# Differentiability Properties of the Efficient $\left(\mu, \sigma^{2}\right)$-Set in the Markowitz Portfolio Selection Method 

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## DIFFERENTIABILITY PROPERTIES OF THE EFFICIENT $\left(\mu, \sigma^{2}\right)$ SET IN THE MARKOWITZ PORTFOLIO SELECTION METHOD

## 1 Introduction

The standard portfolio selection problem with linear constraints may be formulated as follows. An investor wants to invest an amount of one unit in the securities $1, \ldots, n$. If he invests an amount $x_{j}$ in security $j(j=1, \ldots, n)$ the $x_{j}$ should satisfy the conditions

$$
\begin{equation*}
\mathcal{A} X=B, \tag{1.1}
\end{equation*}
$$

$$
\begin{equation*}
X \geq \mathcal{O} \tag{1.2}
\end{equation*}
$$

with $\mathcal{A}$ an $(m \times n)$ - matrix with full rank, $B$ an $m$-vector and $X^{\prime}=\left(x_{1}, \ldots, x_{n}\right) ;(1.1)$ includes the condition

$$
\begin{equation*}
\sum_{j=1}^{n} x_{j}=1 . \tag{1.3}
\end{equation*}
$$

The yearly return on one dollar invested in security $j$ equals $\underline{r}_{j}$ with $\mu_{j}=\mathcal{E}_{j}$; the covariance matrix of the random variables $\underline{r}_{j}$ is $\mathcal{C}$. The yearly return $\underline{r}(X)$ on a portfolio $X$ equals

$$
\begin{equation*}
\underline{r}(X)=\sum_{j=1}^{n} x_{j} \underline{r}_{j} \tag{1.4}
\end{equation*}
$$

with $M^{\prime}=\left(\mu_{1}, \ldots, \mu_{n}\right)$, the expected yearly return $\mathcal{E} \underline{r}(X)$ equals $M^{\prime} X$ and will be denoted by $\mu(X)$, so

$$
\begin{equation*}
\mu(X)=M^{\prime} X ; \tag{1.5}
\end{equation*}
$$

the variance $\sigma^{2}(\underline{r}(X))$ equals $X^{\prime} \mathcal{C} X$ and will be denoted by $\sigma^{2}(X)$, so

$$
\begin{equation*}
\sigma^{2}(X)=X^{\prime} \mathcal{C} X \tag{1.6}
\end{equation*}
$$

For equivalent formulations of the conditions (1.1), (1.2) cf. H.M. Markowitz (1987) p. 24-27, for nonlinear constraints J. Kriens and J.Th. van Lieshout (1988).

A feasible portfolio $\bar{X}$ is called efficient if it is a solution of both

$$
\begin{equation*}
\min _{X}\left\{\sigma^{2}(X) \mid \mu(X) \geq \mu(\bar{X}) \wedge \mathcal{A} \bar{X}=B \wedge \bar{X} \geq \mathcal{O}\right\} \tag{1.7}
\end{equation*}
$$

and

$$
\begin{equation*}
\max _{X}\left\{\mu(X) \mid \sigma^{2}(X) \leq \sigma^{2}(\bar{X}) \wedge \mathcal{A} \bar{X}=B \wedge \bar{X} \geq \mathcal{O}\right\} \tag{1.8}
\end{equation*}
$$

All efficient portfolios can be derived by computing

$$
\begin{equation*}
\min _{X}\left\{X^{\prime} \mathcal{C} X-\lambda M^{\prime} X \mid \mathcal{A} \bar{X}=B \wedge \bar{X} \geq \mathcal{O}\right\} \tag{1.9}
\end{equation*}
$$

for all $\lambda \geq 0$; cf. H.M. Markowitz (1959) p. 315-316, or for a precise and more general statement of the theorem underlying the algorithm J. Kriens and J.Th. van Lieshout (1988).

With $U^{\prime}=\left(u_{1}, \ldots, u_{m}\right)$ and $V^{\prime}=\left(v_{1}, \ldots, v_{n}\right)$ as Lagrange multipliers of (1.1) and (1.2), respectively, the Kuhn-Tucker conditions of (1.9) run

$$
\begin{gather*}
-2 \mathcal{C} X-\mathcal{A}^{\prime} U+V=-\lambda M  \tag{1.10}\\
\mathcal{A} X \quad=B \tag{1.11}
\end{gather*}
$$

$$
\begin{equation*}
V^{\prime} X=0, X \geq \mathcal{O}, V \geq \mathcal{O}, U \text { free. } \tag{1.12}
\end{equation*}
$$

Loosely speaking we can describe an algorithm to solve this system for all $\lambda \geq 0$ as follows. Start with choosing $\lambda=0$, thus with determining the minimum possible variance,
and next raise $\lambda$ to get (new) efficient portfolios. For specific values of $\lambda$ there is a change in the basis. Let these values be $\bar{\lambda}_{1}, \ldots, \bar{\lambda}_{k}$, the corresponding efficient solutions be $\bar{X}_{1}, \ldots, \bar{X}_{k}$ with mean-variance combinations $\left(\mu\left(\bar{X}_{1}\right), \sigma^{2}\left(\bar{X}_{1}\right)\right), \ldots,\left(\mu\left(\bar{X}_{k}\right), \sigma^{2}\left(\bar{X}_{k}\right)\right)$. The sequence $\bar{X}_{1}, \ldots, \bar{X}_{k}$ is called the set of corner portfolios, the set of all $\left(\mu(\bar{X}), \sigma^{2}(\bar{X})\right)$ points in the ( $\mu, \sigma^{2}$ )-plane corresponding to efficient portfolios $\bar{X}$ is the set of efficient ( $\bar{\mu}, \bar{\sigma}^{2}$ ) combinations of the problem, or the efficient frontier.

This last set satisfies the following properties:
a. between the $\left(\bar{\mu}, \bar{\sigma}^{2}\right)$ points of two adjacent corner portfolios $\bar{X}_{i}$ and $\bar{X}_{i+1}\left(\neq \bar{X}_{i}\right)$ it is part of a strictly convex parabola;
b. on the interior of the segments mentioned in a, the relation

$$
\begin{equation*}
\left(\frac{d \sigma^{2}}{d \mu}\right)_{\left(\bar{\mu}, \bar{\sigma}^{2}\right)}=\bar{\lambda} \tag{1.13}
\end{equation*}
$$

holds; it is strictly increasing as a function of $\mu$;
c. in the ( $\bar{\mu}, \bar{\sigma}^{2}$ ) points corresponding to corner portfolios, the left hand derivative $\left(\frac{d \sigma^{2}}{d \mu}\right)_{L}$ and the right hand derivative $\left(\frac{d \sigma^{2}}{d \mu}\right)_{R}$ exist and satisfy

$$
\begin{equation*}
\left(\frac{d \sigma^{2}}{d \mu}\right)_{L} \leq\left(\frac{d \sigma^{2}}{d \mu}\right)_{R} \tag{1.14}
\end{equation*}
$$

From b it follows that on those segments there is a one to one correspondence between the values of $\bar{\lambda}$ and $\bar{\mu}$. In corner portfolios this is only true if $\left(\frac{d \sigma^{2}}{d \mu}\right)_{L}=\left(\frac{d \sigma^{2}}{d \mu}\right)_{R}$, which implies differentiability of the ( $\mu, \sigma^{2}$ ) curve. For proofs cf. H.M. Markowitz (1987), p. 176 and J. Kriens and J.Th. van Lieshout (1988).

Section 2 of this paper contains a more precise discussion of the algorithm to solve (1.10),..,(1.12) for every $\lambda \geq 0$, section 4 necessary and sufficient conditions for the equality sign in (1.14). In preparation for the second topic we present a slightly adapted form of the explicit formulae for $\bar{X}, \mu(\bar{X})$ and $\sigma^{2}(\bar{X})$ as derived by J. Kriens and J.TH. van Lieshout (1988) in section 3.
Section 5 compares with other literature and section 6 considers the standard portfolio case supplied with one riskless asset. Throughout the whole paper we assume $\mathcal{C}$ positive definite.

## 2 The algorithm

In order to present a more precise discussion of the algorithm we first prove the following lemma.

## Lemma 2.1

If in a portfolio selection problem

1) $\forall_{j} \sigma^{2}\left(\underline{r}_{j}\right)>0$
2) there are no linear relations between the returns $\underline{r}_{j}$,
portfolios $\bar{X}_{1}$ and $\bar{X}_{2}\left(\neq \bar{X}_{1}\right)$ with $\mu\left(\bar{X}_{1}\right)=\mu\left(\bar{X}_{2}\right)$ and $\sigma^{2}\left(\bar{X}_{1}\right)=\sigma^{2}\left(\bar{X}_{2}\right)$ cannot be efficient.

## Proof

Let

$$
\bar{X}=\alpha \bar{X}_{1}+(1-\alpha) \bar{X}_{2} \quad(0<\alpha<1)
$$

then

$$
\begin{aligned}
& \underline{r}(\bar{X})=\alpha \underline{r}\left(\bar{X}_{1}\right)+(1-\alpha) \underline{r}\left(\bar{X}_{2}\right) \\
& \mu(\bar{X})=\mu\left(\bar{X}_{1}\right)=\mu\left(\bar{X}_{2}\right) \\
& \sigma^{2}(\underline{r}(\bar{X}))=\alpha^{2} \sigma^{2}\left(\underline{r}\left(\bar{X}_{1}\right)\right)+2 \alpha(1-\alpha) \rho \sigma\left(\underline{r}\left(\bar{X}_{1}\right)\right) \sigma\left(\underline{r}\left(\bar{X}_{2}\right)\right)+ \\
& (1-\alpha)^{2} \sigma^{2}\left(\underline{r}\left(\bar{X}_{2}\right)\right)=\sigma^{2}\left(\bar{X}_{1}\right)\left[\alpha^{2}+2 \alpha(1-\alpha) \rho+(1-\alpha)^{2}\right]=\sigma^{2}\left(\bar{X}_{1}\right) f(\alpha) .
\end{aligned}
$$

For $\rho \neq 1, f(\alpha)<1$ for $0<\alpha<1$, so $\sigma^{2}(\bar{X})<\sigma^{2}\left(\bar{X}_{1}\right)$ and $\bar{X}_{1}$ and $\bar{X}_{2}$ are not efficient. $\rho\left(\underline{r}\left(\bar{X}_{1}\right), \underline{r}\left(\bar{X}_{2}\right)\right)=1$ iff all realizations of $\left(\underline{r}\left(\bar{X}_{1}\right), \underline{r}\left(\bar{X}_{2}\right)\right)$ are situated on a straight line, so all points $\left(\sum_{j=1}^{n} x_{j 1} r_{j}, \sum_{j=1}^{n} x_{j 2} r_{j}\right)$ are on a straight line. This means

$$
\exists_{a} \exists_{d} \forall_{R}\left(\sum_{j=1}^{n} x_{j 2} r_{j}\right)=a+d\left(\sum_{j=1}^{n} x_{j 1} r_{j}\right) .
$$

Let

$$
\forall_{j} a_{j}=d x_{j 1}-x_{j 2},
$$

then
$\forall_{R} a+\sum_{j=1}^{n} a_{j} r_{j}=0$.

We discern four cases:
a) $\forall_{j} a_{j}=0 \Rightarrow \forall_{j} d x_{j 1}=x_{j 2}$, leading with (1.3) to $d=1$ and $\bar{X}_{1}=\bar{X}_{2}$, which contradicts $\bar{X}_{1} \neq \bar{X}_{2}$;
b) $a_{i} \neq 0, \forall_{j \neq i} \quad a_{j}=0 \Rightarrow \forall_{R} \quad a+a_{i} r_{i}=0$ and $r_{i}$ is fixed, so $\sigma^{2}\left(\underline{r}_{i}\right)=0$, which contradicts condition 1);
c) $a_{i} \neq 0, a_{k} \neq 0, \forall_{j \neq i, k} a_{j}=0 \Rightarrow \forall_{R} a+a_{i} r_{i}+a_{k} r_{k}=0$, which contradicts condition 2);
d) More than two $a_{i} \neq 0$; conclusion as under c).

So $\rho\left(\underline{r}\left(\bar{X}_{1}\right), \underline{r}\left(\bar{X}_{2}\right)\right) \neq 1$ and the lemma is proved.

Remark 2.1. From the proof it follows that conditions 1) and 2) are also necessary. Moreover the conditions 1) and 2) hold iff $\mathcal{C}$ is positive definite.

From lemma 2.1 it is clear that for $\mathcal{C}$ positive definite the corner portfolios $\bar{X}_{1}, \ldots, \bar{X}_{k}$ are uniquely determined. However, there are not always as many different corner portfolios as there are different bases during the computations; different bases may yield the same portfolio and also different values $\bar{\lambda}_{i}$ may yield the same portfolio. In this respect the notation in section 1 is misleading.

Starting the algorithm with $\lambda=0$ and next raising $\lambda$, the algorithm produces a series of bases. Bases which hold for just one value of $\lambda$ are dropped so that only bases corresponding to nondegenerate $\lambda$-intervals are left.
Denote for a given basis of the system (1.10) ,..., (1.12), the set of basic $x$-variables by $\left(X_{b}\right)_{i}$. In section 3 we will show that the values $\left(\bar{X}_{b}\right)_{i}$ of the basic $x$-variables satisfy

$$
\begin{equation*}
\left(\bar{X}_{b}\right)_{i}=A_{i}+D_{i} \bar{\lambda} \tag{2.1}
\end{equation*}
$$

for all $\bar{\lambda}$ in the corresponding interval; the constants $A_{i}$ and $D_{i}$ will be computed explicitly. So if $\mathcal{C}$ is positive definite the whole efficient frontier is uniquely determined.

With corner portfolios there correspond at least two vectors $D_{i}$, the vector corresponding to the "old" basis and the vector corresponding to the "new" basis. But there may be more associated vectors $D_{i}$, either because there exists an equivalent basis for the "old" or for the "new" basis producing the same corner portfolio $\left(\bar{X}_{b}\right)_{i}$, or because the series of vectors $D_{i}$ contains one or more vectors $D=\mathcal{O}$. In the latter case the same vector $\left(\bar{X}_{b}\right)_{i}$ is produced for different values of $\lambda$. If the "new" basis is uniquely determined, then for efficient portfolios which are not corner portfolios the vector $D_{i}$ is uniquely determined.

## 3 Explicit expressions for efficient portfolios

Starting from the Kuhn-Tucker conditions for the solution of (1.9), J. Kriens and J.TH. van Lieshout (1988) derive an expression for the basic variables which, if $\mathcal{C}$ is positive definite, holds for every efficient portfolio. We present their results in a slightly adapted form.

For a fixed value $\bar{\lambda}$ of $\lambda(1.10), \ldots,(1.12)$ run

$$
\begin{equation*}
-2 \mathcal{C} X-\mathcal{A}^{\prime} U+V=-\bar{\lambda} M \tag{3.1}
\end{equation*}
$$

$$
\begin{equation*}
\mathcal{A} X \quad=B \tag{3.2}
\end{equation*}
$$

$$
\begin{equation*}
V^{\prime} X=0, X \geq \mathcal{O}, V \geq \mathcal{O}, U \text { free. } \tag{3.3}
\end{equation*}
$$

The equations (3.1) and (3.2) can be summarized as

$$
\begin{array}{|ccc|c|}
\hline X^{\prime} & U^{\prime} & V^{\prime} &  \tag{3.4}\\
\hline-2 \mathcal{C} & -\mathcal{A}^{\prime} & \mathcal{J} & -\bar{\lambda} M \\
\mathcal{A} & \mathcal{O} & \mathcal{O} & B \\
\hline
\end{array}
$$

If

$$
\begin{equation*}
Z_{b}^{\prime}=\left(X_{b}^{\prime}, U^{\prime}, V_{b}^{\prime}\right) \tag{3.5}
\end{equation*}
$$

denotes a set of basic variables for a given efficient portfolio (3.4) can be partitioned into

| $X_{b}^{\prime}$ | $X_{n b}^{\prime}$ | $U^{\prime}$ | $V_{b}^{\prime}$ | $V_{n b}^{\prime}$ |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| $-2 \mathcal{C}_{b_{1}}$ | $-2 \mathcal{C}_{n b_{1}}$ | $-\mathcal{A}_{b}^{\prime}$ | $\mathcal{O}$ | $\mathcal{J}$ | $-\bar{\lambda} M_{b}$ |
| $-2 \mathcal{C}_{b_{2}}$ | $-2 \mathcal{C}_{n b_{2}}$ | $-\mathcal{A}_{n b}^{\prime}$ | $\mathcal{J}$ | $\mathcal{O}$ | $-\bar{\lambda} M_{n b}$ |
| $\mathcal{A}_{b}$ | $\mathcal{A}_{n b}$ | $\mathcal{O}$ | $\mathcal{O}$ | $\mathcal{O}$ | $B$ |

The matrix $-2 \mathcal{C}$ is partitioned into the square matrices $-2 \mathcal{C}_{b_{1}}$ and $-2 \mathcal{C}_{n b_{2}}$ corresponding to basic and non-basic $x$-variables and into $-2 \mathcal{C}_{b_{2}}$ and $-2 \mathcal{C}_{n b_{1}}$ with $\mathcal{C}_{b_{2}}=\mathcal{C}_{n b_{1}}^{\prime} \cdot \mathcal{A}_{b}, M_{b}$ and $\mathcal{A}_{n b}, M_{n b}$ also correspond to basic and non-basic variables respectively. The matrix of coefficients of basic variables is

$$
\mathcal{B}=\left(\begin{array}{ccc}
-2 \mathcal{C}_{b_{1}} & -\mathcal{A}_{b}^{\prime} & \mathcal{O}  \tag{3.7}\\
-2 \mathcal{C}_{b_{2}} & -\mathcal{A}_{n b}^{\prime} & \mathcal{J} \\
\mathcal{A}_{b} & \mathcal{O} & \mathcal{O}
\end{array}\right)
$$

To facilitate computations Kriens and van Lieshout reshuffle (3.7) into

$$
\mathcal{B}_{v}=\left(\begin{array}{ccc}
-2 \mathcal{C}_{b_{1}} & -\mathcal{A}_{b}^{\prime} & \mathcal{O}  \tag{3.8}\\
\mathcal{A}_{b} & \mathcal{O} & \mathcal{O} \\
-2 \mathcal{C}_{b_{2}} & -\mathcal{A}_{n b}^{\prime} & \mathcal{J}
\end{array}\right)
$$

The values of the basic variables are

$$
\bar{Z}_{b}=\left(\begin{array}{l}
\bar{X}_{b}  \tag{3.9}\\
\bar{U} \\
\bar{V}_{b}
\end{array}\right)=\mathcal{B}_{v}^{-1}\left(\begin{array}{l}
\mathcal{O} \\
B \\
\mathcal{O}
\end{array}\right)-\bar{\lambda} \mathcal{B}_{v}^{-1}\left(\begin{array}{l}
M_{b} \\
\mathcal{O} \\
M_{n b}
\end{array}\right)
$$

We find explicit expressions for these values by computing $\mathcal{B}_{v}^{-1}$ :
with

$$
\begin{align*}
& \left(\begin{array}{cc}
-2 \mathcal{C}_{b_{1}} & -\mathcal{A}_{b}^{\prime} \\
\mathcal{A}_{b} & \mathcal{O}
\end{array}\right)^{-1}=  \tag{3.11}\\
& \left(\begin{array}{cc|c}
-\frac{1}{2} \mathcal{C}_{b_{1}}^{-1}+ & \frac{1}{2} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} \mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} & \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} \\
& -\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} \mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} & -2\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1}
\end{array}\right)
\end{align*}
$$

Substituting (3.11) into (3.10) and the result into (3.9), we find

$$
\begin{equation*}
\bar{X}_{b}=A+D \bar{\lambda} \tag{3.12}
\end{equation*}
$$

with

$$
\begin{equation*}
A=\mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} B \tag{3.13}
\end{equation*}
$$

and

$$
\begin{equation*}
D=\frac{1}{2}\left[\mathcal{C}_{b_{1}}^{-1}-\mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} \mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1}\right] M_{b} \tag{3.14}
\end{equation*}
$$

The corresponding values $\mu\left(\bar{X}_{b}\right)$ and $\sigma^{2}\left(\bar{X}_{b}\right)$ are

$$
\begin{equation*}
\mu\left(\bar{X}_{b}\right)=M_{b}^{\prime} A+M_{b}^{\prime} D \bar{\lambda} \tag{3.15}
\end{equation*}
$$

$$
\begin{equation*}
\sigma^{2}\left(\bar{X}_{b}\right)=A^{\prime} \mathcal{C}_{b_{1}} A+D^{\prime} \mathcal{C}_{b_{1}} D \bar{\lambda}^{2} \tag{3.16}
\end{equation*}
$$

(note that the coefficient of $\bar{\lambda}$ equals 0 ).
If the vector $\binom{M}{\mathcal{O}}$ is linear independent of the basis (3.7), it can be shown that

$$
\begin{equation*}
M_{b}^{\prime} \cdot D \neq 0 . \tag{3.17}
\end{equation*}
$$

To prove this Kriens and van Lieshout study problem (1.7) with $\mathcal{A} X \leq B$. With obvious adaptations in the notation, the Kuhn-Tucker conditions of this problem are in our case

$$
\begin{align*}
-2 \mathcal{C} X-\mathcal{A}^{\prime} U+M \lambda+V & =\mathcal{O}  \tag{3.18}\\
&  \tag{3.19}\\
\mathcal{A} X & =B
\end{align*}
$$

$$
\begin{equation*}
M^{\prime} X-y_{m+1} \quad=\bar{\mu} \tag{3.20}
\end{equation*}
$$

$$
\begin{equation*}
X^{\prime} V=y_{m+1} \cdot \lambda=0, X \geq \mathcal{O}, V \geq \mathcal{O}, y_{m+1} \geq 0, \lambda \geq 0, U \text { free. } \tag{3.21}
\end{equation*}
$$

Because $\binom{M}{\mathcal{O}}$ is assumed to be linear independent of $\mathcal{B}_{v}$, the vector $Z_{b}$ (3.5) completed with $\lambda$, forms a basic solution of (3.18), . .,(3.21). Reordering in the same way as in (3.8) the matrix of basic vectors changes into

$$
\mathcal{B}_{v}^{*}=\left(\begin{array}{cc}
\mathcal{B}_{v} & K  \tag{3.22}\\
L^{\prime} & \mathcal{O}
\end{array}\right)
$$

with

$$
L^{\prime}=\left(\begin{array}{lll}
M_{b}^{\prime} & \mathcal{O}^{\prime} & \mathcal{O}^{\prime} \tag{3.23}
\end{array}\right)
$$

and

$$
\begin{equation*}
K^{\prime}=\left(M_{b}^{\prime} \mathcal{O}^{\prime} \quad M_{n b}^{\prime}\right) \tag{3.24}
\end{equation*}
$$

Using the existence of $\left(\mathcal{B}_{v}^{*}\right)^{-1}$, (3.17) can be proved.
Remark 3.1. The condition $\binom{M}{\mathcal{O}}$ linear independent of the basis (3.7) is incorrectly suppressed by J. Kriens and J.TH. van Lieshout (1988). J. Kriens (1989) provides a counter example.

## 4 Necessary and sufficient conditions for differentiability of the efficient frontier

Because of property $b$ in section 1 we can restrict the discussion to the points $\left(\mu\left(\bar{X}_{i}\right), \sigma^{2}\left(\bar{X}_{i}\right)\right)$, in the sequel to be denoted by $\left(\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}\right)$. Furthermore we only discuss nondegenerate models.

## Condition 1.

The efficient frontier (e.f., for short) is differentiable in the point ( $\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}$ ) iff one value $\bar{\lambda}$ corresponds to it.

## Proof.

Follows directly from (1.13) and (1.14).

## Condition 2.

The e.f. is differentiable in the point $\left(\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}\right)$ iff no corresponding $\bar{X}_{b}$-vector can be represented by (2.1) with $D=\mathcal{O}$.

## Proof.

Necessary: $D=\mathcal{O}$ implies the same vector $\bar{X}_{b}$ and thus the same point $\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}$ ) for more than one value of $\bar{\lambda}$.
Sufficient: $D \neq \mathcal{O} ; \bar{\lambda}_{1} \neq \bar{\lambda}_{2} \Rightarrow \bar{X}\left(\bar{\lambda}_{1}\right) \neq \bar{X}\left(\bar{\lambda}_{2}\right)$ and so different points $\left(\bar{\mu}, \bar{\sigma}^{2}\right)$, cf. lemma 2.1.

## Condition 3.

The e.f. is differentiable in the point $\left(\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}\right)$ iff no corresponding $\bar{X}_{b}$-vector can be represented by (2.1) with $M_{b}^{\prime} . D=0$.

## Proof.

Follows from $D \neq \mathcal{O} \underset{\rightarrow}{\leftrightarrows} M_{b}^{\prime} . D \neq 0$.
$\rightarrow$ if $\bar{\lambda}$ changes, $\bar{X}_{b}$ changes and $\mu\left(\bar{X}_{b}\right)$ must change (lemma 2.1), so $M_{b}^{\prime} . D \neq 0$ (cf.
$\leftarrow$ trivial.

## Condition 4.

The e.f. is differentiable in the point $\left(\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}\right)$ iff $\left(\mathcal{B}_{v}^{*}\right)^{-1}$ exists.

## Proof.

Follows from $\left(\mathcal{B}_{v}^{*}\right)^{-1}$ exists $\underset{\rightarrow}{\leftrightarrows} M_{b}^{\prime} . D \neq 0$.
$\rightarrow$ see J. Kriens and J.TH. van Lieshout (1988) p. 190-191.
$\leftarrow$ if $M_{b}^{\prime} . D \neq 0$, all elements of $\left(\mathcal{B}_{v}^{*}\right)^{-1}$ exist and $\mathcal{B}_{v}^{*} \cdot\left(\mathcal{B}_{v}^{*}\right)^{-1}=\mathcal{J}$.

## Condition 5.

The e.f. is differentiable in the point $\left(\bar{\mu}_{i}, \bar{\sigma}_{i}^{2}\right)$ iff $\binom{M}{\mathcal{O}}$ is linear independent of the vectors of $\mathcal{B}_{v}$.

Proof.
$\binom{M}{\mathcal{O}}$ linear independent of the vectors of $\mathcal{B}_{v} \stackrel{\leftarrow}{\rightarrow}$ inverse of $\mathcal{B}_{v}^{*}$ exists.

## 5 Relations with statements on differentiability in the literature

The theorem stated by J. Vörös (1987) and J. Kriens (1989) are easily checked through applying the conditions of section 4 . We combine these theorems in one new theorem. Define $\mu_{\text {min }}:=\min _{i} \mu_{i}, \mu_{\text {max }}:=\max _{i} \mu_{i}$,

$$
\begin{align*}
& \mathcal{M}=\left(m_{i j}\right):=\mathcal{C}_{b}^{-1}  \tag{5.1}\\
& f:=\sum_{i=1}^{k} \sum_{j=1}^{k} m_{i j}  \tag{5.2}\\
& d:=\sum_{i=1}^{k}\left(\sum_{j=1}^{k} m_{i j} \mu_{j}\right) . \tag{5.3}
\end{align*}
$$

## Theorem 5.1

If in the investment problem subject to (1.2) and (1.3), $\mathcal{C}$ positive definite, a corner portfolio with $\mu \in\left(\mu_{\min }, \mu_{\max }\right)$ has $k(\geq 1) x$-variables in the basis, then the set of efficient ( $\bar{\mu}, \bar{\sigma}^{2}$ ) points is nondifferentiable in the corresponding ( $\bar{\mu}, \bar{\sigma}^{2}$ ) point if and only if there exists a representation of $\bar{X}_{\dot{b}}=\left(\bar{x}, \ldots, \bar{x}_{k}\right)$ with $\forall_{1 \geq i . j \geq k} \mu_{i}=\mu_{j}$.

## Proof.

We distinguish between $k=1$ and $k>1$.

## Sufficient.

$k=1$. Suppose $\bar{x}_{i}>0$, then $\bar{x}_{i}=1, \mathcal{C}_{b_{1}}=\left(c_{i i}\right), \mathcal{A}_{b}=(1), M_{b}=\left(\mu_{i}\right)$. Substitution of these values into (3.14) leads to

$$
\begin{align*}
D & =\frac{1}{2}\left[\mathcal{C}_{b_{1}}^{-1}-\mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1} \mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1}\right] M_{b}  \tag{5.4}\\
& =\frac{1}{2} c_{i i}^{-1}\left[1-\left(c_{i i}^{-1}\right)^{-1} c_{i i}^{-1}\right] \mu_{i}=0 .
\end{align*}
$$

So $D=\mathcal{O}$ and from condition 2 nondifferentiability follows.
$k>1$. Let the representation with $k$ variables in the basis be

$$
\bar{X}_{b}=\left(\begin{array}{l}
\bar{x}_{1} \\
\vdots \\
\bar{x}_{k}
\end{array}\right), \mathcal{C}_{b_{1}}=\left(\begin{array}{ccc}
c_{11} & \ldots & c_{1 k} \\
\vdots & & \vdots \\
c_{k 1} & \ldots & c_{k k}
\end{array}\right), \mathcal{A}_{b}^{\prime}=\left(\begin{array}{c}
1 \\
\vdots \\
1
\end{array}\right), M_{b}=\left(\begin{array}{l}
\mu_{1} \\
\vdots \\
\mu_{k}
\end{array}\right)
$$

then

$$
\begin{equation*}
\left(\mathcal{A}_{b} \mathcal{C}_{b_{1}}^{-1} \mathcal{A}_{b}^{\prime}\right)^{-1}=\frac{1}{f} \tag{5.5}
\end{equation*}
$$

and $D$ can be rewritten as

$$
D=\frac{1}{2} \mathcal{M}\left[\mathcal{J}-\frac{1}{f}\left(\begin{array}{ccc}
\sum_{i} m_{i 1} & \ldots & \sum_{i} m_{i k}  \tag{5.6}\\
\vdots & & \vdots \\
\sum_{i} m_{i 1} & \ldots & \sum_{i} m_{i k}
\end{array}\right)\right]\left(\begin{array}{l}
\mu_{1} \\
\vdots \\
\mu_{k}
\end{array}\right)
$$

If $\mu_{1}=\ldots=\mu_{k}$, then $D=\mathcal{O}$ and condition 2 leads again to nondifferentiability.

## Necessary.

$k=1$. Trivial.
$k>1$. If there is nondifferentiability then there exists a representation with $D=\mathcal{O}$. For this vector (5.6) is equivalent to

$$
\left[\mathcal{J}-\frac{1}{f}\left(\begin{array}{ccc}
\sum_{i} m_{i 1} & \ldots & \sum_{i} m_{i k}  \tag{5.7}\\
\vdots & & \vdots \\
\sum_{i} m_{i 1} & \ldots & \sum_{i} m_{i k}
\end{array}\right)\right]\left(\begin{array}{l}
\mu_{1} \\
\vdots \\
\mu_{k}
\end{array}\right)=\left(\begin{array}{c}
0 \\
\vdots \\
0
\end{array}\right)
$$

or

$$
\left(\begin{array}{l}
\mu_{1} \\
\vdots \\
\mu_{k}
\end{array}\right)=\frac{1}{f}\left(\begin{array}{ccc}
\sum_{j} & \left(\sum_{i} m_{i j}\right) & \mu_{j} \\
\vdots & & \vdots \\
\sum_{j} & \left(\sum_{i} m_{i j}\right) & \mu_{j}
\end{array}\right)=\left(\begin{array}{c}
\frac{d}{f} \\
\vdots \\
\frac{d}{f}
\end{array}\right)
$$

so nondifferentiability implies $\mu_{1}=\ldots=\mu_{k}$.

## Remark 5.1

In the case of constraint (1.1) instead of (1.3), $X_{b}$ cannot contain only one $x$-variable if
(1.1) contains two or more independent constraints.

## Remark 5.2

Theorem 5.1 combines the theorems 5.1 and 5.2 in J. Kriens (1989) and generalizes the case $k>1$ to situations in which the basis contains $x$-variables with value 0 . The theorem also generalizes theorem 2 by J. Vörös (1987).

## Remark 5.3

The theorem can likewise be proved by directly applying condition 5 from section 4 .

## 6 The standard portfolio selection problem with one riskless asset.

The standard portfolio selection problem with conditions (1.2) and (1.3) can also be formulated as

$$
\begin{equation*}
\min _{X} \sigma^{2}(X)=X^{\prime} \mathcal{C} X \tag{6.1}
\end{equation*}
$$

subject to

$$
\begin{equation*}
X \geq \mathcal{O} \tag{6.2}
\end{equation*}
$$

$$
\begin{equation*}
\sum_{j=1}^{n} x_{j}=1 \tag{6.3}
\end{equation*}
$$

$$
\begin{equation*}
M^{\prime} X=\mu, \tag{6.4}
\end{equation*}
$$

using $\mu$ as a parameter; the optimal solution is denoted as $\bar{X}(\mu)$.
Now, consider the standard portfolio case with one riskless asset: minimize (6.1) subject to (6.2),

$$
\begin{equation*}
\sum_{j=1}^{n} x_{j}+y=1 \tag{6.5}
\end{equation*}
$$

$$
\begin{equation*}
M^{\prime} X+i y=\mu, \tag{6.6}
\end{equation*}
$$

where $y$ is the share of capital invested in the riskless asset and $i$ is the rate of interest; we allow $y$ to be positive, 0 or negative.
We can easily state that for $\mu=i$ the optimal solution runs $\bar{y}=1, \bar{X}(i)=\mathcal{O}$ with $\sigma^{2}(\underline{r}(\bar{X}(i)))=0$. Thus we can restrict to the case $\mu>i$; furthermore we assume $i<\max _{j}\left\{\mu_{j}\right\}$. Let again $X_{b}^{\prime}=\left(x_{1}, \ldots, x_{k}\right)$ represent the set of basic $x$-variables and
$X_{n b}^{\prime}=\left(x_{k+1}, \ldots, x_{n}\right)$ the set of non-basic $x$-variables. Denote the Lagrange multipliers of (6.5) and (6.6) by $u_{1}$ and $\lambda$ respectively, and let $I_{n}=\left(\begin{array}{c}1 \\ \vdots \\ 1\end{array}\right)$ with $n$ elements. The Kuhn-Tucker equations for the problem (6.1), (6.2), (6.5), (6.6) are:

$$
\begin{align*}
& 2 \mathcal{C}_{b_{1}} X_{b}+I_{k} \cdot u_{1}-M_{b} \lambda=\mathcal{O}  \tag{6.7}\\
& 2 \mathcal{C}_{b_{2}} X_{b}+I_{n-k} \cdot u_{1}-M_{n b} \lambda \geq \mathcal{O}  \tag{6.8}\\
& \quad-u_{1}+i \lambda=0  \tag{6.9}\\
& X \geq \mathcal{O}  \tag{6.2}\\
& I_{k}^{\prime} X_{b}+y=1  \tag{6.10}\\
& M_{b}^{\prime} X_{b}+i y=\mu \tag{6.11}
\end{align*}
$$

From (6.7) we have

$$
\begin{equation*}
X_{b}=-\frac{1}{2} u_{1} \mathcal{C}_{b_{1}}^{-1} I_{k}+\frac{1}{2} \lambda \mathcal{C}_{b_{1}}^{-1} M_{b} . \tag{6.12}
\end{equation*}
$$

With (5.2), (5.3) and

$$
\begin{equation*}
e:=\sum_{i=1}^{k} \sum_{j=1}^{k} m_{i j} \mu_{i} \mu_{j} \tag{6.13}
\end{equation*}
$$

we can derive from (6.10) and (6.11)

$$
\begin{equation*}
I_{k}^{\prime} X_{b}=-\frac{1}{2} f u_{1}+\frac{1}{2} d \lambda=1-y \tag{6.14}
\end{equation*}
$$

$$
\begin{equation*}
M_{b}^{\prime} X_{b}=-\frac{1}{2} d u_{1}+\frac{1}{2} e \lambda=\mu-i y . \tag{6.15}
\end{equation*}
$$

## Lemma 6.1

The expression $f i^{2}-2 d i+e$ is always positive, except in the case $\forall_{i \in\{1, \ldots, k\}} \mu_{i}=i$.

## Proof

$\left(M_{b}-i I_{k}\right)^{\prime} \mathcal{C}_{b_{1}}^{-1}\left(M_{b}-i I_{k}\right)=f i^{2}-2 d i+e=0$ iff $M_{b}=i I_{k} \leftrightarrows \forall_{i \in\{1, \ldots, k\}} \mu_{i}=i$ (cf. also J. Vörös (1987)).

As $\forall_{i \in\{1, \ldots, k\}} \mu_{i}=i$ implies $\mu=i$, the case we excluded, $f i^{2}-2 d i+e$ is always $>0$ in our model.

## Lemma 6.2

For a given set of basic $x$ - variables $X_{b}$ the problem (6.1), (6.2), (6.5), (6.6) has a unique solution.

## Proof

Eliminating $y$ from (6.14), (6.15) and using (6.9) we find

$$
\begin{equation*}
\lambda=\frac{2(\mu-i)}{f i^{2}-2 d i+e} \tag{6.16}
\end{equation*}
$$

and

$$
\begin{equation*}
u_{1}=\frac{2 i(\mu-i)}{f i^{2}-2 d i+e} . \tag{6.17}
\end{equation*}
$$

From these equations and (6.12) it follows that the solution is unique.

In the remainder of this section we exploit the well-known property that in the $(\mu, \sigma)$ plane the e.f. of the model $(6.1),(6.2),(6.5),(6.6)$ is a straight line through the point $\mu=i, \sigma=0$ which touches the e.f. of the risky assets of the model $(6.1), \ldots,(6.4)$ if this e.f. is differentiable (cf. e.g. Th.E. Copeland and J.F. Weston (1988) p. 179-180). This property implies that we can find the e.f. of the risky assets by using $i$ as a parameter: with every value of $i$ there corresponds one point of the e.f. of risky assets and so one
set of basic variables $X_{b}$. Formulae (6.12) and (6.8) provide a simple procedure for deriving the corner portfolios of the risky assets. Therefore we rewrite (6.12) and (6.8) by substituting (6.16) and (6.17) into

$$
\begin{align*}
& \mathcal{C}_{b_{1}}^{-1} M_{b}-i \mathcal{C}_{b_{1}}^{-1} I_{k} \geq \mathcal{O}  \tag{6.18}\\
& \mathcal{C}_{b_{2}} \mathcal{C}_{b_{1}}^{-1} M_{b}-M_{n b}+i\left(I_{n-k}-\mathcal{C}_{b_{2}} \mathcal{C}_{b_{1}}^{-1} I_{k}\right) \geq \mathcal{O} \tag{6.19}
\end{align*}
$$

The algorithm runs as follows.

Step 1: Determine $\max \left\{\mu_{j}\right\}$ and fill up the sets $X_{b}$ and $X_{n b}$.
Step 2: Find the smallest value of $i$ for which (6.18) and (6.19) hold.

Step 3: If $i=-\infty$ then stop. Otherwise remove the variable from $X_{b}$ into $X_{n b}$ if $X_{b} \geq \mathcal{O}$ gives the smallest $i$, or inversely. Repeat step 2.

If we apply this algorithm to the well-known Markowitz example

$$
M=\left(\begin{array}{l}
1 \\
3 \\
5
\end{array}\right) \quad \mathcal{C}=\left(\begin{array}{rrr}
3 & 3 & -1 \\
3 & 11 & 23 \\
-1 & 23 & 75
\end{array}\right)
$$

then $\mu_{3}=\max \left\{\mu_{j}\right\}, X_{b_{1}}=\left(x_{3}\right)$ and we find successively

$$
X_{b_{2}}=\binom{x_{2}}{x_{3}}, X_{b_{3}}=\left(x_{2}\right), x_{b_{4}}=\binom{x_{1}}{x_{2}}, X_{b_{5}}=\left(\begin{array}{c}
x_{1} \\
x_{2} \\
x_{3}
\end{array}\right), X_{b_{b}}=\binom{x_{1}}{x_{3}} .
$$

The expressions (6.7) (or (6.12)) and (6.8) are a special form of the equations (3.1). Raising the value $\bar{\lambda}$ of $\lambda$ from 0 to the largest relevant value in the standard algorithm is the same as lowering $i=\frac{u_{1}}{\lambda}$ from the largest relevant value to $-\infty$ in the algorithm just presented. Both algorithms produce exactly the same steps, albeit in a reverse order.

The model of this section gives us the opportunity to present another proof of theorem 5.1.

## Another proof of theorem 5.1.

If the e.f. of the risky assets is nondifferentiable in a point $P$ there are different interest rates $i$ from where we can draw subgradients to $P$. For this interval of values $i$ the return on the portfolio of risky assets is the same i.e. independent of $i$. Using (6.15) we get

$$
\begin{equation*}
M_{b}^{\prime} X_{b}=-\frac{1}{2} d u_{1}+\frac{1}{2} e \lambda=\mu \tag{6.20}
\end{equation*}
$$

with $\mu$ the expected return of the corresponding portfolio. We substitute (6.16) and (6.17) for $\lambda$ and $u_{1}$ to find

$$
\begin{equation*}
(d-f \mu) i-(e-d \mu)=0 . \tag{6.21}
\end{equation*}
$$

This can hold for the whole interval of $i$-values iff

$$
\begin{equation*}
d-f \mu=0 \text { and } e-d \mu=0 \tag{6.22}
\end{equation*}
$$

from which follows

$$
\begin{equation*}
f \mu^{2}-2 d \mu+e=0 \tag{6.23}
\end{equation*}
$$

and with lemma 6.1: $\mu_{i}=\mu_{j}=\mu$ for all $x_{i}, x_{j} \in X_{b}$.

Considering the case $\mu_{i}=\mu_{j}=\mu$ for all $x_{i}, x_{j} \in X_{b}$ we find with (5.2), (5.3), (6.13)

$$
\begin{equation*}
e=\mu^{2} f \text { and } d=\mu f \tag{6.24}
\end{equation*}
$$

and with (6.12)

$$
\begin{equation*}
X_{b}=-\frac{1}{2}\left(u_{1}-\mu \lambda\right) \mathcal{C}_{b_{1}}^{-1} I_{k} . \tag{6.25}
\end{equation*}
$$

Substitution of the equations (6.16) and (6.17) leads to

$$
\begin{equation*}
X_{b}=\frac{1}{f} \mathcal{C}_{b_{1}}^{-1} I_{k}, \tag{6.26}
\end{equation*}
$$

which is independent of $i$, so subgradients can be drawn to a whole interval of $i$-values.

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