mathcal{S}_{+}$ belongs to $\overline{\mathcal{S}}$. In particular, since $P(T)$ is finite and positive definite, $\bar{S} \in \overline{\mathcal{S}}_{+}$. Obviously, there is a complete symmetry between the forward and the backward settings; all results of this section have backward versions obtained by merely starting from a minimal backward realization instead. Consequently, all realizations in $\overline{\mathcal{\delta}}$ are minimal. (Indeed, were this not the case, we could, by reversing the procedure above, construct a forward realization of dimension less than $n$, contradicting our original minimality assumption.) Moreover, it is not hard to see that the two subclasses $\mathcal{S}_{+}$and $\overline{\mathcal{S}}_{+}$are in one-to-one correspondence.

In order to extend the one-to-one correspondence between forward and backward realizations beyond $\delta_{+}$and $\overline{\mathcal{S}}_{+}$we shall have to enlarge the classes $\mathcal{\delta}$ and $\overline{\mathcal{S}}$ slightly in the following way. Let $\hat{\mathcal{S}}$ be the class of all analytic realizations (1.1) of $y$ defined on $[0, T-\epsilon]$ for $a n y \epsilon>0$ and having (3.2a) for its Kalman-Bucy filter on each such interval, and let $\hat{\mathscr{S}}_{0_{\hat{e}}}$ be the corresponding extension of $\mathscr{S}_{0}$. The classes $\hat{\overline{\mathcal{S}}}$ and $\hat{\overline{\mathcal{S}}}_{0}$ are defined analogously with respect to (3.13a) and all intervals $[\epsilon, T]$. We shall call the elements of $\hat{\mathscr{\delta}}$ and $\hat{\bar{\delta}}$ generalized realizations. Clearly, $\mathcal{\delta} \subset \hat{\mathscr{E}}$ and $\overline{\mathcal{S}} \subset \stackrel{\overline{\mathcal{E}}}{ }$. Then the forward-backward construction above can be redone in the light of Lemma 2.3 to yield the following theorem, which also summarizes some of the pertinent facts on this topic.

Theorem 3.1: To each realization (1.1) in 5 there corresponds a generalized backward realization (3.10) in $\hat{\overline{\mathcal{S}}}$ such
that $\bar{P}=P^{-1}, \quad \bar{B}=P^{-1} B, \quad \bar{x}=P^{-1} x$, and $d \bar{w}=d w-$ $B^{\prime} P^{-1} x d t$. Likewise to each backward realization (3.10) in $\bar{\S}$ there is a generalized realization (1.1) in $\hat{\mathcal{E}}$ such that $P=\bar{P}^{-1}, B=\bar{P}^{-1} \bar{B}, x=\bar{P}^{-1} \bar{x}$, and $d w=d \bar{w}+\bar{B}^{\prime} \bar{P}^{-1} \bar{x} d t$. For each such pair $(S, \bar{S})$ of forward and backward (generalized) realizations, relation (3.11) holds for each $t$ for which both $S$ and $\bar{S}$ are defined.

Since $\bar{P}_{*}(T)=0$, the backward filter realization $\bar{S}_{*}$ has a forward counterpart only in this generalized sense, and it has the form

$$
\left(S^{*}\right)\left\{\begin{array}{l}
d x^{*}=A x^{*} d t+B^{*} d w^{*} ; \quad x^{*}(0)=\bar{P}_{*}(0)^{-1} \bar{x}_{*}(0)  \tag{3.17}\\
d y=C x^{*} d t+R^{1 / 2} d w^{*}
\end{array}\right.
$$

with state covariance function $P^{*}=\bar{P}_{*}^{-1}$ satisfying

$$
\begin{equation*}
\dot{P}^{*}=A P^{*}+P^{*} A^{\prime}+B^{*} B^{* \prime} ; \quad P^{*}(0)=\bar{P}_{*}(0)^{-1} \tag{3.18}
\end{equation*}
$$

on $[0, T]$. Obviously, $P^{*}(t) \rightarrow \infty$ as $t \rightarrow T$. (Note that $B^{*}$ is unbounded.) The following lemma explains the "super star" notation.

Lemma 3.4: Let $P$ be the state covariance function of $a$ realization $S \in \mathcal{S}$. Then

$$
\begin{equation*}
P_{*}(t) \leqslant P(t) \leqslant P^{*}(t) \tag{3.19}
\end{equation*}
$$

for all $t \in[0, T)$.
Proof: Since $Q_{*}(t)$ is a covariance matrix, $Q_{*}(t) \geqslant 0$. But $Q_{*}=P-P_{*}$, and therefore $P(t) \geqslant P_{*}(t)$. An analogous argument in the backward setting yields $\bar{P}(t) \geqslant \bar{P}_{*}(t)$, i.e., $P(t)^{-1} \geqslant P^{*}(t)^{-1}$, from which $P(t) \leqslant P^{*}(t)$ follows.

Relation (3.19) induces a partial ordering of $\hat{\mathcal{S}}, S_{*}$ being the smallest and $S^{*}$ the largest element; the same holds for $\hat{S}_{0}$, for both $S_{*}$ and $S^{*}$ belong to this subclass. (It can be shown that $\mathcal{S}$ and $\mathcal{S}_{0}$ have lattice structures, but this goes beyond the scope of this paper.) Since $S^{*} \in \hat{\mathscr{E}}_{0}, x^{*}$ satisfies a Kalman-Bucy type equation

$$
\begin{equation*}
d x^{*}=A x^{*} d t+B^{*} R^{-1 / 2}\left(d y-C x^{*} d t\right) ; \quad x^{*}(0)=\xi^{*} \tag{3.20a}
\end{equation*}
$$

where $\xi^{*}=\bar{P}_{*}(0)^{-1} \bar{x}_{*}(0)$, and $B^{*}$ can be determined from any other realization $S \in \delta$ through (3.9c), (3.9d), setting $B^{*}=B_{0}$ and $\Pi_{0}=\Pi^{*}$. The corresponding solution $Q_{0}$ of the matrix Riccati equation (3.9d) is, in view of (3.9b), $Q_{0}=P-P^{*}$, which is nonpositive definite (Lemma 3.4). For the smoothing problem it will be more convenient to express $B^{*}$ in terms of a nonnegative definite solution of (3.9d) instead, and therefore we define $Q^{*}:=-Q_{0}$, i.e.,

$$
\begin{equation*}
Q^{*}=P^{*}-P, \tag{3.20b}
\end{equation*}
$$

in terms of which (3.9c), (3.9d) yields

$$
\begin{equation*}
B^{*}=-\left(Q^{*} C^{\prime}-B D^{\prime}\right) R^{-1 / 2} \tag{3.20c}
\end{equation*}
$$

$$
\left\{\begin{array}{l}
\dot{Q}^{*}=A Q^{*}+Q^{*} A^{\prime}+\left(Q^{*} C^{\prime}-B D^{\prime}\right) R^{-1}  \tag{3.20d}\\
\quad \cdot\left(Q^{*} C^{\prime}-B D^{\prime}\right)^{\prime}-B B^{\prime} \\
Q^{*}(0)=\Pi^{*}-\Pi
\end{array}\right.
$$

where $\Pi^{*}=\bar{P}_{*}(0)^{-1}$. Clearly, $Q^{*}(t) \rightarrow \infty$ as $t \rightarrow T$. The definition (3.20b) enables us to interpret $Q^{*}$ as an error covariance function, much in analogy with the Kalman-Bucy filter. In fact,

$$
\begin{equation*}
Q^{*}(t)=E\left\{\left[x(t)-x^{*}(t)\right]\left[x(t)-x^{*}(t)\right]^{\prime}\right\} \tag{3.21}
\end{equation*}
$$

for all $t \in[0, T)$. This is an immediate consequence of the following lemma, which we shall also need in Section IV.

Lemma 3.5: Let $x$ be the state process and $P$ the state covariance function of any realization in $\mathcal{S}$. Then

$$
\begin{align*}
& E\left\{x(t) x_{*}(t)^{\prime}\right\}=P_{*}(t),  \tag{3.22}\\
& E\left\{x(t) x^{*}(t)^{\prime}\right\}=P(t)
\end{align*}
$$

and

$$
\begin{equation*}
E\left\{\left[x(t)-x_{*}(t)\right]\left[x^{*}(t)-x(t)\right]^{\prime}\right\}=0 . \tag{3.23}
\end{equation*}
$$

Proof: In view of the definition (3.1), $H_{t}(x-$ $\left.x_{*}\right) \perp H_{t}\left(x_{*}\right)$ and therefore the first of relations (3.22) follows. The analogous relation in the backward setting reads $E\left\{\bar{x}(t) \bar{x}_{*}(t)^{\prime}\right\}=\bar{P}_{*}$. Hence, $E\left\{x(t) \underline{x}^{*}(t)^{\prime}\right\}=$ $P E\left\{\bar{x}(t) \bar{x}_{*}(t)^{\prime}\right\} P^{*}=P$, for $\bar{x}=P^{-1} x$ and $P^{*}=\bar{P}_{*}^{-1}$. Then (3.23) is an immediate consequence of (3.22).

In Section IV we shall need to invert both $Q_{*}(t)$ and $Q^{*}(t)$ for arbitrary $t \in[0, T)$. This is possible for all realizations $S \in \mathcal{\delta}$ such that $P_{*}(t)<P(t)<P^{*}(t)$ for all $t$ on this interval. We shall call the class of all such $S$ the interior of $\mathcal{S}$ and denote it int $\mathcal{S}$.

Lemma 3.6: The interior of $\mathcal{S}$ is nonempty.
Proof: Let $Q_{*}$ be the error covariance (3.2c) corresponding to a realization $S \in \mathcal{S}_{+}$. Then $Q_{*}(0)=\Pi>0$. A simple reformulation of (3.2d) yields

$$
\begin{equation*}
\dot{Q}_{*}=\Gamma_{*} Q_{*}+Q_{*} \Gamma_{*}^{\prime}+\left(B_{*} R^{-1 / 2} D-B\right)\left(B_{*} R^{-1 / 2} D-B\right)^{\prime} \tag{3.24}
\end{equation*}
$$

where $\Gamma_{*}$ is the feedback matrix

$$
\begin{equation*}
\Gamma_{*}=A-B_{*} R^{-1 / 2} C \tag{3.25}
\end{equation*}
$$

of the Kalman-Bucy filter (3.2). The Lyapunov type equation (3.24) can be integrated to yield an expression of the same general form as (2.6). From this it is seen that $Q_{*}(0)>0$ implies that $Q_{*}(t)>0$ for all $t \in[0, T]$. It remains to show that $Q^{*}(t)>0$ for all $t \in[0, T)$. To this end first note that the corresponding backward realization $\bar{S}$ belongs to $\bar{\delta}_{+} ;$this is clear from the discussion leading to Theorem 3.1. Then we can repeat the argument above to see that $\bar{Q}_{*}(t)>0$ for all $t \in[0, T]$. But $Q^{*}=P\left(\bar{P}-\bar{P}_{*}\right) P^{*}$ $=P \bar{Q}_{*} P^{*}$. Since $P>0$ and $P^{*}>0$ on $[0, T), Q^{*}(t)>0$ for all $t \in[0, T)$.
Corollary 3.6.1: Let $Q=P^{*}-P_{*}$. Then $Q(t)>0$ for all $t \in[0, T)$.

Corollary 3.6.2: Let $S \in \mathcal{S}_{+}$. Then $Q_{*}(t)>0$ for all $t \in$ $[0, T]$.

We shall now demonstrate that the two processes $x_{*}$ and $x^{*}$ together contain all the relevant information on $y$ needed in estimating the state process $x$ of an arbitrary realization $S \in \mathcal{S}$. To this end first note that (3.1) can be written

$$
\begin{equation*}
\hat{E}\left\{H_{t}(x) \mid H_{t}^{-}(d y)\right\}=H_{t}\left(x_{*}\right), \tag{3.26}
\end{equation*}
$$

and that (3.11) and (3.12) yield

$$
\begin{equation*}
\hat{E}\left\{H_{t}(x) \mid H_{t}^{+}(d y)\right\}=H_{t}\left(x^{*}\right) \tag{3.27}
\end{equation*}
$$

for all $t \in[0, T)$. Now define the orthogonal complements $N_{t}^{-}:=H_{t}^{-}(d y) \ominus H_{t}\left(x_{*}\right)$ and $N_{t}^{+}:=H_{t}^{+}(d y) \ominus H_{t}\left(x^{*}\right)$, respectively. Then we obtain the orthogonal decomposition

$$
\begin{equation*}
H(d y)=N_{t}^{-} \oplus H_{t}^{\square} \oplus N_{t}^{+} \tag{3.28}
\end{equation*}
$$

where $H_{t}^{\mathrm{D}}$ is the frame space

$$
\begin{equation*}
H_{t}^{\square}=H_{t}\left(x_{*}\right) \vee H_{t}\left(x^{*}\right) \tag{3.29}
\end{equation*}
$$

(where $A \vee B$ denotes the closed linear hull in $H$ of $A$ and B) (cf. [22], [24], [26]).

Lemma 3.7. (cf. [27]): Let $x$ be the state process of $a$ realization in $\delta$. Then, for $t \in[0, T)$,

$$
H_{t}(x) \subset H_{t}^{\circ} \oplus[H(d y)]^{\perp}
$$

where $[H(d y)]^{\perp}$ is the orthogonal complement of $H(d y)$ in H.

Proof: Clearly, $H_{t}(x) \perp N_{t}^{-}$. To see this note that the components of $x(t)-x_{*}(t)$ are orthogonal to $H_{t}^{-}(d y) \supset$ $N_{t}^{-}$and that the components of $x_{*}(t)$ belong to $H_{t}\left(x_{*}\right) \perp N_{t}^{-}$. In the same way we show that $H_{t}(x) \perp N_{t}^{+}$.

## IV. The Smoothing Problem

Consider an arbitrary realization (1.1) in the class $\mathcal{S}$. The basic problem before us is to determine the smoothing estimate

$$
\begin{equation*}
\hat{x}(t)=\hat{E}\{x(t) \mid H(d y)\} \tag{4.1}
\end{equation*}
$$

for each $t \in[0, T)$ and to interpret it in terms of stochastic realizations. Let $\Sigma$ denote the corresponding estimation error covariance, i.e.,

$$
\begin{equation*}
\Sigma(t)=E\left\{[x(t)-\hat{x}(t)][x(t)-\hat{x}(t)]^{\prime}\right\} \tag{4.2}
\end{equation*}
$$

Of course, this problem is interesting only if the realization $S$ is external. However, by not restricting our analysis to external realizations, as a by-product we shall obtain some interesting results on internal models also.

In view of Lemma 3.7, $\hat{x}(t) \in H_{t}^{\text {口 }}$, and consequently there are two matrix functions $K_{*}$ and $K^{*}$ such that

$$
\begin{equation*}
\hat{x}(t)=K_{*}(t) x_{*}(t)+K^{*}(t) x^{*}(t) . \tag{4.3}
\end{equation*}
$$

The components of the estimation error $x(t)-\hat{x}(t)$ are clearly orthogonal to $H(d y)$ and hence in particular to the components of $x_{*}(t)$ and $x^{*}(t)$. Therefore, $E\left\{x(t) x_{*}(t)^{\prime}\right\}$ $=E\left\{\hat{x}(t) x_{*}(t)^{\prime}\right\}$ and $E\left\{x(t) x^{*}(t)^{\prime}\right\}=E\left\{\hat{x}(t) x^{*}(t)^{\prime}\right\}$. By Lemma 3.5, the first of these relations yields $P_{*}=K_{*} P_{*}+$ $K^{*} P_{*}$ and consequently,

$$
\begin{equation*}
K_{*}(t)+K^{*}(t)=I \tag{4.4}
\end{equation*}
$$

for all $t \in(0, T)$ because $P_{*}(t)$ is nonsingular on this interval. The second relation yields

$$
\begin{equation*}
P(t)=K_{*}(t) P_{*}(t)+K^{*}(t) P^{*}(t) \tag{4.5}
\end{equation*}
$$

for all $t \in[0, T)$. Then solving (4.4) and (4.5) for $K_{*}$ and $K^{*}$ we obtain $K_{*}=Q^{*} Q^{-1}$ and $K^{*}=Q_{*} Q^{-1}$, where as before $Q_{*}=P-P_{*}, Q^{*}=P^{*}-P$, and $Q=P^{*}-P_{*}$. Note that $Q(t)$ is nonsingular for all $t \in[0, T)$ (Corollary 3.6.1) and that

$$
\begin{equation*}
Q(t)=Q_{*}(t)+Q^{*}(t) \tag{4.6}
\end{equation*}
$$

Theorem 4.1: Let $x$ be the state process of a realization (1.1) of class $\mathfrak{S}$. Then the smoothing estimate (4.1) is given by

$$
\begin{equation*}
\hat{x}(t)=\left[I-Q_{*}(t) Q(t)^{-1}\right] x_{*}(t)+Q_{*}(t) Q(t)^{-1} x^{*}(t) \tag{4.7}
\end{equation*}
$$

and the error covariance function (4.2) by

$$
\begin{equation*}
\Sigma(t)=Q_{*}(t)-Q_{*}(t) Q(t)^{-1} Q_{*}(t) \tag{4.8}
\end{equation*}
$$

for all $t \in[0, T)$.
Proof: Relation (4.7) was derived above for $t \in(0, T)$; for $t=0$ (4.7) follows from (4.19) below. To prove (4.8) note that

$$
\begin{equation*}
x-\hat{x}=\left(I-Q_{*} Q^{-1}\right)\left(x-x_{*}\right)+Q_{*} Q^{-1}\left(x-x^{*}\right) \tag{4.9}
\end{equation*}
$$

By Lemma 3.5 the two terms of (4.9) are orthogonal and therefore, observing (3.2c) and (3.21),

$$
\Sigma=\left(I-Q_{*} Q^{-1}\right) Q_{*}\left(I-Q^{-1} Q_{*}\right)+Q_{*} Q^{-1} Q^{*} Q^{-1} Q_{*}
$$

which, in view of (4.6) yields (4.8).
Relation (4.5) should be compared with the decomposition in [46, Theorem 6]. Note, however, that $K_{*}(t)$ and $K^{*}(t)$ are projections if and only if the realization $S$ is internal. To see this observe that $\left(K^{*}\right)^{2}=K^{*}$, i.e., $Q_{*} Q^{-1} Q_{*}=Q_{*}$, if and only if $\Sigma=0$ (Theorem 4.1).
Theorem 4.1 is a generalization of results given in [20]-[22]. Following the procedure in [22] we obtain an alternative derivation by observing that $x_{*}(t)$ and

$$
\begin{equation*}
z(t)=x^{*}(t)-x_{*}(t) \tag{4.10}
\end{equation*}
$$

are orthogonal (to see this, note that $x_{*}(t)=$ $\left.\hat{E}\left\{x^{*}(t) \mid H_{t}^{-}(d y)\right\}\right)$ and applying Lemma 2.1. In fact, since $\hat{x}(t)=\hat{E}\left\{x(t) \mid H_{t}^{\square}\right\}\left(\right.$ Lemma 3.7) and $H_{t}^{\square}=H_{i}\left(x_{*}\right) \oplus H_{i}(z)$,

$$
\begin{equation*}
\hat{x}(t)=\hat{E}\left\{x(t) \mid x_{*}(t)\right\}+\hat{E}\{x(t) \mid z(t)\} \tag{4.11}
\end{equation*}
$$

Then using Lemmas 2.1 and 3.5 and the fact that

$$
\begin{equation*}
Q(t)=E\left\{z(t) z(t)^{\prime}\right\} \tag{4.12}
\end{equation*}
$$

we obtain

$$
\begin{equation*}
\hat{x}(t)=x_{*}(t)+Q_{*}(t) Q(t)^{-1} z(t), \tag{4.13}
\end{equation*}
$$

which is precisely (4.7).
If, for the moment, we restrict our attention to realizations in the interior of $\mathcal{S}$ we obtain the following wellknown result.

Corollary 4.1: Let $S \in$ int $\varsigma$, let $x$ be the state process of $S$, and let $\hat{x}$ be the corresponding smoothing estimate (4.1). Then, for each $t \in[0, T)$

$$
\begin{equation*}
\hat{x}(t)=\Sigma(t)\left[Q_{*}(t)^{-1} x_{*}(t)+Q^{*}(t)^{-1} x^{*}(t)\right] \tag{4.14}
\end{equation*}
$$

where $x_{*}$ and $x^{*}$ are given by (3.2) and (3.20), respectively, and the smoothing error covariance $\Sigma$ by

$$
\begin{equation*}
\Sigma(t)^{-1}=Q_{*}(t)^{-1}+Q^{*}(t)^{-1} \tag{4.15}
\end{equation*}
$$

Proof: Since $S \in \operatorname{int} \delta, Q_{*}$ and $Q^{*}$ are invertible. By writing (4.8) as $\Sigma=Q_{*} Q^{-1}\left(Q-Q_{*}\right)$ and using (4.6), it is seen that

$$
\begin{equation*}
\Sigma=Q_{*} Q^{-1} Q^{*} . \tag{4.16}
\end{equation*}
$$

Inverting this and again using (4.6) yields (4.15). From (4.16) we also see that $Q_{*} Q^{-1}=\Sigma\left(Q^{*}\right)^{-1}$. Then $I-$ $Q_{*} Q^{-1}=\Sigma\left[\Sigma^{-1}-\left(Q^{*}\right)^{-1}\right]=\Sigma Q_{*}^{-1}$. Hence, (4.14) follows from (4.7).
Relations (4.14) and (4.15) together with (3.2) and (3.20) is the Mayne-Fraser two-filter formula [5], [6], which has received considerable attention in the literature [7]-[9], [13]-[17]. Although this algorithm is easy to derive formally [9], its probabilistic justification has caused considerable difficulty, partly due to the fact that $Q^{*}(t) \rightarrow \infty$ as $t \rightarrow T$. The system (3.20) has usually been interpreted as a backward filter, and in [14]-[17] it is presented as the limit of such a filter as a certain covariance matrix function tends to infinity. However, in our stochastic realization setting (3.20) has a very natural interpretation: it is simply the maximum-variance forward realization $S^{*}$. By using the identity

$$
\begin{equation*}
x^{*}(t)=\bar{P}_{*}(t)^{-1} \bar{x}_{*}(t) \tag{4.17}
\end{equation*}
$$

we can instead write the smoothing formula (4.14) in terms of two Kalman-Bucy filters, one (3.2) evolving forward and the other (3.13) evolving backward in time. (Note that then (4.14) is defined on the whole interval $[0, T]$.) This fact was pointed out in [14], [15], [17], in which papers the backward estimate

$$
\begin{equation*}
\hat{x}_{b}(t)=\hat{E}\left\{x(t) \mid H_{t}^{+}(d y)\right\} \tag{4.18}
\end{equation*}
$$

was used in place of $\bar{x}_{*}$, a choice that may at first sight seem more natural. The reader should, however, note that

$$
\begin{equation*}
\hat{x}_{b}(t)=P(t) P^{*}(t)^{-1} x^{*}(t) \tag{4.19}
\end{equation*}
$$

is not invariant over $\mathcal{S}$ and is therefore less suitable for our purposes. It is not hard to see that

$$
\begin{equation*}
\left(Q^{*}\right)^{-1}=\left[\left(Q^{*}\right)^{-1}+P^{-1}\right] P\left(P^{*}\right)^{-1} \tag{4.20}
\end{equation*}
$$

and consequently (4.14) may also be written

$$
\begin{align*}
& \hat{x}(t)=\Sigma(t)\left\{Q_{*}(t)^{-1} x_{*}(t)\right. \\
&\left.+\left[Q^{*}(t)^{-1}+P(t)^{-1}\right] \hat{x}_{b}(t)\right\}, \tag{4.21}
\end{align*}
$$

which is the formula presented in [14], [15], [17]. The partitioned smoothing formula [12], [13] also can be seen to be equivalent to (4.14), and it can be used to derive all the equations of the Mayne-Fraser procedure. In the early papers [7], [8], relation (4.14) was introduced via a formula [47] for optimal weighting of two estimates with orthogonal errors. No justification of this orthogonality was given in [8], and the argument in [7] is incomplete due to problems with the endpoint condition. (A more satisfactory treatment has recently been presented in [48].) However, the stochastic realization theory provides a natural justification of this procedure. Indeed, (3.23) is the required orthogonality condition.

The smoothing formulas (4.7) and (4.14) are both based on the nonorthogonal decomposition (3.29), whereas (3.13) corresponds to the orthogonal decomposition

$$
\begin{equation*}
H_{t}^{\square}=H_{t}\left(x_{*}\right) \oplus H_{t}(z) \tag{4.22}
\end{equation*}
$$

(where, in either case, Lemma 3.7 justifies the restriction to the finite dimensional frame space $H_{t}^{D}$ ). We shall now take a closer look at representations of the latter type. It follows from (3.4) and (3.17) that $z$ as defined by (4.10) is the solution of

$$
\begin{equation*}
d z=\Gamma_{*} z d t-Q C^{\prime} R^{-1 / 2} d w^{*} ; \quad z(0)=x^{*}(0) \tag{4.23}
\end{equation*}
$$

where $\Gamma_{*}$ is the feedback matrix (3.25) of the Kalman-Bucy filter (3.2). To see this, note that the input process $w^{*}$ of the maximum variance realization $S^{*}$ is related to the innovation process $w_{*}$ through the relation

$$
\begin{equation*}
d w_{*}=R^{-1 / 2} C z d t+d w^{*} \tag{4.24}
\end{equation*}
$$

and that $B^{*}-B_{*}=-Q C^{\prime} R^{-1 / 2}$. We shall need the backward counterpart of (4.23). Observing that $Q$ is the covariance function of $z$, Lemma 2.3 yields the following equation for $\bar{z}=Q^{-1_{z}}$ :

$$
\begin{equation*}
d \bar{z}=-\Gamma_{*}^{\prime} \bar{z} d t-C^{\prime} R^{-1 / 2} d w_{*} ; \quad \bar{z}(T)=0 \tag{4.25}
\end{equation*}
$$

for, in view of (4.24), $w_{*}$ is the backward counterpart of $w^{*}$ with respect to (4.23). Note that $\bar{z}$ is defined on the whole interval $[0, T]$. The covariance matrix $\bar{Q}=Q^{-1}$ of $\bar{z}$ satisfies

$$
\begin{equation*}
\dot{\bar{Q}}=-\Gamma_{*}^{\prime} \bar{Q}-\bar{Q} \Gamma_{*}-C^{\prime} R^{-1} C ; \quad \bar{Q}(T)=0 \tag{4.26}
\end{equation*}
$$

The estimate $\hat{x}$ is then obtained from (4.13).

Theorem 4.2: Let $x$ be the state process of an arbitrary realization in $\mathcal{\delta}$. Then the smoothing estimate $\hat{x}(t)$ satisfies

$$
\begin{equation*}
\hat{x}(t)=x_{*}(t)+Q_{*}(t) \bar{z}(t) \tag{4.27}
\end{equation*}
$$

for all $t \in[0, T]$, where $x_{*}$ is given by (3.2) and $\bar{z}$ by (4.25) and (3.3). The process $\bar{z}$ is related to $x_{*}$ and $x^{*}$ through the relation

$$
\begin{equation*}
\bar{z}(t)=\bar{Q}(t)\left[x^{*}(t)-x_{*}(t)\right] \quad \text { for } t \in[0, T) \tag{4.28}
\end{equation*}
$$

Relation (4.27) is the smoothing formula of Bryson and Frazier [2]. (Also see [3], [4] and, in particular, [9].) What is new here is its interpretation (4.28) in terms of the minimum and maximum variance realizations $S_{*}$ and $S^{*}$. Theorem 4.2 can also be regarded as a generalization of a result presented in [21], and the basic techniques used there provide an alternative approach to deriving the above result.

Corollary 4.2: The smoothing estimate (4.27) satisfies the stochastic differential equation

$$
\begin{align*}
d \hat{x}=A \hat{x} d t+B\left(I-D^{\prime} R^{-1} D\right) & B^{\prime} \bar{z} d t \\
& +B D^{\prime} R^{-1}(d y-C \hat{x} d t) \tag{4.29}
\end{align*}
$$

with initial condition $\hat{x}(T)=x_{*}(T)$. If $S \in \mathcal{S}_{+}, \bar{z}$ can be replaced by $Q_{*}^{-1}\left(\hat{x}-x_{*}\right)$ in (4.29).

Proof: Inserting (3.2a), (3.2d), and (4.25) into

$$
d \hat{x}=d x_{*}+Q_{*} d \bar{z}+\dot{Q}_{*} \bar{z} d t
$$

and using (3.2b) yields (4.29). If $S \in \delta_{+}, Q_{*}^{-1}$ exists (Corollary 3.6.2), and (4.27) can be solved for $\bar{z}$.

We shall now study two different special cases of (4.29). First, let $B D^{\prime}=0$; this is a standard assumption in the smoothing literature. Then $\hat{x}$ is differentiable, and (4.29) reduces to

$$
\begin{equation*}
\frac{d \hat{x}}{d t}=A \hat{x}+B B^{\prime} \bar{z} ; \quad \hat{x}(T)=x_{*}(T) \tag{4.30}
\end{equation*}
$$

For realizations $S \in \mathscr{S}_{+}$(4.30) reduces to the smoothing formula of Rauch, Tung, and Striebel [3]

$$
\begin{equation*}
\frac{d \hat{x}}{d t}=A \hat{x}+B B^{\prime} Q^{-1}\left(\hat{x}-x_{*}\right) ; \quad \hat{x}(T)=x_{*}(T) . \tag{4.31}
\end{equation*}
$$

Secondly, assume that $D$ is square. Then $D$ is full rank and $D^{\prime} R^{-1} D=I$. Hence,

$$
\begin{equation*}
d \hat{x}=A \hat{x} d t+B D^{-1}(d y-C \hat{x} d t) ; \quad \hat{x}(T)=x_{*}(T) \tag{4.32}
\end{equation*}
$$

which defines a realization in $\delta_{0}$. Note that the original realization $S$ need not be internal; it may have an initial condition $x(0) \notin H(d y)$.

The problem of smoothing can be regarded as that of finding the "internal part" of the state process. Given a realization $S \in \mathcal{S}_{+}$, we shall next look at the structure of the "external part," i.e., the smoothing error $\tilde{x}:=x-\hat{x}$. To this end, first note that, given a realization (1.1), there exists an orthogonal $p \times p$-matrix $V(t)$ for each $t \in[0, T]$
such that

$$
\left[\begin{array}{c}
B(t)  \tag{4.33a}\\
D(t)
\end{array}\right]=\left[\begin{array}{cc}
B_{1}(t) & B_{2}(t) \\
R^{1 / 2}(t) & 0
\end{array}\right] V(t)
$$

where $B_{1}$ is $n \times m$ and $B_{2}$ is $n \times(p-m)$. Next let

$$
\left[\begin{array}{l}
d u  \tag{4.33b}\\
d v
\end{array}\right]=V d w
$$

define a pair of orthogonal increment processes $u$ and $v$, of dimensions $m$ and $p-m$, respectively. Obviously (4.33b) satisfies (1.2).

Theorem 4.3: Let $x$ be the state process of a realization $S \in \delta_{+}$and let $B_{2}$ and $v$ be defined by (4.33). Then the smoothing error $\tilde{x}$ is given by

$$
\left\{\begin{array}{l}
\hat{x}(t)=Q_{*}(t) \eta(t)  \tag{4.34a}\\
d \eta=-\Gamma_{*}^{\prime} \eta d t+Q_{*}^{-1} B_{2} d \zeta ; \quad \eta(T)=\eta_{T}
\end{array}\right.
$$

where $\eta_{T}=Q_{*}^{-1}(T)\left[x(T)-x_{*}(T)\right]$ and $\zeta$ is a $(p-m)$ dimensional orthogonal increment process of type (1.2) such that $H(d \zeta) \perp H(d y)$. Equation (4.34b) is a backward Markovian representation in the sense that $\eta_{T} \perp H(d \zeta)$ and the increments of $\zeta$ are given by

$$
\begin{equation*}
d \zeta=d v-B_{2}^{\prime} Q_{*}^{-1}\left(x-x_{*}\right) d t \tag{4.35}
\end{equation*}
$$

Proof: Define $z_{*}:=x-x_{*}$. Replacing $B d w$ and $D d w$ in (1.1) by $B_{1} d u+B_{2} d v$ and $R^{1 / 2} d u$, respectively, and noting that the innovation process $w_{*}$ in (3.4) is given by

$$
\begin{equation*}
d w_{*}=d u+R^{-1 / 2} C z_{*} d t \tag{4.36}
\end{equation*}
$$

and that $B_{1}-B_{*}=-Q_{*} C^{\prime} R^{-1 / 2}$ (Lemma 3.1), it is just a matter of simple calculations to see that $z_{*}$ satisfies

$$
d z_{*}=\Gamma_{*} z_{*} d t-Q_{*} C^{\prime} R^{-1 / 2} d u+B_{2} d v ; \quad z_{*}(0)=\xi
$$

Since $S \in \delta_{+}, Q_{*}(t)^{-1}$ exists for all $t \in[0, T]$ (Corollary 3.6.2). By Lemma 2.3 and (4.36), $\bar{z}_{*}=Q_{*}^{-1} z_{*}$ satisfies the backward Markovian representation

$$
\begin{array}{r}
d \bar{z}_{*}=-\Gamma_{*}^{\prime} \bar{z}_{*} d t-C^{\prime} R^{-1 / 2} d w_{*}+Q_{*}^{-1} B_{2} d \zeta \\
\bar{z}_{*}(T)=\eta_{T} \tag{4.37}
\end{array}
$$

where $\zeta$ is given by (4.35). Since $H(d \zeta) \perp H\left(d w_{*}\right)$ (by construction) and $H\left(d w_{*}\right)=H(d y), H(d \xi) \perp H(d y)$ as required. Now, in view of (4.27), $\tilde{x}=z_{*}-Q_{*} \bar{z}$, i.e., (4.34a) holds with $\eta:=\bar{z}_{*}-\bar{z}$. Then (4.34b) follows from (4.25) and (4.37).

As a corollary we see that the state process of any realization $S \in \mathcal{S}_{+}$can be decomposed into three orthogonal terms

$$
\begin{equation*}
x(t)=x_{*}(t)+Q_{*}(t) \bar{z}(t)+Q_{*}(t) \eta(t), \tag{4.38}
\end{equation*}
$$

each of which is the output of a stochastic system whose dynamical behavior is determined by the function $\Gamma_{*}$. This is seen from (4.25), (4.34) and the fact that (3.2a) can be written

$$
\begin{equation*}
d x_{*}=\Gamma_{*} x_{*} d t+B_{*} R^{-1 / 2} d y ; \quad x_{*}(0)=0 . \tag{4.39}
\end{equation*}
$$

(Note that both (4.25) and (4.34b) are backward representations. If we transform to the forward setting the systems matrices will be $\Gamma_{*}$ rather than $-\Gamma_{*}^{\prime}$.)

The internal realizations play an interesting role in the theory of smoothing. These are precisely the representations (1.1) for which the smoothing problem is trivial, i.e., $\tilde{x} \equiv 0$. The next theorem shows that (subject to a mild regularity condition) $S \in \mathcal{S}$ is internal if and only if $B_{2} \equiv 0$ and the components of the initial conditions of (1.1) belong to $H(d y)$. In view of Theorem 4.3, this is to be expected.

Theorem 4.4: A realization $S \in \mathcal{S}$ such that $\binom{B}{D}$ has full rank is internal if and only if $S \in \delta_{0}$.

Proof: We only need to prove the "only if" part; the "if" part was proved in Section III. Let $S$ be internal. Since the condition $x(0) \in H(d y)$ holds trivially, it only remains to show that $B_{2}$, as defined by (4.33a), is identically zero. In view of the fact that $x=\hat{x}$, comparing (1.1) and (4.29) shows that the identity

$$
\begin{equation*}
B\left(I-D^{\prime} R^{-1} D\right) B^{\prime} z d t+B D^{\prime} R^{-1} D d w=B d w \tag{4.40}
\end{equation*}
$$

must hold. It is not hard to see that $B D^{\prime} R^{-1} D=\left(B_{1}, 0\right) V$ and $B\left(I-D^{\prime} R^{-1} D\right) B^{\prime}=B_{2} B_{2}^{\prime}$, and therefore (4.40) can be written

$$
B_{2} B_{2}^{\prime} \bar{z} d t=B_{2} d v
$$

which cannot hold unless $B_{2}=0$. Then the full rank condition implies that $p=m$.

## References

[1] J. L. Doob, Stochastic Processes. New York: Wiley, 1953.
[2] A. E. Bryson and M. Frazier, "Smoothing for linear and nonlinear dynamic systems," Aero. Syst. Div., Wright-Patterson Air Force Base, OH, TDR 63-119, pp. 353-364, 1963.
[3] H. E. Rauch, F. Tung, and C. T. Striebel, "Maximum likelihood estimates of linear dynamic systems," AIAA J., vol. 3, pp. 1445-1450, Aug. 1965.
[4] J. S. Meditch, "On optimal fixed point smoothing," Int. J. Contr., vol. 6, no. 2, pp. 189-199, 1967.
[5] D. Q. Mayne, "A solution of the smoothing problem for linear dynamic systems," Automatica, vol. 4, pp. 73-92, 1966.
[6] D. C. Fraser, "A new technique for optimal smoothing of data," Sc.D. dissertation, MIT, Cambridge, 1967.
[7] R. K. Mehra, "On optimal and suboptimal linear smoothing," in Proc. Nat. Elec. Conf. XXIV, 1968, pp. 119-124.
[8] D. C. Fraser and J. E. Potter, "The optimum linear smoother as a combination of two optimum linear filters," IEEE Trans. Automat. Contr., vol. AC-14, pp. 387-390, 1969.
[9] T. Kailath and P. A. Frost, "An innovations approach to leastsquares estimation-Part II: Linear smoothing in additive noise," IEEE Trans. Automat. Contr., vol. AC-13, pp. 655-660, 1968.
[10] L. E. Zachrisson, "On optimal smoothing of continuous-time Kalman processes," "Inform. Sci., vol. 1, pp. 143-172, 1969.
[11] A. Lindquist, "A theorem on duality between estimation and control for linear stochastic systems with time delay," J. Math. Anal. Appl., vol. 37, no. 2, pp. 516-536, 1972.
[12] D. G. Lainiotis, "Optimal linear smoothing: Continuous data case," Int. J. Contr., vol. 14, no. 6, pp. 1137-1148, 1971.
[13] $\frac{\text { Case, "Partitioned estimation algorithms, II: Linear estimation," }}{}$ Inform, Sci, vol. 7, pp. 317-340, 1974.
[14] L. Ljung and T. Kailath, "A unified approach to smoothing formulas," Automatica, vol. 12, pp. 147-157, 1976.
[15] T. Kailath and L. Ljung, "Backwards Markovian models for second order stochastic process," IEEE Trans. Inform. Theory, vol. IT-22, pp. 488-491, 1976.
[16] G. S. Sidhu and U. B. Desai, "New smoothing algorithms based on reversed-time lumped models," IEEE Trans. Automat. Contr., vol. AC-21, pp. 538-541, 1976.
[17] -,"Reversed-time lumped models, smoothing algorithms and Fredholm resolvents," in Proc. JHU Conf: on Information Science and Systems, Johns Hopkins Univ., Baltimore, MD, Mar.-Apr., 1976.
[18] J. S. Meditch, "A survey of data smoothing for linear and non-linear dynamic systems," Automatica, vol. 9, pp. 151-162, 1973.
[19] G. Picci, "Stochastic realization of Gaussian processes," Proc. IEEE, vol. 64, pp. 112-122, 1976.
[20] A. Lindquist and G. Picci, "On the stochastic realization problem," SIAM J. Contr. Optimiz., vol. 17, pp. 365-389, May 1979.
[21] M. Pavon, "Stochastic realization and invariant directions of the matrix Riccati equation," SIAM J. Contr. Optimiz., to be published.
[22] G. Ruckebusch, "Représentations Markoviennes de processus Gaussiens stationnaires," Ph.D. dissertation, Univ. of Paris, VI, 1975.
[23] A. Lindquist and G. Picci, "On the structure of minimal splitting subspaces in stochastic realization theory," in Proc. 1977 Conf. on Decision and Control, New Orleans, LA, 1977, pp. 42-48.
[24] ,"A state-space theory for stationary stochastic processes," in Proc. 21st Midwest Symp. on Circuits and Systems, Ames, IA, Aug. 1978.
[25] $\xrightarrow{\text { 1 }}$, "A Hardy space approach to the stochastic realization problem," in Proc. 1978 Conf. on Decision and Control, San Diego, CA, Jan. 1979.
[26] $\longrightarrow$, "Realization theory for multivariate stationary Gaussian process I: State space construction," presented at the 4th Int. Symp. on the Mathematical Theory of Networks and Systems, July 1979, Delft, Holland.
[27] A. Lindquist, G. Picci, and G. Ruckebusch, "On minimal splitting subspaces and Markovian representations," Math. Syst. Theory, vol. 12, pp. 271-279, 1979.
[28] M. Pavon, "Constant directions and singular stochastic realization: The scalar case," in Proc. 1st Colloque AFCET-SMF, Sept. 1978, Palaiseau, France.
[29] Y. A. Rozanov,"On two selected topics connected with stochastic systems theory," Appl. Math. Optimiz., vol. 3, pp. 73-80, 1976.
[30] G. Ruckebusch, "Représentations Markoviennes de processus Gaussiens stationnaires," C. R. Acad. Sci. Paris, ser. A, vol. 22, pp. 649-651, 1976.
[31] -, "Représentations Markoviennes de processus Gaussien stationnaires et applications statistiques," in Journées de Statistiques des Processus Stochastiques (Lecture Notes in Mathematics). Grenoble: Springer-Verlag, 1977, p. 636.
[32] "A state space approach to the stochastic realization problem," in Proc. 1978 Intern. Symp. on Circuits and Systems, New York.
[33] "A geometric approach to the stochastic realization problem," Cent. Math. Appl., Ecole Polytechnique, Internal Rep. 41, Nov. 1978.
[34] B. D. O. Anderson, "The inverse problem of stationary covariance generation," J. Statist. Phys., vol. 1, pp. 133-147, 1969.
[35] P. Faurre, "Réalisations Markoviennes de processus stationnaires," IRIA (LABORIA), Le Chesnay, France, Res. Rep. 13, Mar. 1973.
[36] J. C. Willems, "Dissipative dynamical systems, Part II: Linear systems with quadratic supply rates," Arch. Rational Mechan. Anal., vol. 45, pp. 352-393, 1972.
[37] M. Clerget, "Systèmes lineaires positifs non stationnaires," IRIA (LABORIA), Le Chesnay, France, Res. Rep. 65, Apr. 1974.
[38] F. Germain, "Algoritmes de calcul de realisations stationnaires," IRIA (LABORIA), Le Chesnay, France, Res. Rep. 66, Apr. 1974.
[39] R. W. Brockett, Finite Dimensional Linear Systems. New York: Wiley, 1970.
[40] L. M. Silverman and H. E. Meadows, "Controllability and observability in time-variable linear systems," SIAM J. Contr., vol. 5, no. 1, pp. 64-73, 1967.
[41] D. M. Wiberg State Space and Linear Systems (Shaum's Outline Series). New York: McGraw-Hill, 1971.
[42] D. G. Lainiotis, "General backward Markov models," IEEE Trans. Automat. Contr., vol. AC-21, pp. 595-598, Aug. 1976.
[43] T. Kailath, "A note on least squares estimation by the innovations method," SIAM J. Contr., vol. 10, no. 3, pp. 477-486, 1972.
[44] B. D. Anderson and P. J. Moylan, "Spectral factorization of a finite-dimensional non-stationary matrix covariance," IEEE Trans. Automat. Contr., vol. AC-19, pp. 680-692, Dec. 1974.
[45] A. Lindquist, "Optimal filtering of continuous-time stationary processes by means of the backward innovation process," SIAM J. Contr., vol. 12, pp. 747-754, 1974.
[46] J. C. Willems, "Least squares stationary optimal control and the algebraic Riccati equation," IEEE Trans. Automat. Contr., vol. AC-16, pp. 621-634, 1971.
[47] R. Deutsch, Estimation Theory. Englewood Cliffs, NJ: PrenticeHall, 1965.
[48] J. Wall, Jr., "Control and estimation for large-scale systems having spatial symmetry," Ph.D. dissertation, MIT, Cambridge, 1978.


Faris A. Badawi was born in Attil, West Bank of Jordan, on December 18, 1946. He received the B.S. and the M.S. degrees in mathematics from the American University of Beirut, Beirut, Lebanon, in 1971 and 1975, respectively.

From 1971 to 1975 he was with the Training Department at the Arabian American Oil Company (ARAMCO), Dhahram, Saudi Arabia as an Advanced Bilingual Teacher. Since 1975, he has been pursuing his studies towards the Ph.D. degree in mathematics at the University of Kentucky, Lexington. His research interests lie primarily in the fields of stochastic systems theory, estimation, and control.


Anders Lindquist (M'77) was born in Lund, Sweden, on November 21, 1942. He received the Ph.D. degree from the Royal Institute of Technology, Stockholm, Sweden.

He was employed in various teaching and research positions between 1969 and 1972 at the Royal Institute of Technology. In 1972 he was appointed a Docent of Optimization and Systems Theory there. During the academic year of 1972-1973 he was a Visiting Assistant Professor at the University of Florida, during the Fall
semester of 1973 he served as an Associate Professor (Research) in the Division of Applied Mathematics at Brown University, and during the Spring of 1974 he was a Visiting Associate Professor at the University of Kentucky, where in August 1974 he was appointed as Associate Professor of Mathematics. He has also held visiting positions at the University of Lund, Sweden, State University of New York at Albany, and the University of Padua, Italy. During 1975-1976, he was a SIAM Visiting Lecturer.

Dr. Lindquist is a member of the Society of Industrial and Applied Mathematics (SIAM) and the IEEE Control Systems Society.


Michele Pavon was born in Venice, Italy, on October 12, 1950. He received the "laurea" degree in mathematics from the University of Padua, Padua, Italy, in 1974, and the Ph.D. degree in mathematics from the University of Kentucky, Lexington, in 1979.
During the 1974-1975 academic year, he held a C.N.R. Fellowship at the Institute of Algebra and Geometry of the University of Padua. His research interests lie in the area of systems theory and stochastic programming. He is presently with the Department of Mathematics, University of Kentucky.

# Short Papers 

## Multiobjective Trajectory Optimization for Electric Trains

SAROSH N. TALUKDAR, MEMBER, IEEE, AND ROBERT L. KOO


#### Abstract

This paper develops a method for generating Pareto efficient trajectories that provide optimal tradeoffs between two conflicting attribu-tes-the total energy consumed and the total time taken to complete each trajectory. Straightforward formulations of the multiohjective optimization problem in these attributes are difficult to solve because of certain nonlinearities in train models and certain constraints on train trajectories. A discrete reformulation is developed to circumvent these difficulties and produce computationally feasible algorithms. The results from the algorithms can be used to develop operating strategies for existing systems and to compare hardware alternatives in planning new systems. An illustration is included.


## Introduction

In competing with other travel modes, the operators of intercity and rapid transit electric trains find themselves under pressure to conserve both the total energy, $W$, and the total travel time, $T$, required for each journey. These are conflicting objectives. Reductions in energy must usually be paid for with increases in travel time and vice versa. Such situations demand compromises and tradeoffs. The best tradeoffs occur

[^3]along the Pareto efficient (noninferior) surface of the feasible region in attribute space [1], [2]. The determination of this surface constitutes a multiobjective optimization problem.

Prior work in the area (e.g., [3]-[7]) has either tended to neglect objective conflicts and important systems nonlinearities or has produced site specific solutions. In the succeeding material we will develop a general multiobjective methodology for generating trajectories that are Pareto efficient in the attributes, $W$ and $T$.

## Problem Description

A representative trajectory of a train on an intercity journey is shown in Fig. 1. The frequent speed changes between stations are required to accommodate speed limits which vary with the quality of the track and its curvature. In general, the constraints imposed by speed limits may be stated in the form

$$
\begin{equation*}
0 \leqslant c \leqslant v_{\max }(x) \tag{1}
\end{equation*}
$$

where
c is the train's velocity
$x \quad$ is its distance from the beginning of its journey
$v_{\text {max }}(x)$ is the speed limit at the point $x$.
Propulsion systems for electric trains contain motor drives to develop and deliver tractive force to the track and conditioners to control the flow of power to the motors and thereby, their tractive output. A variety of conditioner-motor combinations are feasible [8]. Some allow power flows only in one direction; others permit two way flows (during braking the motor acts as a generator and power is pumped back into the network-a phenomenon called regeneration).


[^0]:    Manuscript received April 2, 1979; revised August 17, 1979. Paper recommended by A. Ephremides, Chairman of the Estimation Committee. This work was supported by the Air Force Office of Scientific Research, Air Force Systems Command, USAF, under Grant AFOSR-78-3519.

    The authors are with the Department of Mathematics, University of Kentucky, Lexington, KY 40506.

[^1]:    ${ }^{1}$ Some of these shortcomings have been pointed out in a recent thesis by Wall [48], brought to our attention after the submission of this paper.
    ${ }^{2}$ For example, the Moore-Penrose pseudo-inverse can be used.

[^2]:    ${ }^{3}$ It is not hard to see that the concept of minimality used here is equivalent to assuming both that i) the imput-output map of (1.1a) is minimal and that ii) the family of state spaces $\left\{H_{t}(x) ; t \in[0, T]\right\}$ is minimal in the sense of the geometric state space theory outlined in [27].

[^3]:    Manuscript received December 20, 1978; revised June 19, 1979. Paper recommended by A. Manitius, Chairman of the Optimal Systems Committee. This work was supported in part by the Department of Transportation under Contract DOT-OS-60219.
    The authors are with the Department of Electrical Engineering, Carnegie Institute of Technology, Carnegie-Mellon University, Pitsburgh, PA 15213.

